A COURSE IN CELESTIAL MECHANICS
VOLUME 2
by

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This is the second volume of "a course in celestial mechanics" and must be regarded as an immediate continuation of the first volume. This volume is concerned with the general theory of perturbative motion, the methods of evaluation of the perturbations of planets and comets and principles on the theory of the motion of the Moon.

Tilis book may serve as a text-book for undergraduate and postgraduate students and being a sufficiently complete monograph on celestical rechanics, it may be of interest to all scientists working in this field.

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Translation Editor's Note: The reader is advised to consult the foreign
text for greater legibility of the graphic material.
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## PRERACR

Whereas the first volume of this course was entirely devoted to the two-body preblem and to the methods for determining the orbitals, this volume is mainly given to the perturbation theory and its application in celestial mecharics.

In this book, it has been my aim to give beginners an orientation in modern celestial mechanics and to introduce them to the periodic literature. Secondly, I hope that this book will be accepted as a comprehensive and practical treatise on the solution of some of those problens which usually meet astronomers.

The material of this book is mainly based on the general and spectal courses I gave in the University of Leningrad during the past six years. The book is divided into four sections. The first section is devoted to the study of the general properties of motion of materfal points mutually interacting according to Newton's law of gavitation and, in particular, to the most important properties of perturbative motion. This section may be considered as an introduction to the ensuing sections, in which the acrual determination of the perturbed coordinates is discussed.

The second section is given to the determination of the perturbed coordinates using nethods based on the numerical intergration of differential quations. Here, as well as in volume $I$, I alm to give an exhuastive manual for carrying out these operations. In volume I, I did not deal with a detailed account of the theory of numerical integration of differential equations, I gave but a brief summary of this important subject in an appendix to volume $I$ in order to fill as soon as possible one of the most important gaps in our literature. In volume II, I developed
the subject to the necessary degree, and gave only pertinent examples. I preferred not to bother the reader by referring continuously to the above mentioned appendix and $I$ found it better to expose all the theory of integration of the equations even at the cost of repeating $n$ nunber of pages. This allowed me to exclucle the above mentioned appendix from the new edition of volume $I$, which is being prepared at the present time.

After explaining in detail the methods of the numerical integration of the equations and illustrating this by means of several examples on how to use these methods, $I$ considered in detail the application of the numerical integration methods to the study of the unperturbed motion. I then applied these methods to the evaluation of perturbations, and here $I$ could not forego the illustration of examples, but the reader who stucied in detail the preceeding two chapters will not find a need for these examples.

The third and fourth sections of the book are concerred with the analytic methods of evaiuating perturbations. In these sections, I do not attempt to give a comprehensive account of the methods to be applied, as they arc already found in special monographs which $I$ am not trying to replace and thus restricted myself to the complete presentation of the main points of each of these important methods. Considerable space is devoted to the study of the motion of the Moon. This is not only one of the most well-developed areas of elestial mechanics, but is also considered not less important than the motion theory of planets. I do not need to mention that the theory of the motion of the Moon is used in star astronomy in the study of multiple stars, We do not forget that the work by Hill. is considered to be one of the most important reference sources in celestial mechanics during the past decade.

Volume II is hence mainly devoted to the study of the methods of celestial mechanics. More details on the results, namely the comparison between theory and observation, will be given in volume III. There, ore special methods (periodic orbitals, : :thods of Guiden and krendel) and also the theory of the motion of stars will be given.

I tried to make the standard of presentation suitable for the selfstudy of the subject and I avoided complicated mathematical methods as much as possible. I hope that lecturers will be able to choose from this book various topics according to the standard and the degree of completeness they reaisire.

I have carefully chosen notations for quantities, for which no standard notations exist and $I$ would like to point out that the introduction of a complete set of noterions is not only difficult, but also not alwavs useful. For example, I used different notations in the motion theory of planets and in the theory of motion of the Moon. This should not cause the reader any trouble and should help him in reading the special ifterature.

I do not claim a complete and systematic bibliography and more details on the available bihliography could be found by investigating the well known literature.

Finally, I note that the chapters devoted to the determination of orbitals obtained from the various observations, which $I$ had planned to include in volume $1 I$, were included in the new edition of volume $I$, thus allowing the contents of the present volume to be more homogeneous.

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## PART ONE

PRINCTPLES OF THE THEORY OF PERTURBATTVE MOTION
CHAPTER I
THE n-BODY PROBLEM

## 1. The integrals of the n-body problem

Let $11 s$ consider $n$ material points of masses $m_{0}, m_{1}, \ldots m_{n-1}$. We denote by $\xi_{i}, \boldsymbol{\gamma}_{i}$ and $\boldsymbol{\zeta}_{i}$ the coordinates of point $m_{i}$ relative to an arbitrary system of axes. Let $\boldsymbol{\Delta}_{i j}$ be the distance between points $m_{i}$ and $m_{j}$, so that

$$
\Delta_{11}^{2}=\left(\xi_{1}-\xi_{1}\right)^{n}-p\left(\eta_{1}-\eta_{1}\right)^{2}+\left(\eta_{1}-\vdots\right) .
$$

The force of eravitation by which point $m_{j}$ acts on puint $m_{i}$ is equal to $k^{2} m_{i} m_{j} \Delta_{i j}^{-2}$, where $k^{2}$ is the coefficient of gravity (vol. I, SS 3 and/6). The projections of this force on the coordinate axes are

The equations of motion of point $m_{i}$ are then given by


$m_{1} \frac{d \cdot r_{2}}{d!}-k=\sum_{1} m_{1} m_{1} \frac{j_{i} j_{i}}{J_{i,}^{i}}$.
where terms with $i=j$ should be dropped in the summation. Fquation (1), with $1=0,1, \ldots, n-1$ form a system of the 6 n-order, the integration of which gives a complete solution of the $n$-body problem. These equations may also be written as
once the force function
is introduced.
Since we are dealing with a system on which no external force acts, the general equations of mechanics allow us to derive for the system 6 integrals of motion of the centre of mass, and 3 integrals of area. These integrals can be easily obtained from equation (1). Substituting $i$ in each of equations (1) where $i=0$ to $i=n-1$ to obtain
where the dots denote differentiations with respect to time.
Then
where $\alpha_{1}, \beta_{1}, \gamma_{1}, \alpha_{2}, \beta_{2}$ and $\gamma_{2}$ are arbitrary constants. Equations (4) and (j) show that the centre of mass of the system moves with a linear and uniform velocity, These relations are called the integrals of motion of the centre of mass.

In order to derive the integrals of area, let us consider the following relations, which can be easily obtained $f_{1}$ om equations (1):

$$
\begin{aligned}
& \ddot{1} m_{1}\left(\xi_{i} \ddot{r}_{1}-r_{1,} \bar{i}_{1}\right)=0 \text {. }
\end{aligned}
$$

Integrating these equations and denoting by $C_{1}, C_{2}$ and $C_{3}$ the new arbitrary constants, we obtain the integrals of area

$$
\begin{align*}
& \sum_{1} m_{1}\left(r_{1} \dot{m}_{1} \quad \ddot{p}_{1} \dot{\eta}_{1}\right) \cdots C_{1} \\
& \because m_{1}\left(\sigma_{i}\right)-C_{0}  \tag{6}\\
& \sum m_{1}\left(\bar{c}_{i} \dot{I}_{2}-\eta_{i, i}\right)=C_{2} .
\end{align*}
$$

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OF POOR OF POOR QUALITY
The left-hand side of equations (6) are the projections on the coordinate axes of the angular momentum of the system, which equals the sum of angular momenta of all the points. In this manner, equations (6) show that the magnitude and direction of the angular momentum of the system are conseried. The plane perpendicular to the angular momentum of the system is given by the following equation
and called Laplace's invariable plane.
Each of the terms of the sums in the left hand side of equation (6) can be interpreted $a^{-}$the projection of the areal velocity of a given point $m$ on the respective coordinate $p l a n e$, multiplied by the mass of the particle. In other words, the left hand side of each of these equations is the sum of the mass projections of areal velocities of all particles of the system on one of the coordinate planes. If the sum of mass areal velocities is projected on an arbitrarily chosen plane, the projection will evidently be equal to

$$
\ddot{V}^{\prime} C_{1}^{2}+C_{2}^{2}+r_{3}^{2} \cos B .
$$

where $\beta$ is the angle between the normal to this plane and the angular monentum of the system. From here, it follows that the invariable plane may be defined as the plane for which the sum of mass nrojections of the areal velocities of all the points of the system is maximum.

The remaining integral of motion owes its existence to the situation that the force function, given by equation (3), does not depend explicity
 and $\zeta_{i}^{\prime}$ respectivel $y$, and by adding the resulting equations, we obtain

The integration of this equation directly leads to the following integral of kinetic energy

$$
\begin{equation*}
\because \searrow m_{1}\left(i_{i}+\dot{\eta}_{i}^{\prime}+\dot{y}_{1}\right)=u+h . \tag{8}
\end{equation*}
$$

where $h$ is a new constant. The quantity

$$
r=\begin{aligned}
& 1 \\
& 2
\end{aligned} m_{i}\left(i_{i}, i_{1}^{\prime}+F_{i}\right)
$$

is act ually the kinetic energy of the system. The potential energy is equal to - U. Then, the total energy is

$$
M=T-U
$$

and the integral (8), assuming $H=h$, expresses the law of conservation of energy for the system under consideration.

We conclude that the general theorems of mechanics lead for the n-body problem to ten integrals of motion. Various attempts to find other integrals have not been successful. Tn the year 1887, Bruns has shown that any first integral, algebraic with respect to the coordinates $\xi_{i}, \eta_{i}$ and $\zeta_{i}$ and their derivatives $\dot{\xi}_{i}, \bar{\eta}_{i}$ and $\dot{\zeta}_{i}$ should be a consequent of the tan integrals obtained above even for three body problem. In the year 1889, Pofncare also found that for three body problems there exists no other first integrals that could be definitely expressed by single valued analytic functions. We shall not stop on the proof of these theorems. Knowing that they exist, we shall
not search for other iniegrals of motion, other than the above mentioned ten. If other integrals of motion were obtainable, they would be too complicated to be of practical use.
2. Reduction of the n-body problem to the integration of a system of order $6 \mathrm{n}-12$ and two quadratures

The 6 integrals of motion of the centre of mass, the 3 integrals of area and the integral of kinetic energy found in the previous section reduce the order ofsystem (1) by ten units. We can reduce the order of the system oy one more unit when we make use of the time independence of the gravitational forces. For this purpose, we have to exclude from the equations of motion the increment dt. The system of equations obtained in this manner is then integrated, while time is determined by evaluation of a quadratare. The problem is then reduced to the solution of a system of the order $6 n-11$ and one quadrature.

We can again reduce the crder of the system ty one more unit by making use of the property of the forces that they depend only on the interpoint distances. To do this, we introduce the generalized cuordinates in the following way. We draw a straight line passing through one of the points of the system and fix its direction. We then draw a plane through this line passing by another point in the system. We denote by $\varphi$ the azimuthal angle of this plane with respect to an arbitrary direction and let this angle be one of the generalifed coordinates. The motion of the whole system will be defined by the coordinate $\varphi$ and the coordinates which define the position of the system relative to the plane. We then prove that the coordinate is cyclic, i.e. that the Lagrange equations of the system finvolve only the tine derivative $\varphi$ but not the coordinate $\varphi$ itself. We choose the $z$-axis in the direction of the fixed straightline and let the $x$-axis be in the rotating plane. We denote by $x_{k}, y_{k}$.and $z_{k}$ the coordinatee of point
$m_{k}$ in the rotating coordinate system. The components of velocity of point $m_{k}$ will then be:

$$
\dot{x}_{k}-y_{k} \dot{\dot{p}}, \quad \dot{y}_{k}+1-x_{k} \dot{f}, \quad \dot{b}_{k}
$$

The kinetic energy of the system will be given by

$$
r=\frac{1}{2} \sum_{k}^{1} m_{k}\left|\left(x_{k} \quad v_{k} \dot{y}\right)^{2}+\left(y_{k}+x_{k}\right)^{2} \cdot+z_{A}\right| \cdot
$$

and will thus depend on $\varphi$ but not on $\varphi$. Similarly, the Lagrangian of the system

wi. 11 yield
d. comat.

This equation allows us to exclude $\varphi$ from the remaining equations of motion. Integrating the system of equations obtained in this manner, rk will be able to know the motion of the material points relative to the rotating plane. From the last equation, we obtain an expression for $\dot{\varphi}$ as a function of the other coordin: es of the systen. Solving one quadrature, we obtain the value of th angle $\varphi$ which defines the position of the rotating pla $e$.

Th.e ahove procedure for the exclusion of the angle $\varphi$ from the equations of motion has been named "method of elimination of nodes" by Jacobi in his study of the three-body problem.

The integration of the system (1) is finally reduced to the integration of a system of $6 n-12$ equations and two quadratures. We re not going to do this reduction here since it is only necessary for
the complete solution of the n-body problem ${ }^{1}$. We shall only consider the reduction of the order of the system by ( units, using integrals (4) and (5), which is of practical importance.

## 3. Equations of relative motion

Let us return back to equations (1) wifich describe the motion of $n$ bodies relative to an arbitrary system of fixed axes. Let us make use of the integrals of motion, given by equations (4) and (5) to exclude from equations (1) three arbitrary coordinates, say $\boldsymbol{\xi}_{0}, \mathcal{Y}_{0}$ and $\boldsymbol{\xi}_{0}$ as well as their derivatives with respect to time. For this purpuse, we introduce the new coordinates $x_{i}, y_{i}$ and $z_{i}$ so that

These are the coordinates of point $m_{i}$ relative to three axes passing by point $m_{0}$ parallel to thefixed axes. Noting that $x_{0}=y_{0}=z_{0}=0$, we rewrite equations (1) as follows

Details on thedifferent methods of reduction of the order of the system of differential equations mentioned above can be found in the article: E.T. Whittakex, Prinziplen der Storungstheorie und allgemeine Theorie der Bahnkurven in dynamischen Problem, Enzyklopadie der mathem. Wiesenschaften, $\mathrm{Bd} . \mathrm{VI}_{2}, 512-556$, and al.so in the book "Analytic dynamics" by the same author where a whole chapter is devoted to the above problem. The reduction of the n-body problem to a system of $16 \mathrm{n}-12$ equations has been done by T.L. Bennett (Messenger of Math. (2), 644, 1901).

Where $\sum^{\prime}$ denotes a summation over $j$ in which terms with $j=0$ and $j=1$ are dropped. Introducing the notation $\Delta_{0 j}=r_{j}$, we obtain the following equations for the motion of point $m_{i}$ relative to point $m_{0}$
where

Introducing the notation

$$
\left.k_{1} n \sum_{1} n_{1} \left\lvert\, \begin{array}{ccc}
1 & 1_{1} x, & y_{1}, y!z_{2} z_{1}  \tag{10}\\
1,
\end{array}\right.\right)
$$

we rewrite equations (9) as follows

$$
\begin{align*}
& \begin{array}{ll}
d x_{1} \\
d i^{2} & R_{2}\left(m_{1} \mid m_{1}\right) \\
r_{i}^{3} & d x_{1} \\
r_{1}
\end{array} \\
& d y_{1}+k^{2}\left(m_{1}+m_{i}\right)_{i}^{y_{i}^{\prime}}=\begin{array}{l}
d k_{1} \\
d y_{i}
\end{array} \tag{11}
\end{align*}
$$

Let us assume that the masses of all points except points $m_{0}$ and $m_{i}$ equal zero. In thiscase $R_{f}=0$ and equations (11) turn out to be the well known Kepler problem of two interacting bodies (vol. I, Ch. II). In Astronomy, the Kepler problem is usually referred to as the nonperturbative motion, and all deviatious from titare called perturbative. For this reason, the functions $R_{i}$ will be referred to as perturbation
functions. The derivatives of these functions with respect to the coordinates of $m_{i}$ are equal to the components of the relative acceleration aquired by the body $m_{i}$ from its interaction with the rest $f^{f}$ the particles of the relative acceleration acquired by the body $m_{f}$ from its interaction with the rest of the particles of the system, except the central body $m_{0}$.

When we solve the ( $6 n-6$ )-order system, given by equations (11), we have a complete information about the motion of all points of the system relative to point $m_{0}$. After that, it is easy to obtain the absolute motion of all bodies of the system. Actually, equations (5) lead to

$$
\begin{align*}
& 7: / F_{u}+\underline{y}_{1}^{n} m_{1}^{\prime} x_{1}=-a_{1} \ell+a \\
& M r_{i 0}+\stackrel{n-1}{\frac{1}{1}} m_{1} y_{1}=\beta_{1} t+\beta_{2}  \tag{12}\\
& M K_{0}+-_{1}^{\sum_{1}} n_{1} z_{1}=\gamma_{1} t+\gamma_{2},
\end{align*}
$$


where

$$
M=m_{0}+\sum_{1}^{n} m_{i}^{1}
$$

From these equations, we can find the values of $\xi_{0}, \mathcal{F}_{0}$ and $\xi_{0}$.
We can find four integrals of motion for the system (11) which corespond to the integrals (6) and (8) of the absolute motion. The first of equations (6) gives
or, in other words

We determine the values of $\xi_{0}, \eta_{0}$ and $\zeta_{0}$ from equations (12). Substituting here these values, we ohtain the following integrals of equation (11)
where

$$
\begin{aligned}
& C_{1}^{\prime} \quad M C_{1}+\beta_{1} B_{2} \quad \beta_{2} r_{2} \\
& C_{2}^{\prime}-M C_{2}+r_{1} a_{2}-r_{2} a_{1} \\
& C_{3}^{\prime}-M C_{3} 1-a_{1} \beta_{2}-a_{2} \beta_{1} .
\end{aligned}
$$

carrying out similar transformations on the integral of kinetic energy given by equation (8), we obtain another integral for the equations of relative motion.

Equations (11) are widely used in celestral mechanics, espectally in the study of the motion of planets and comets. In these cases, the central body $m_{0}$ is usuaily taken to be the sun. With this choice, each term of the perturbation function (10) is proportional to the mass $m_{j}$ of one of the planets. Therefore, the right-hand sides of equations (11) are so small that their influence can be treated as a perturbation. 4. A second form frr the equations of relative motion.

The form of equations of relative motion obtained in the previous section is not always convenient since these equations invilive a particular perturbation function $R_{i}$ for each body. Sometimes a different form of equations of relative motion is used, which is based on the following choice of relative coordinates:
(1) Lraw three coordinate axes through the first point. $m_{0}$, parallel to the fixed axes and define the position of point $m_{1}$ in this
system by the coordinates $\mathrm{x}_{1}, \mathrm{y}_{1}$ and $\mathrm{z}_{1}$.
(2) Draw three other axes through the centre of gravity $G_{1}$ of points $m_{0}$ and $m_{1}$ parallel to the previous axes and define the position of point $m_{2}$ relative to these axes by the coordinates $x_{2}, y_{2}$ and $z_{2}$.
(3) Define the position of point $m_{3}$ by the coordinates $x_{3}, y_{3}$ and $z_{3}$ relative to a system of axes parallel to the previous ones and having their origin at the centre of gravity of points $m_{0},{ }^{10} 1$ and $\mathrm{m}_{2}$; and so or.
In this way, every subsequent point $m_{i+1}$ is related to the centre of gravity $G_{i}$ of all previous points $m_{0}, m_{1}, \ldots m_{i}$.

Let $X_{i}, Y_{i}$ and $Z_{i}$ be the coordinates of point $G_{i}$, so that

$$
M_{1} x_{1}=m_{11} \xi_{11}+m_{1} \dot{亏}_{1}+\cdots \cdot f-m_{1}^{\xi},
$$

where

$$
m_{i}=m_{1}: \cdot m_{1} \mid \ldots \ldots+m_{1}
$$

By definition

$$
\lambda_{i}=6_{i}-\lambda_{i-1}
$$

so that

$$
\begin{align*}
& \text { in, } \left.x_{1}=H_{1},-1 m_{1} n_{0}, m_{1}, 1 \quad \mid m_{1} \vdots, 1\right) \tag{1.1}
\end{align*}
$$

In order to express the old coordinates, $\sum, \eta$ and $\mathcal{\eta}$, in terms of the new ones, $x, y$ and $z$, we note that

$$
M_{1} X_{1} \cdot H_{1}, \lambda_{1}==M_{i}
$$

or

$$
\begin{equation*}
\left.M_{1} X_{1}-H_{1} X_{1}, \quad\left(M_{1}-H_{1}\right)\right)_{1} \tag{15}
\end{equation*}
$$

Hence,

$$
\left(M_{1}-M_{1},\right)_{i} \cdot M_{1}\left(E_{1,1}-x_{1,1}\right) \cdot M_{i}, I_{i}-x_{1},
$$

so that

$$
\xi_{1+1}-\xi_{i}=x_{i+1}-M_{i, 1} M_{1}^{-i} x_{1} .
$$

Adding these equations term by term in successive order of index, we cbtain

$$
\bar{c}_{1}-\xi_{1}=x_{1}, \quad m_{1} x_{1}+\begin{gather*}
m_{1} 1_{1} x_{1-1} \\
\xi_{1+1}-\xi_{1}-x_{1+1}  \tag{16}\\
M_{1}
\end{gather*} M_{1}, m_{1} x_{1}
$$

We now write the differential equations of motion in terms of the new variables. Differentiating equation (14) twice, and using equations (2), we obtain

On the other hand, the following relations follov from equations (14)

so that

Finally, we obtain the following differential equations for the relative motion of the system
where

$$
H_{1}=\begin{gathered}
m_{1} M_{1} \quad 1 \\
A_{1}
\end{gathered}
$$

Equations (16) allow us to express the force function $U$ in terms of the new variables. Equations (17) are successfully applied in the study of the motion of satellites and systems of multiple stars. In order to study the motion of the Moon, for example, it is convenient to choose point $m_{0}$ to be the Earth, point $m_{1}$ the Moon and point $m_{2}$ the Sun. Denoting by $x_{1}, y_{1}$ and $z_{1}$ the coordinates of the Moon relative to the Earth, and by $x_{2}, y_{2}$ and $z_{2}$ the coordinates of the Sun relative to the centre of mass of the Moon and Earth, and putting

$$
a_{1}=\frac{m_{1} m_{0}}{m_{0}+m_{1}} \quad . \quad a_{4} \cdot \frac{m_{3}\left(m_{0}+m_{1}\right)}{m_{0}-1-m_{1}+m_{3}}
$$

we obtain the following equations of motion
where

$$
11 \cdots h_{1}=\left(\begin{array}{c:ccc}
m_{1} m_{2} & m_{0} m_{1} & m_{0} m_{2} \\
\nu_{1:} & 1_{1} & & \nu_{112}
\end{array}\right)
$$

It has been already pointed out (vol. I, $f(7)$ that the motlun of the sun relative to the centre of gravity of the gisten earth-moon is approximately elliptical. Therefore, the coordinates $x_{2}, y_{2}$ and $z_{2}$
sam approximately be found by solvins 3 two-body problem. This approximation essentially simplifies the solusin. f the system [18].

We finaliy show how to express the inetic energy in terms of the new coordinates. We aubstitute $x_{1}+\because$ for $\mathcal{F}_{i}$ in equation (15) and .btain

$$
\left.\left(M_{1}-H_{1}\right):=\vdots \quad-X_{1}, 1\right)
$$

Squaring bohis isies of this i. in and of equation (15) yields

$$
\begin{aligned}
& \left(1 \overline{1}_{1}-1 f_{1}\right): M_{1}\left(X_{1} \cdots X_{1}\right)^{2} \\
& \left(M_{1}-M_{1}\right)^{2}: \quad\left(M_{1} X_{1}-M_{1.1} X_{1}\right)^{2}:
\end{aligned}
$$

Eliminating here the product $X_{1} X_{i-1}$, we get

Suming this equation from $1=1$ to $1=n-1$, we obtain

$$
\sum_{1}^{1} m_{i} \vdots_{i}{\underset{1}{1} m_{1} H_{1} y_{1}, H_{n}, x_{n}, \cdots, H_{n}, i_{n}, ~}_{n_{1}}
$$

or, since $M_{0}=m_{0}$ and $X_{0}=\xi_{0}$,

$$
\because_{"}^{\prime} m_{1} \because_{1}-\because_{1}^{\prime} \because_{1} x^{2}+-1 t_{1}, x_{n}^{2} .
$$

Adding this equation term by term to the corresponding equations for the other coordinates, we find

This relation has been obtained owing to rhe linear relations between the coordinates $\mathcal{E}, \eta$ and $\zeta$ and the coordinates $x, y$ and $z$. Hence, $a$ similar relation holds for the derivatives of chese coordinates, such that

We note that equations (17) have the same form as equations (2). Therefore, we may obtain from equations (17) the integrais of ; real and kinetic energy by replacingithe masses $m_{i}$ by $\mu<i$ in the corresponding integrals obtained for the absolute motion.

The most general linear transformations of coordinates, which preserve the form of the integrals of motion in the three-body problem are given by Hopfner ${ }^{\text {(1) }}$.

## 5. Jacobi's formula

ORIGINAL PAG.
OF POOR QUA:
The Eorce function for the n-budy problem, given by equation (3), is a homogeneous function of ccordinates. Jacobi made use of this property to show wat the kinetic energy fintegral nay se obtained fin a very simple form.

Since $U$ is a homogeneous fanction of $\xi_{i i} \gamma_{i}$ and $\mathcal{Z}_{i}$ of the $(-1)^{\text {th }}$ order, then

Therefore, multiplying both sides of equations (2)respectively by
$\varepsilon_{i}, \mathcal{F}_{i}$ and $\tau_{i}$ and addirg, ve fiind

Adding this equation to the kinetic energy integral (8) ytelds

$$
{ }^{n} \because_{1}^{\prime} m_{1}\left(i_{1} \vdots+r_{1} r_{1}+\xi_{1} r_{1}+\xi_{i}+1-1, r_{1}\right) \cdots U+2 h,
$$

or
(1)

Hopfner, Uber eine Verallgemeinerung der relativer kinonischen Koordinaten von Jac'ol, Astr. Nachr., 195, 1913, 257-262.
or, finally

$$
\begin{equation*}
d=\sum_{1}^{n} m\left(\because_{i}^{2}-i \quad r_{1}^{2} \quad \because\right) \quad \because 11: 1 h \tag{119}
\end{equation*}
$$

The sum involved here

$$
J-\sum_{1}^{n} m_{i}\left(\because-i-r_{i}+\because\right)
$$

is the polar moment of inertia of our system. It is well known that $J$ can be expressed in terms of the squares of the interpoint distances as well as the quantif $y$

$$
J_{1} \cdot: \because(X-;: \because:
$$

where $M$ is the sum of a 11 masses $t_{i}$, and $X, Y$ and $Z$ are the coordinates of the centre of mass of the system. Making use of the following identity
where each combination of the symbols $i$ and $j$ in the right-hand side appears only once. Adding this identity to two similar identities for the variables $\eta$ and $\tau$, and noting that

$$
\begin{equation*}
M X=\searrow m_{i} \dot{i}_{i}, \quad M Y=\Sigma m_{i}, \quad M \%=\Sigma m_{i} V_{1} \tag{20}
\end{equation*}
$$

we obtain

$$
A H J-A M J_{0}==\sum_{1.1} m_{i} m_{1} S_{i}
$$

Hence, using the integrals (5), sharactarizing the motion of the centre of gravity of the system, we obtain

$$
M J=\left(x_{1} t+x_{2}\right)^{2}+\left(\beta_{1} i \cdot 1 \hat{r}_{0},+\left(\gamma_{1} f-1 i_{i 2}\right)^{2} \mid \sum_{1,1} m_{i} m_{1} د_{1,}^{3} .\right.
$$

Substituting this expression for $J$ in equation (19) and denoting by $h^{\prime \prime}$ the new constant, we obtain

$$
\begin{equation*}
d: R=2 U \cdot \mid \cdot 4 i^{\prime} \tag{2}
\end{equation*}
$$

where:

$$
R=-1 \not \sum_{i, 1}^{1} m_{4} m_{1} د_{11}^{2} .
$$



This formula is known as Jacoti's formula. A particular case of this formula was obtained by Lagrange (1772) for the three-body problem. However, the general case was obtained by Jacobi in 1842.

The following results of equation (21) was obtained by Jacobi and was probably the first application of the qualitative methods in celestial mechanics. Let us irtegrate equation (21) from 0 to $t$ and obtain

$$
d R \quad R_{0}^{\prime}+\left(\because a+4 n^{\prime}\right) r
$$

where $R_{0}^{\prime}$ stands for $\frac{d R}{d t}$ at $t=0$, and $\alpha$ for the lower limit of the function $[i$. Evidently $\alpha$ can be set equal zero. Another integration in the same limits yields.

This inequality shows that the motion of the system is stable only if $h^{\prime}<0$. Actualiy, if $h^{\prime} \geqslant 0$, then $\alpha+2 h^{\prime}>0$ ind the right-hand side of this inequality indefinitely increases as $t \rightarrow \infty$. In this case, at least one of the mutual distances $\Delta_{i j}$ should tend to infinity.

In the two-body case, equation (21) becomes
where $r$ is the distance between the two bodies and a is the semimajor axis of the relative orbital. If $a<0$, the relative motion proceeds, through a hyperbola so that $r \rightarrow \infty$ when $t \rightarrow \infty$.

## 6. Laplace's invariable plane

It was shown in $S 1$ that when $n$ material points are entirely under the action of their mutual gravitation, there exists a plane which conserves its direction in space. The plane is determined by equation (7), where the coefficients $C_{1}, C_{2}$ and $C_{3}$ are given by equations (6). However, equations (6) are not of practical use since they require the knowledge of the absolute motinn of all points of the system. Also, the integrais of area of the relative notica, given in the form of equations (13) are not useful since they involve the quantities: $\alpha_{1}, \alpha_{2}, \beta_{1}, \ldots \ldots$ which characterize the absolute motion of the centre of mass.

We shall see now that the direction of the invariable plane may be found in terms of onily the relative coordinates and velocities of the points of the systum. The reason for this is very simple. Equations (6) or (13) determine the values of the quantities $C_{1}, C_{2}$ and $C_{3}$. Hnwever, it is sufficient to the ratios $C_{1}: C_{2}: C_{3}$ in order to define the position of the invariable plane.

Let us introduce a new coordinate system having its origin at the centre of mass of points $m_{0}, m_{1}, \ldots m_{n-1}$ and the directions of its axis in the space fixed. The coordinates of point $m_{i}$ in this system are denoted by $x_{i}, y_{i}$ and $z_{j}$, and the absoluce coordinates of the centre of mass by $X, Y$ and $Z .$. Using equations (5) añ (17), we obtain

$$
M X^{-} a_{1} t+\alpha_{2}, \quad M Y-\beta_{1} t+\beta_{2}, \quad M Z=\gamma_{1} t+\gamma_{2}
$$

so that

$$
\ddot{X}=\ddot{Y}=\ddot{Z}=0
$$

We now express the old coordinates in equations (1) in terms of the new coordinates using the following relations

$$
\varepsilon_{i}=x+x_{i}, \quad y_{i}=y+y_{i}, \quad z_{i}-\dot{z}: \cdot z_{i}
$$

We then obtain the following equations for the motion of point $m_{i}$ relative to the centre of mass of the system

Since these equations have the same form as equations (1), we can immediately write for them the integrals of area

$$
\begin{aligned}
& \Xi m_{1}\left(y_{i} \dot{z}_{i}-x_{i} y_{i}\right)-\widetilde{C}_{i} \\
& \Xi m_{i}\left(x_{i} \dot{x}_{i}-x_{1} z_{i}\right)=C_{3}^{\prime} \\
& \Xi m_{i}\left(x_{i} \dot{y}_{1}-y_{i} \dot{x}_{1}\right)=C_{i}^{\prime \prime}
\end{aligned}
$$

The constants $C_{1}^{\prime \prime}, C_{2}^{\prime \prime}$ and $C_{3}^{\prime \prime}$ fix the position of the Laflace's plane passing through the arbitrary point $X^{0}, y^{\circ}$ and $z^{0}$ such that its equation is

$$
C_{2}^{\prime \prime \prime}\left(x-x^{\prime \prime}\right)+C_{2}^{\prime}\left(y-y^{0}\right)-1 C_{3}^{\prime \prime}\left(z-z^{\prime \prime}\right)=0
$$

Thus, to determine the position of Laplace's plane, it is s"fficient to know the coordinates $x_{i}, y_{i}$ and $z_{i}$ and the components of velocity $x_{i}, y_{i}$ and $z_{i}$ for all points of the system at any $t i m e$.

In order to study the motion of the bodies in the solar system, it is more natural to choose the Laplace plane as a basis rather than use the ecliptics of a given epoch $(1750.0,1850.0$, or 1900.0$)$. However, the application of the Laplace plane is met with certain difficulties. The
position of this plane depends on th masses of the planets which are only approximately known. Consequently, it is only possible to approximately know the position of the Laplace plane. When new and more exact detercitnation of the masses of planets are made one must cansequen : $\perp$ y change the basic plane. Another difficulty comes from the fact that the sun and planets are not material points. The angular momentum, defined by the quantities $C_{1}^{\prime \prime}, C_{2}^{\prime \prime}$ and $C_{3}^{\prime \prime}$ may be changed by the values of angular momenta acquired by individual bodi:'s of the system, for example during tidal processes.

Relative to the ecliptic and equinox 1850.0 the position of the invariable plane is given by the elements

As expected, this plane slightly differs from the plane of the orbital of Jupiter and is situated between this planc and the orbital plane of Saturn.

## CHAPTER II

## THF EQUATIONS OF MOTION EXPRESSED IN POLAR COORDTNATES

## \#. The equations of motion expressed in cylindrical coordinates

It is well known that rectangular coordinates are used side by side with polar coordinates for the determination of the positions of material points. In the following, we express the equations of motion in different polar coordinates. We start by the simplest case of a stationary spherical polar coordinate system.

Let us consider the motion of a material point $P$, which we shall call-planet, relative to other point $S$, which we shall call-sun. We choose the origin of the coordinate system at point $S$, and call the plane $x y$ the ecliptic. We denote by $r$ and $\hat{\mathcal{O}}$ respectively, the radius vector of point $P$ and its projection in the plane $x y$, and by $v$ the longitude of this plane as measured from the $x$-axis. In this case

$$
\begin{aligned}
& r: \quad \therefore-\therefore
\end{aligned}
$$

In order to express the equations of motion of point $P$ in terms of the polar coordinates $\varphi, v$ anc $e$, it is best to start with the Lagrange equation

Let us put $q_{1}=\rho, q_{2}=v$ and $q_{3}=s$, and note that the kinetic energy is given in terins of these coordinates by

$$
r \quad: \quad m(\because:+: \cdot \because=-1
$$

where $m$ is the mass of point $P$. Let $u s$ denote by $P, T$ and $Z$ the components of accleration of point $P$ in the direction of the projection of the radius
vector in the plane $x y$, in the direction perpendicular to this projection In the plane an. on the z-axis. We thus obtain

$$
\varphi_{1}=m P, \quad \theta_{2} \cdot m_{p} T, \quad \varphi_{1}=m \% .
$$

In this manner the equations of notion will be given by

$$
\begin{align*}
& d f+\binom{d \prime}{d f^{2}}^{2}=\rho \\
& \frac{d}{d t}\left(p=\frac{d v}{d t}\right)-p r  \tag{1}\\
& d=2 .
\end{align*}
$$



When the force function is given by $m 0$, then the previous equations may be written as

$$
\begin{aligned}
& \frac{\ddot{x}^{2} \rho}{\Delta t^{2}} \rho\left(\frac{v v}{t t}\right)^{2}=\frac{\partial 1}{\partial \theta} \\
& a\left(\rho^{2} \frac{d r}{a i}\right)=\frac{i l}{i!} \\
& \frac{d^{2} z}{1.1}-\frac{1 / 1}{1!}
\end{aligned}
$$

Equations (1) have been used by several authors to study the motion of the Moon. In this case, instead of coordinate $z$, the following quantity is introduced

$$
\text { , } \quad \frac{!}{\because}
$$

which represents the tangent of the MOON"S LATITUDF.
If the perturbation of motion of point $P$ is taken into account, then

$$
\begin{equation*}
\| \because \frac{k^{3}}{r}: k \tag{3}
\end{equation*}
$$

Here the first term corresponds to the attraction by the sun and the second term to the perturbation -unction. We notice here that the coefficient $k^{2}$ must be replaced by $k^{2}(1+m)$ if mass $m$ of planet $p$ cannot be neglected in comparison with the mass of the sun, which is assumed to be equal to unity. Substituting equation (3) into equations (2), we obtain

$$
\begin{align*}
& d t\left(r^{2} \frac{d v}{d t}\right)=\begin{array}{l}
d / t \\
d v
\end{array}  \tag{1}\\
& d z+k^{2} \quad \begin{array}{ll}
d / 2 \\
r^{3} & d z
\end{array}
\end{align*}
$$

These equations are appiied in the calculation of the perturbation of the planets and comets using the methods of numerical integration of differential equations.

## 8. The Clepo-Laplace equaitions

Let us consider the case when the perturbation function $R$ in equations (4) vanishes. In this case, the unperturbed motion takes place in the invariable plane passing by $S$. Choosing this plane as the $x y-p l a n e$, we set $Z=0$ and $\rho=r$. Then, the equations of notion read

The general solution of these equation is given (vol. 1, Ch. II) by the well-known formulae

where $a, e, v_{0}$ and $t_{0}$ are arbitrary constants. The inspection of these formulae indicates that it is easier to express $r$ and $t$ by functions of $v$ rather than express $r$ and $v$ by functions of $t$. It is therefore advisable to choose the longitude $v$ as the independent variable in
equations (4) instead of the time $t$. Assuming that the perturbed motion is slightly different from the unperturbed, we expect that this replacemeat will also simplify the solution in the case of perturbed motion.

When the radius vector $r$ is expressed by a function of $v$, it satisifies a rather difficult equation, obtained by excluding $t$ from' equations ( $4^{\prime}$ ). On the other hand, the inverse quantity

$$
\left.\left.u-\frac{1}{r} \frac{1}{a\left(1-e^{-}\right)} \right\rvert\, 1+e \cos (i)-v_{0}\right) \mid
$$

satisfies a very simple equation, namely

$$
d \cdot u+i u=\frac{1}{a(1--c)} d .
$$

Taking this into consideration, let us rearrange the equations of notion given in the previous section, choosing
and $t$ as the unknown quantities and the longitude $v$ as the independent variable. Assuming

$$
;_{d t}^{2} d \prime
$$

so that

$$
\begin{array}{ll}
d v & H u^{\prime \prime}, \\
d l &
\end{array}
$$

we easily replace the derivatives with respect to $t$ by derivatives with respect to v. Since

$$
\begin{aligned}
& d_{b} \quad d\binom{u}{d i} \quad \| u^{2} \quad d\left(\begin{array}{l}
d u^{2} \\
d v \\
d v
\end{array}\right)= \\
& =-11 u^{2}: d ⿲\binom{d \prime \prime}{d v}=
\end{aligned}
$$

the first of equations (1) becomes

The second of these equations gives

$$
\left\|_{d!}^{i \| l}=\right\|: \%
$$

$$
1.1
$$

Substituting into the third of these equations $z=s u^{-1}$, and noting that
we obtain

$$
H^{2} u^{:}\left(\begin{array}{l}
d^{\prime} s \\
d v^{2}
\end{array} \vdots s\right)-H_{d y}^{d H} u_{d,}^{d s}=Z-P_{s},
$$

where the second derivative with respect to $u$ is eliminated using equation (4"). Applying equation (5) to exclude the derivatives of H , we finally obiain the equations of notion in the form

Integrating these equations, we express the quantities $u$, $s$ and $H$ in terms of functions of $v$. We still have to determine the time $t$. For this purpose, we use the following equation

$$
\begin{equation*}
d t=\left\|^{\prime}\right\| \tag{7}
\end{equation*}
$$

It is easy to exclude the auxiliary function $H$ from the above equations. Actually, it follows from equation (5) that

$$
H z==h^{2}+2 \int \pi u{ }^{\prime} d v
$$

where $h$ is a constant of integration. Then equations (6) and (7) may be replaced by

$$
\begin{align*}
& \left.\left(n^{2}+2 \int T u^{: a} d v\right)\left(\begin{array}{l}
d^{\prime} u \\
d \nu^{2}
\end{array}+u\right)=u:\left(\ldots-p-1 u^{\cdot 1} \begin{array}{l}
d u \\
d \nu
\end{array}\right)\right) \\
& \left(h^{2}+2 \int T u{ }^{-s} d v\right)\left(\begin{array}{l}
d^{2} s \\
d v^{2}
\end{array}+s\right)=-u^{-2}\left(-p_{s}-T_{d v}^{d s}+z\right)  \tag{8}\\
& d u^{-}=u^{-1}\left(n^{2}+2 \int T u^{-3} d v\right)^{\prime} \quad .
\end{align*}
$$

We finally rewrite the equations obtained for the case when the force function ${ }^{\circ}$ is present. In this case

It follows from the euqallty

$$
\|(u, n, s) \cdot u\left(\frac{1}{!}, n, \frac{z}{\vdots}\right)
$$


that

Hence, we condlue that

$$
\begin{align*}
& d_{d u}^{d \prime} \cdot u^{3}\left(n^{2}+2 \int u=a l d d u\right) \tag{9}
\end{align*}
$$

These equations were first obtained by Laplace, although the principal idea of using the longitude as the independent varłable was due to clepo. Clepo derived equation (9) for the particular case of $s=0$. He applied them to study the perturbation produced by the Sun on the motion of the moon assuming that the moon is moving in an ecliptic plane.

Equations (6) were widely used by Adams in his contributions to the theory, of the. moon's motion.

## 9. Application of the Clepo-Laplace equations to the study of motion

## in a resisting system.

Let us assume that a planet moving around the sun is subject to a resistance of magnitude $\propto \mathrm{mVr}^{-2}$, where m is the mass of the planet, $V$ its velocity, $r$ its distance from the sun and $\mathcal{\alpha}$ is a small coefficient constant. Let the direction of this resistance be alcng the tangent to the trajectory of the planet and opposite to the direction of its motion. The motion of the planet is evidently in a plane. Choosing this plane to be the $x y$ plane, we rewirte equations (6) as follows

$$
\begin{gather*}
d^{2} u \\
d v^{2}+u-=H^{-2} u^{-3}\left(-\Gamma-\Gamma u^{-1} d u\right.  \tag{10}\\
d v) \\
H \frac{d H}{d u} \cdot \Gamma u^{-1}, \quad \frac{d v}{d t}=H u^{2}
\end{gather*}
$$

where

$$
r \cdot u^{-1}
$$

Let us evaluate the components of acceleration caused by the resistance of the medium in the direction of the radius vector and along the peryendicular to the radius vector in the orbital plane. The cosines of the angles between these directions and the positive direction of the tangent to the orbit are respectively equal to

$$
r V^{-1} \text { and } r V^{-1}
$$

where

$$
b^{\prime}=1^{\prime} \overline{r^{2}} r^{2} r^{2}
$$

Then, the components of acceleration will be

$$
\text { ar }{ }^{2} \dot{r}=-1 \cdot x / h u^{:} \begin{aligned}
& \dot{d} u \\
& d u
\end{aligned}
$$

and

$$
-x r^{-1} v=\ldots-a H \mu
$$

Consequent1y

$$
P=-k^{2}(1-1-m) u^{2}+a / / u^{d u}{ }_{d u}^{d u}
$$

Substituting these expressions in equations (10), we obtain

$$
\begin{aligned}
& d H \\
& d v-1,
\end{aligned}
$$

and thus

$$
\begin{equation*}
\because \quad H=-n \quad . \quad \pi \tag{11}
\end{equation*}
$$

where $h$ is a constant of integration. Furthermore, the first of equations (10) gives

$$
d: u+: a \cdots k^{\prime}(1+m) \|^{-}
$$

When the coefficient $\alpha$ is so small that terms of order $\alpha^{2}$ are negligible, we obtain

$$
d:\|\cdot 1 \cdot\| \cdot k \cdot \|^{\cdots}\left(1 \cdot 2 x n^{-1}(\cdot) .\right.
$$

We have dropped the factor $(1+m)$ in writing the above equations, since this factorcan always be included in the coefficient $k^{2}$.

The general integral of the last equation takes the form

$$
\begin{equation*}
u=k^{2} h \quad\left|1+2 x h^{\prime} v+=\cos (v \quad \cdots)\right| \tag{12}
\end{equation*}
$$

where $e$ and $\omega$ are arbitrary constantis. We can compare the orbit given by this equation with the elliptic orbit

$$
\begin{equation*}
u=p_{u}^{-1}\left[1-\left|-c_{0} \cos \left(11-r_{u}\right)\right|\right. \tag{13}
\end{equation*}
$$

which describes the motion of the planet in the absence of the resistance of the medium. Evidently, we can replase the coustant elcments $p_{0} z_{0}$ and
To in equation (13) by functions of $v$ so that this equation becomes identical to equation (12). This can be acilieved in the zollowing way: Let the coordinates and their velocities take the values $u_{0}, v_{0}, u_{0}$ and $v_{0}$ at time $t=t_{0}$. We then evaluate the elements o: the elliptic motion in terms of these values (vol. I, ch. IV). This would be the motion of the planet if the resistance of the medium was absent at the mozent $t_{0}$. Such a notion is called osculator' 'i.e. touching) relative to the motion under consideration. The elements $p_{0}, e_{0}$ and 770 corresponding to this orbit are called osculating elements for moment $t_{0}$. Let us derive expressions for the osculating elements as functions of $v$. We note that; in the moment $t_{0}$ the quantities

$$
\|-r^{2} \dot{b} n_{d i}^{d u}=\dot{d} \dot{u}^{-1}
$$

should have the same values for both the real and osculating motions. For the osculatory (elliptic) motion we have

$$
H=k V^{\prime} p_{0}
$$

whereas, for the motion in a resisiting medium, $H$ will be given by equation (11). Hence $h^{2} p_{11}:\left(h-a 1_{0}\right)^{\prime \prime}$ :

$$
\begin{equation*}
\dot{R}=h=\quad \mu_{0}^{-1}\left(1-2 a h^{-1} v_{0}\right) . \tag{1.1}
\end{equation*}
$$

Substituting here the values of $u$ and $\frac{d u}{d t}$ at $v=v_{0}$, given by equations (12) and (13), we obtein

Taking equation (14) into account, and limiting ourselves to the first powers of $e<$, we obtain

$$
\begin{gathered}
c_{11} \cos \left(u_{0} \cdot r_{0}\right)=\left(1-2 x h^{\prime} u_{0}\right)=\cos \left(u_{0}-\cdots\right) \\
c_{0} \sin \left(l_{0}-\pi_{0}\right)-\left(1-2 x h^{1} v_{10}\right) e \sin \left(v_{0} \quad(1,1)-9 x h^{-1} .\right.
\end{gathered}
$$

Combining these equations, we finally obtain

$$
\begin{gathered}
c_{0} \cos \left(\pi_{0}-\omega\right)=\varepsilon-2 x h^{-1} \epsilon l_{0}-2 x h^{-1} \sin \left(l_{1}-\omega\right) \\
c_{0} \sin \left(\pi_{0}-\omega\right)=2 x h^{-1}=\cos \left(i_{n}-\omega\right) .
\end{gathered}
$$

When $\alpha=0$, these relations become

$$
e_{1}=\cdots \varepsilon, \quad \pi_{0}==\omega,
$$

Therefore, within accepted accuracy limits, the above relations may be replaced by

$$
\begin{gathered}
c_{0}=\varepsilon-2 x h^{-1}\left|\because v_{0}+-\sin \left(v_{0}-w\right)\right| \\
x_{0}=w!\cdot 2 a h^{-1} r \cos \left(\nu_{0}-w\right) .
\end{gathered}
$$

These formulae show that the peribeiion longitude $\mathbb{T}_{6}$ of the osculating orbit is a periodic function of $v_{0}$ and consequently of $t_{0}$. The elenent $\epsilon_{o}$ will vary not only periodically but also secularly, owing to the presence of a term proportional to $v_{0}$.

Increasing $v_{0}$ by $2 \pi$, the eccentricity $e_{0}$ is increased by

$$
\begin{equation*}
\therefore \quad-1 \pi n^{-1} c=1 \pi h^{\prime} \because \tag{!}
\end{equation*}
$$

In other words, in the presence of a resisting medium acting according to the above mentioned law, the eccentricity will decrease after each revolution of the planet by a quality equal to $4 \pi \alpha h^{-1}$. In the same way we can obtain from equation (14) stating that

$$
p_{n} \quad \therefore \quad n_{2}\left(1-\cdots h h^{\prime} r\right)
$$

that after each revolution of the planet the parameter $p_{o}$ is changed by

$$
\Delta p==-1: 2 k \quad \text { ' }
$$

since

$$
p_{11}=-u_{0}\left(1 \quad-c_{0}\right)_{1}
$$

then, assuming that these elements are infinitesimal, we ebtain

$$
\begin{array}{ccc}
\Delta p & \Delta a \\
p_{v} & a_{11} & \cdots e, ~ \\
1-e_{i 1}^{2}
\end{array} .
$$

Therefore

$$
\begin{aligned}
& 1 a \quad \text { Iri } 1: r_{i}^{2} \\
& \text { い, } \quad 1 \quad \therefore \quad \therefore
\end{aligned}
$$

Concluding, let us find the corresponding variation of the abege daily motion defined;by

$$
n, \cdots\left\{a_{0} \quad \stackrel{s}{\vdots} .\right.
$$

We obtain

Thus, if a planet or comet moves in a medium whose resistance is linearly proportional to the velocity and inversly proportional to the square of the distance from the sum, then the eccentricity of the osculating orbit decreases and the average daily motion increases. The magnitudes of these changes are given in the first approximation by equations (15) and (16).

## CHAFTER III

## THE METHOD OF VARIATION OF ARBITRARY CONSTANTS

10. The osculating elements

Let $u s$ denote by $x, y$ and $z$ the heliocentric elliptic coordinates of a planet (or comet), $P$, having mass $m$. If the only force acting on this planet is the gravitational force of the sun (whose mass is set equal unity), then the equations of motion, derived in $\$ 3$, are given by

$$
\begin{array}{l:l}
x: k^{\prime}(1-m) x r=0 \\
\ddot{y}: R^{2}(1-f-m) y r^{-:}=0  \tag{I}\\
\ddot{z-1} k^{\prime \prime}(1-m)<r^{-}=0 .
\end{array}
$$

In order to simplify, we shall replace the term $\mathrm{k}^{2}(1+m)$ by $\mathrm{k}^{2}$. Due to the fact that the factor $(1+m)$ is always accompanied by $k^{2}$, it can always be included when necessary.

If a force, $\mathbb{D F}$, having components $m F_{x}, m F y$ and $m F_{z}$, acts on the planet in addition to the sun's gravitational force, equation (1) should then be replaced by the following equations of motion

$$
\begin{align*}
& \bar{i}+\cdots r^{-\quad i} \tag{2}
\end{align*}
$$

The motion determined by equations (1) is called the uperturbed or Kepler motion, whilst the motion described by equations (2) is called the perturbed motion. In the perturhation theory, one usually deals with motion along an approximatel.y elleptic orbit. This is why the unperturbed motion is sometimes called the elleptic motion.

The complete salution of equation (1) is well known. In order to be precise we will limit ourselves to a motion along an ellipse. We then express the solution in the following way:

$$
\begin{align*}
& \pi=n\left(t-t_{0}\right)+H_{0}  \tag{3}\\
& n \quad k J^{-\frac{1}{=}}  \tag{4}\\
& r-e \sin E \cdot A  \tag{इ}\\
& r=a(1-e \cos (i) \tag{i}
\end{align*}
$$

$$
\begin{align*}
& \text { 4.... } v-f-u \tag{7}
\end{align*}
$$




These equations express the unknown coordinates $x, y$ and 2 in functions of time and six arbitrary constants $a, e, M_{0}, \infty, \Omega$ and $j$.

A similar integration of the equations of perturbed motions, eqs. (2), in terms of known functions is not possible. Therefore, one has to solve these equations in a different manner. One often makes use of the fact that the perturbing accelerntion is in most cases considerably less than the acceleration caused by the gravitation of the sun. One can then study the perturbed motion using the method of successive approximations. The unperturbed motion is taken as the first approximation, then, by adding corrections ("perturbations" or "inequalities") to it, one gradually approaches the correct description of the real motion. The application of this method simplifies essentially the appropriate choice of thefunctions of time that determine the motion. It usually happens that it is more useful to use instead of the unknown functions $x, y$ and $z$, other quantities that can decermine the position of the ilanet. Tin particular, the osculating elements of the orbit can be used for this purpose.

$$
\begin{align*}
& \text { Equations (3) }- \text { (8) giving } \\
& \qquad \begin{array}{l}
x \cdots f_{1}\left(i, a, c, H_{0,}, \cdots, \cdots, i\right) \\
y \cdot f,\left(t, a, r, . l_{,}, \cdots, \cdots, i\right) \\
: f,(f, a, r, u, \cdots, \cdots, i)
\end{array}
\end{align*}
$$

represent an elliptical:motion when the quantities $a, e, \ldots, \pm$ are considered to be constant numbers. However, these equations can represent any arbitrary types of motion when the quantities n, e,..., i are considered to be properly chosen functions of time. The functions $a(t), e(t), \ldots, i(t)$ representing the perturbed motion in equation (9), are called the instantaneous elements. The totality of these elements determine the instantaneous orbit of the planet $P$. Once the instantaneous orbit is known, one can evaluate the coordinates of $P$ for any subsequent moment using the formulae that describe the elliptical motion.

We have only the conditions, given by equations (9), to determine the six functions $a(t), e(t), \ldots, i(t)$. Thus, we require that these functions satisfy another three supplementary conditions. It is required that not only the coordinates but also their derivatives $x, y$ and $z$ should be expressed in terms of the instantaneous elements by the expressions obtained for the elleptical motion. These conditions can easily be obtained from equations (3)-(8) in the form of

$$
\begin{align*}
& \dot{r}=\begin{array}{c}
k \sin p \\
i p
\end{array}  \tag{111}\\
& \text { i. ivir } \tag{il}
\end{align*}
$$

where

$$
u-n: \ddot{n} \quad p \quad \alpha(1 \cdots c)
$$

Equations (9) and (12) define the quantities $a(t), e(t), \ldots$ in functions of time. We call these functions the osculating elements and the corresponding orbit, which continuously changes its direction and form, the osculating orbit. The solution of equations (9) and (12) relative to the elements is given in Vol. 1 Ch . IV. One sees here
that these equations have one and only one solution.
We see that we can apply the equations of the unperturbed motions to express the osculating elements for any moment $t$ in terms of the vaires of $x, y, z, \dot{x}, \dot{y}$, and $\dot{z}$ in this moment. Therefore, the osculating elements can be interpreted as the elements of that unperturbed motion which would replace the perturbed motion if the perturbing acceleration vanished at this moment.

In the following, we recall theformulae which lead to the solution of equations (9) and (12) relative to the elements. From the following relations

$$
\begin{align*}
& k \sqrt{ } / p \sin i \sin 4=y z \cdots z \\
& k \sqrt{j} \sin i \cos 9 \quad x z-z x  \tag{1;3}\\
& k \sqrt{ } p \cos i=x y-y x
\end{align*}
$$

we find the parameter $p$ and the longitude of node $\Omega$, together with the slope of the orbit 1 . The kinetic energy integral

$$
x^{2}+\dot{y}=\dot{c}+h\left(\begin{array}{ll}
2 & 1  \tag{1t}\\
r & a
\end{array}\right)
$$

gives us the semi-major axis 2 and allows us to find the eccentricity $e$ from the following relation

$$
\begin{equation*}
p: \quad a\left(1-c^{2}\right) \tag{15}
\end{equation*}
$$

We obtain the true anomaly $v$ from equation (10) and then find the perihelion distance from the noded using the relations

$$
\begin{align*}
& r \sin (11 \mid \cdot \omega) \quad \therefore \operatorname{cosec} i \tag{16}
\end{align*}
$$

which can easily be obtained from equations (8). Finally, we find the average anomaly of the epoch $M_{o}$ using equations (7), (5) and (3).
11. Differential equations for the determination of the osculating elements

In the previous section we have seen that, in order to study the motion of planet $P$, it is possible to use instead of coordinates $x, y$ and $z$, the six elements $a, e, M_{o}, \omega, \Omega$ and 1 . This change of variables is useful because the elements, which remain constant during the unperturbed motion, slowly vary during the perturbed motion, at least if the perturbing acceleration is small as compared to the acceleration produced by the sun. For this reason, the determination of elements a,e, ... using the method of successive approximations is more conventent than the determination of coordinates $x, y$ and $z$.

Let us now derive the differential equations which determine elements $a, e, . .$. i. For this purpose, we substitute in equations (2), and rewrite them as
the expressions, given by Eqs. (9) and (12), that express $x, y, z$, $x, y$ and $z$ in terms of the new unknowns $a, e, M_{o}, m, \Omega$ and $i$. However, direct substitution will lead us to a series of complicated calculations and we therefore choose an indirect method which leads us more easily to our target.

Let us assume that the following relation holds

This relation can be deduced from equations (3)-(8), (10), (11) and (12). Differentiating equatio: (18) with respect to time gives

We now consider that an unperturbed motion results when the perturbing accleration $F$ vanfshes at moment $t$. We denote the corresponding coordinates and components of velocity by $\xi^{\prime}, \eta, \bar{\zeta}, \dot{\xi}, \dot{\eta}$ and $\zeta^{\circ}$. The differential equation (1) describing this motion can be written, in anology to equations (7), 88

$$
\begin{aligned}
& \begin{array}{lllll}
d i & \vdots & d \eta & \therefore i & d i \\
d l & d t & \quad & d t &
\end{array}
\end{aligned}
$$

where

For the moment $t$ under consideration, we have

$$
\begin{aligned}
& \lambda \therefore \xi, y=\eta=\quad \because
\end{aligned}
$$

and hence

We now return back to equation (18), which evidently takes part in the unpeturbed motion. Differentiating this equation, we oltain for the case under consitevation

Subtracting this equality, term by term from the equality given by Eq. (19), we finally obtain
and we thus arrive to the following important conclusion:
Any relation, Eq. (18), between the elements, coordinates and components of velocity leads to a relation of the type of Eq. (21) between the derivatives of the elements and thecomponents of the perturbing acceleration.

The transition from equation (18) to equation (21) will, in short, be called the basic operation.

The longitudie of the node and the slope
Let us apply the basic operation to equations (13) and for abbreviation the following notation is introduce ad

$$
\frac{1}{r_{p}} F_{x} \quad r_{x}^{\prime} . \quad \frac{1}{k_{1} \bar{p}} r_{y}=F_{i}, \quad \frac{1}{k_{1} \bar{p}} F_{z}-F_{2}^{\prime}
$$

We then obtain

$$
\begin{aligned}
& \frac{1}{2} p{ }^{\prime} \cos i_{d t}^{d \prime} \quad-\ln i_{d i}^{d i} \quad x F_{v} \cdots y / y_{4}^{\prime} .
\end{aligned}
$$

Taking equations (8) into consideration, we find

In order to simplify these expressions, let us introduce the components of the perturbing acceleration along the radius vector, in the direction perpendicular to the radius vector in the plane of the osculating orbit and along the normal to the plane of the orbit. Denoting these components by $\underline{S}, \underline{T}$ and $\underline{W}$ and assuming, as before,
we obtain

The coefficients that multiply $F_{x}^{1}, F_{y}^{\prime}$ and $F_{z}^{\prime}$ in the expression of $S^{\prime}$ are evidently equal to $\mathrm{xr} \mathrm{r}^{-1}, \mathrm{yr}^{-1}$ and $\mathrm{zr}^{-1}$.

The corresponding coefficients in the expression of $T^{\prime}$ are obtained from the previous ones by the replacement of $u$ by $u+90^{\circ}$ and, we therefore finally obtain

$$
\begin{align*}
d p & =2 p r \eta^{\prime} \\
d t & \\
\sin i^{d U} & =r \sin u W^{\prime}  \tag{25}\\
d t & \\
d i & =r \cos u W^{\prime} .
\end{align*}
$$

## ORIGINAL PAGFIS <br> The semima; ax ans and the eccentricity OF POOR QUALITY

Applying the basic operaiion to the kinetic energy integral, given by Eq. (14), we obtain

$$
{ }^{\prime} \quad \quad=2 \dot{x} t_{x}+2 \dot{y} t_{y}+2 \dot{2} t_{z}
$$

Taking equations (10), (11), (12) and (22) into account, we find that

$$
\frac{d a}{d t}=2 a^{\prime} e \sin v . S^{\prime}+2 a^{2} p r^{-\cdot} T^{v}
$$

Substituting in euçation (15) we obtain

Taking into account equation (6) as well as the following equation

$$
\begin{equation*}
\text { or }{ }^{1}: 1 \text { i ecoso. } \tag{27}
\end{equation*}
$$

which follows from the equation of the orbit, we obtain

$$
\begin{equation*}
d t-p: \| \frac{1}{d}+p(\operatorname{con} \cdot+\cos E) \Gamma \tag{28}
\end{equation*}
$$

## The perihelion distance from the node

We now apply the basic operation to the second set of equations and we obtain

Where (dv/dt) denotes the derivative corresponding to the dependence
of $v$ on time that enters only through the osculating elements ${ }^{(1)}$.
Substituting here expressions (8) for $x$ and $y$, we obtain.

$$
d(d) \cdot \cdots\binom{l d}{d t} \quad \cos \cdot{ }_{a}^{d}
$$

In order to find (dv/dt), we turn to equation (10). Excluding the eccentricity e from this equation, we rewrite it as

$$
\dot{\operatorname{ctg}} 1=-\frac{k}{V p} e \cos u
$$

or, by using equation (27), as

$$
\operatorname{r\operatorname {cts}u} \underset{\gamma}{k V^{\prime}}-\beta_{p}^{\beta}:
$$

We apply the basic operation to this equation and due to the fact the $r$ is defined by the factor

$$
\text { rr } x x: y \dot{y} \mid z z .
$$

we obtain

$$
\mathbf{S} \operatorname{ctg} v-\frac{i}{\sin ^{2} v}\binom{\dot{c} v}{d u}=\stackrel{k}{2 v^{\prime} p r}\left(1+\frac{r}{\dot{p}}\right){ }_{u t} .
$$

Using equation (23), we obtain

$$
e^{\prime}\binom{d v}{d t}=p \cos \nu s^{\prime}-(r+\rho) \sin v r .
$$

## Consequently


(1) J.t is necessary to note that, in contrast to the radius vector $r$, the true anomaly $v$ cannot be considered as a coordinate. In fact, $v$ not only depends on $x, y$ and $z$ but also on $x, y$ and $z$.

$$
\begin{equation*}
e_{d l}^{d \|}=-p \cos v S^{\prime}+(r \mid p) \sin v r^{\prime}-e \cos i_{d!}^{d!} \tag{29}
\end{equation*}
$$

## The mean anomaly of the epoch

Let us now apply the basic operation to equations (5) and (6). Ne obtain

$$
\begin{aligned}
& \binom{d!1 t}{t!t}=(1-e \cos i)\binom{d i t}{d t}-\sin t i_{d e}^{d t} \\
& \begin{array}{l}
r d a \\
d d
\end{array} d \cos t:_{d e}^{d l} \mid u z^{\prime} \sin t:\binom{d t^{\circ}}{d t} \\
& 10 .
\end{aligned}
$$

Here, we denote by $\left(\frac{d M}{d t}\right)$ and $\left(\frac{d E}{d t}\right)$ the derivatives of the parts of $M$ and $E$ which depend on the variations of the osculating elements. Eliminating $\left(\frac{d E}{d t}\right)$, we obtain

Substituting here the values of hederivatives of the eccentricity and semimajor axis, obtained above, we find

In order to simplify the coefficient $T^{\prime}$, we use the following equaticns

$$
\begin{align*}
& r \sin v \cdot a!/-e^{3} \sin t \\
& r \cos v=a(\cos t \cdot e) \tag{30}
\end{align*}
$$

Substituting equation (6) in the second set of these equations, we obtain

$$
r \cos u \cos E \cdots \cdot a \cos ^{2} E \cdots a e \cos t \cdots r-a \sin ^{2} r
$$

Elliminating sin $E$ by means of equations (30), we obtain

$$
\text { cusucos.. } 1 \quad \begin{array}{r}
\text { p } \\
\text { illi" ". }
\end{array}
$$

Using these relations, we finally obtain

$$
\underset{1-c}{e}\binom{d M t}{d t} \cdots(p \cos 11 \cdots 2 e r) s-(r+p) \text { sin: }: 7
$$

We now apply our basic operation to equation (3). We find

$$
\begin{equation*}
\left(\frac{d M t}{d t^{\prime}}\right)=\frac{d \hat{d} t_{0}}{d t}+\left(t-t_{0}\right) \frac{d n}{d t} \tag{31}
\end{equation*}
$$

The complete differentiation of the same equation gives


$$
\begin{array}{cc}
d M & d M r \\
d r & \ddots
\end{array}+(l-u d n: n .
$$

Therefore

$$
d M=\binom{d .:}{d t} ; n .
$$

Integrating this equation from $t_{0}$ to $t_{1}$ we obtain

$$
\left.M(t) \cdot M_{0}\left(t_{0}\right) \cdot \frac{\int_{t}^{1}(d, d t}{d}\right) d t!\int_{t_{0}}^{i} n d t .
$$

or, expressed as

$$
M(t)=n(a) \quad \dot{\square} \cdot!
$$

where

$$
H_{1}(t) \cdots M_{1}\left(t_{1}\right)+\int_{t_{1}}^{1}\binom{d!}{u} d t
$$

The osculating elements $M_{0}(t)$ and $n(t)$ are functions of time. We use he notation $M_{0}\left(t_{0}\right)$ to stress that the quantity $M_{0}(t)$ corresporils to the epoch of osculation $t_{0}$.

If we cvaluate the position of the planet for an arbitrary noment $t$. we can fint the corresponding mean anomaly $M(t)$ using the equation of elliptical motion

$$
\begin{equation*}
A(t)==n_{0}(t)+n(t)\left(l-\quad t_{1}\right) \tag{31}
\end{equation*}
$$

If follows from equation (31) that the derivative of the osculating element $M$ is given by

$$
\frac{d M l_{11}}{d t}=\binom{d M /}{d t}-\left(t \quad t_{0}\right)_{d l}^{d n}
$$

where, on the basis of equation (4),

$$
\begin{gathered}
d n \\
d t
\end{gathered}=-\begin{gathered}
3 \\
2
\end{gathered} a^{2} \vee a_{a} d t=-\begin{array}{lll}
3 & n a z \\
2 & a d t
\end{array}
$$

or

$$
\frac{d n}{d l}=-3 n a e \operatorname{sil} v S^{\prime}-3 n a p r{ }^{\prime} 7^{\prime}
$$

We see that thederivative of the element $M_{0}(t)$ includes terms relative to time. Hence, this element varies rapidly regardless to the degree of smallness of the $S^{\prime}$ and $T^{\prime}$ factors. This situation leads to a great deal of difficulties in the evaluation of perturbations. It obliges us to use equation (32) and nct equation (34) in order to evaluate the mean anomaly. The function $M_{0}$ involved in this formula is defined on the basis of equation (33) by the following relation

$$
\begin{equation*}
d_{d t}=\frac{M_{0}}{c}(p \cos v-2 c r) S^{\prime}-l^{\prime} 1 \cdots \iota^{2}(r ;-p) \cdot i!v^{\cdots} \tag{35}
\end{equation*}
$$

Equations (32) and (35) are fin practice commonly used. To simplify rheir final forms, we shall denote $\bar{M}_{0}$ simply by $M_{0}$ as far as this dnes not cause any confusion, and once and for all agree to use equation (32) to evaluate the mean anomaly.
12. Comparison between the diffecent formulae

In most practical cases, the slopes of the orbits, 1 , are very small. In this case, some of the formulae derived in the previous two sections are preferably replaced ly others.

Instead of the peribelion distance from the node, we introduce here the perihelion longitude $\pi$ by the following equation.

Then, equation (29) gives


Using equation (24), we transform the first term on the right-hand side of the previous equation into

This term can only decrease if the value of $i$ decreases, whereas the corresponding term in formula (29) will still be large.

Furthermore, we introduce the mean longitude in the orbit by

$$
\text { , }=: \because \quad \because+\cdots \quad M .
$$

which will simply be called the mean longitude. We denote tine mean longitude corresponding to the initial epoch $t_{0}$ by $\lambda_{0}$, so that

$$
\lambda \begin{array}{cc:c}
y_{0} & \pi & : i_{u} \\
t_{1} & n(t-t)
\end{array}
$$

Taking into account equation (32), we obtain

$$
\begin{equation*}
x \cdot c+i_{0}^{n} n \cdot \tag{1.31}
\end{equation*}
$$

where

$$
\text { \& } \quad \therefore \because i l(1) \text {, }
$$

Hence

The quantity $\in$ will also be called the average longitude of the epoch, hoping that this will not lead to any misunderstanding.

We now combine together all the differential equations which determine the osculating elements:

$$
\begin{align*}
& d e=p \sin u S^{\prime}+p(\cos v+\cos t) r^{\prime} \\
& \sin i_{d!}^{d!}=r \sin u I V \\
& d_{i}=r \cos u W^{\prime} \tag{37}
\end{align*}
$$

$$
\begin{aligned}
& \because(r!\mu) \mathrm{sin}: \%
\end{aligned}
$$

where $\quad S=\begin{gathered}1 \\ . v i\end{gathered}, \quad, \quad: \quad 1 . \quad W^{\prime}={ }_{k i ;}^{1} W$,
and $\underline{S}, \underline{T}$ and $\underline{W}$ denote the components of the perturbing accelcration. Since the average longitude is given by equation (36), the following equation should be addec to the previous ones

The integration of this system of differential equations gives the values of the osculating elements in moment $t$. The position of the planet in this moment is obtained by the usual formulae. Starting with equation (36) and
one finds $r$ and $u$. Then, using equations (8), one obtains $x, y$ and $z$.

## 13. The Lagrange equations

In the previous o ction we did not impose any limitations on the perturbing acceleration $F$. We now assume that the accleration is caused' by a force for which a potential exists. In other words, we assume that there exists a function $R$ ruch that

For example, in equation (1) of Chapter $I$, the pecturbation to the motion of one of the planets, caused by' the presence of the others, is expressed, by the perturbation function.

Let us transform the hasis: equations (37) so that they inciude the partial derivatives of the function $R$ instead of the components of the perturbing acceleration $S, \underline{T}$ and $\underline{W}$. For an arbitrary element $a$, the following relations hold
where equathous (22) expressed a:

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can be used to express $F_{x}, F_{y}$ and $F_{z}$ in terms of $\underline{S}, \underline{T}$ and $\underline{W}$. The evaluation of the derivatives of the coordinates with respect to the elements, $\frac{C X}{\delta a}, \frac{\partial y}{\partial a}, \ldots, \frac{\partial x}{\partial x}, \ldots$, is straightforward. 'We first find the derivatives of $r$ and $u$, namely


We then differentiate equations (8) with respect to the elements. In doing so, we should remember that each of the coordinates depends explicity on $\Omega$, together with the relationsihip $u=v+\pi-\Omega$. Similarly, whilst differentiating wit $k$ respect to $\pi$, we must take into account that $u$ depends explicity on $\pi$ and to $v$, since $v$ is a function of

$$
11 \cdot j_{6}^{1} 14!:-r
$$

Therefore,

$$
\begin{gathered}
d \\
d x
\end{gathered}=\frac{1}{d \prime \prime}-d_{6} .
$$

We finally obtain

$$
\begin{aligned}
& \begin{array}{ll}
d R & =n a^{2}
\end{array} \quad p \cos u S^{\prime}+(r+p ; \sin u T i \\
& \begin{array}{ll}
d K \\
d_{1}-n a^{2} \\
V & -e^{2} r \sin u
\end{array} W^{\prime}
\end{aligned}
$$

It remains to substitute the values of $S^{\prime}, T^{\prime}$ anc' $W^{\prime}$ obtained by these equations into equations (37). In equations (37) the following combinations are present
as well as the following combinations

Substituting these expressions into equations (37) and (38), we obtain after some manipulations


ORIGINAL PAGL OR POOR QUALIII Combining the first and second equations, we obtain
which also follows from equations (23). Subsequently, we shall call equations (41), the Lagrange equations.

It is important to note that, in evaluating derivative $\frac{\partial R}{\partial a}$ involved at the end of equations (41), we ignored equation (4) and orly considered the explicit deperdence of $R$ on $a$. This is the method by which equations (40) were derived, and based on these equations, the entire present deduction was developed.

In conclusion, we note that the Lagrange equations obtained here have the following properties:
(1) Time enters the Lagrange eduations only through the derivatives of the perturbation function $R$.
(2) The elements of the orbit are divided into two groups, one consisting
of $a, e$ and $i$ and the other of $\Omega, \pi$ and $\epsilon$. The differential equations, which determine the elements of one of these groups, include the partial derivatives of $R$ with respect to only the elements of the other group.
(3) Let $\alpha$ and $\beta$ be two elements belonging to different groups. If $\frac{d \alpha}{d t}$ contains $\frac{\partial R}{\partial \beta}$, then $\frac{d \beta}{d t}$ will ontain $\frac{\partial R}{\partial \alpha}$, where the coefficients of $\frac{\partial R}{\partial \alpha}$ and $\frac{\partial R}{\partial \beta}$ will be equal in magnitude but of different signs.
14. Another derivation of the Lagrange Equations
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The differential equations (41), or the more general equations (37), are some of the corner-stones of celestial mechanics. Therefore, it is interesting to investigate all questions concerning these equations. We have just obtained an elementary and relatively simple derivation for the Legendre equations. ${ }^{(1)}$ Another interesting method for the deduction of these equations was suggested by Lagrange, to whom we owe the development of the method of variation of arbitrary constants. In the following, we give a brief outline of the derivation suggested by Lagrange. We shall not carry out all the calculations since we have already obtained the final equations.

Following Lagrange, and in keeping with his notations, we consider the following_system_of_equations
(1)

A derivation of the Legendre equations, having a geometrical character, may be found in: S.L. Kazakov, The method of variation of arbitrary constants, scientific Transactions of Moscow University (Sposob variacef proizvolnyh postojannyh, Ucenye Zapiski Moskovskogo Universiteta) 1905 and in: A.I. Krylov, Sur la variation des orbits elliptiques des planets, Proceeding of the Academy of Science (Collection of Transactions) 1905 vol. IV. Izvestija Akademil nauk (Sobranie trudov) . It is shown in the latter paper that one has the right to think that Newton has obtained the above equations namely on the basis of these al uments. Newton, however, published only some theorems which have no direct relation to equations (37).

$$
\begin{array}{cc}
d x-d Q \\
d f-x^{\prime}-0, & d x^{\prime}+\frac{d Q}{d t}+\frac{d x}{d}+x=0  \tag{43}\\
d y & d Q \\
d t-y^{\prime}-y^{\prime}-0, & -y^{\prime}+\frac{d Q}{d t}+\partial y+Y \ldots 0
\end{array}
$$

Let the number of these equations be $2 h$. This should equal the number of the conjugate variables $x, x^{\prime}, y, y^{\prime}, \ldots$ The quantities $\Omega, X, X^{\prime}$, $Y, Y^{\prime}, \ldots$ are functions of $t, x, x^{\prime}, y, y^{\prime}, \ldots$ We assume that we are able to integrate the following equations

$$
\begin{array}{ccc}
d x \\
d l-d x^{\prime} & 0 ; & d x^{\prime}: d y  \tag{41}\\
d y & d t & 0 \\
d t-d y^{\prime} & 0 . & d y^{\prime}, d y \\
d y & 0
\end{array}
$$

These equations are obtained from equations (43) by equating all the suppiementary functions $X, X^{\prime}, Y, Y^{\prime}, \ldots$ to zero. Let the general solution of equations (44) be given by

$$
\begin{equation*}
x=i_{i}\left(t, a, A, \ldots, \ell_{1}\right), \quad x^{\prime} \quad \therefore(t, u, h, \quad, k) \tag{1.1}
\end{equation*}
$$

which involves 2 h arbitrary constants $\mathrm{a}, \mathrm{b}, \ldots \mathrm{g}$.
The expressions given by Eqs. (45) satisfy equations (43) only if the quantities $a, b, \ldots, g$ are treated as functions of $t$. Let us find the differential equations that these functions should satisfy. We stbstitute the following equations
into equations (14), and taking equations (44) into consideration, we obtain

These are the required differential equations. Lagrange suggested that' these equations could be written in a much simpler way. He introduce the so-called Lagrange brackets $[a, a],[a, b] \ldots$ in the following way

It is easy to see that

$$
\begin{gather*}
\{a, a \mid[b, b]  \tag{17}\\
|a, n| \cdot|n, a| 1! \tag{4h}
\end{gather*}
$$

Assuming that
we can replace the system (46) by

$$
119
$$

$$
\begin{aligned}
& 1
\end{aligned}
$$

$$
\begin{align*}
& \text { iss at its ats } \tag{46}
\end{align*}
$$

$$
\begin{aligned}
& \begin{array}{l}
d y^{\prime} d a \\
i d_{d}+\frac{d y^{\prime}}{} d b \\
d i \\
d t
\end{array}+\ldots+\gamma=0 . \\
& \begin{array}{l}
\text { ORIGINAL PAGE IS } \\
\text { OF POOR QUALITY }
\end{array}
\end{aligned}
$$

In order to obtain the first equation, we multiply equations (46) in sequence by $-\frac{\partial x^{\prime}}{\partial a},-\frac{\partial y^{\prime}}{\partial a}, \cdots+\frac{\partial x}{\partial a}+\frac{\partial y}{\partial a}, \ldots .$. and add. The other equations are obtained in a similar manner.

In the light of the relations given by equations (47) and (48), equations (49) are simpler than equations (46). In particular, the third property of che Lagrange brackets is responsible for the simplicity of equations (49). When expressions (45) are substituted in a Lagrange bracket, the independent variable $t$ is eliminated according to this property. In other words

$$
\begin{equation*}
" 1 a .110 \tag{50}
\end{equation*}
$$

To prove this, let us differentiate the bracket $[a, b]$ term by term. We obtain

Taking equation (44) into account, we find

Hence

Equation (40) suggests that one can evaluate the Lagrange brackets at the value $t$ that makes calculations quite simple.

The present method will be particularly simple when we take constants
$a, b, \ldots, g$ to be the initial values $x_{0}, y_{0}, \ldots, x_{0}^{\prime}, y_{o}^{\prime}, \ldots$ corresponding to the value to the independent variable. We immediately evaluate the Lagrange brackets at $t=t_{0}$ due to the property just shown. We obtain

$$
\begin{aligned}
& \left|x_{1,}, y_{0}\right|\left\langle x_{1,1}, y_{1,}\right| \ldots(\ldots)
\end{aligned}
$$

In the same way, we obtain

$$
|y,, y| \quad 1,\left|y, \quad,|\quad| 1, x_{n}\right| \quad \ldots .
$$

and so on. In this manner, equations (49) will have the following form

$$
\begin{aligned}
& \because 1 \\
& \because i \\
& \frac{y_{0}}{a i}=+R_{y_{0}^{\prime}}^{\prime} \quad \frac{\because x_{0}^{\prime}}{\because} \quad \frac{\because 0}{\because}=-R_{y_{0}}
\end{aligned}
$$

However, these equations are not used in celestial mechanics because equations (45) become very compliccied when the constants of integration are chosen according to the above method.

Let us consider the determinant

We construct the complementary determinant $F^{\prime}$, whose elements are the algebraic complements to the corresponding elements of $F$, devided by the value of the determinant $F$. This determinant is given by

$$
\begin{aligned}
& (0, a),(a, b), \cdots,(a, 8) \\
& f^{\prime} \ldots(b, a),(b, b), \ldots,(b, r) \\
& (1, b),(s, b), \ldots, \ldots
\end{aligned}
$$

The elements of the determinant $F^{\prime}$ are called Poisson brackets. Evidently
$f: 1$.

We can express equations (49) in terms of Poison brackets as follows

We now apply the method of variation of arbitrary constants to the perturbation of the motion of a planet thet is given by equations (17). We rewrite these equations as follows

Comparing these equations with equations (43), we find here that

$$
\because=\frac{1}{2}\left(1^{\prime 2} \quad i^{\prime 2}+z^{\prime 2}\right)-\frac{k^{\prime}}{r}
$$

and thus
(52
when $R=0$, we obtain from equations (51) a system corresponding to equations (44) characterizing the uperturbed motion. In solving this latter system, we choose as the integration constants, the usual elements of the elliptic motion $a, e, i, \Omega, \omega$ and $M_{0}$. After long but relatively simple manipulations, we obtain the corresponding Lagrange brackets, which are

The corresponding poisson brackets are immediately obtained by solving the system (49). Taking into consideration equations (52), we finally obtain


In order to obtain the Lagrange eutuations in the form (41), it is sufficient to replace the variables $\omega$ and $M_{0}$ by $\pi$ and $\epsilon$ using the relations
15. 'The perturbations of the elements

The lagrange equations, from which the osculating elements are obtained, are extremely complicated equations. They can only be solved using the method of successive approximation. In the following, we are going to investigate the form in which this solution is obtained.

Let us suppose that there exist only two planets whose masses are $m$ and $m^{\prime}$. We denote the elements of these planets respectively by $a, e, \ldots$ and $a^{\prime}, e^{\prime}, \ldots$ On the basis of the results obtained in the previous section, we assume that the motion of these planets relative to the sun is defined by the following twelve equations


(.1)

In these equations (\$ 3)

$$
\begin{aligned}
& R=\kappa^{2} m^{\prime}\left(\begin{array}{cc}
1 & x_{1}^{\prime}+1 \\
1 & y^{\prime}, f z^{\prime} \\
r^{\prime}:
\end{array}\right)
\end{aligned}
$$

where $x, y, z, r$ and $x^{\prime}, y^{\prime}, z^{\prime}, r^{\prime}$ denote the coordinates and radius vectors of the planets and $\Delta$ the distance between them. It is easy to see that $R$ and $R$ ' depend only on the mutual location of the orb:its, but not on their location relative to the ecliptic. In fact, if the angle between the radius vectors $r$ and $r$ ' is denoted by $H$, then these functions will be given by

They depend only on $r, r$ and $H$ since

$$
A:=r: r^{\prime 2}-2 r r \text { is } H \text {. }
$$

The method of successive spproximations applies when the masses mand $m^{\prime}$ are small. This method yields the unknown functions in the form
of a series of expansion in powers of masses $m$ and $m^{\prime}$, such as


Here, by $a_{0}, e_{0}, \ldots, a_{0}^{\prime}, e_{0}^{\prime} \ldots$ we denote the (constant) values of the osculating elements evaluated in the initial moment, by $\delta a, \delta_{1} e, \ldots$ the functions of time that have $m$ as a multiplying factor, and by
$\delta_{1} a^{\prime}, \delta_{1} e^{\prime}, \ldots$ the functions of time that have $m$ ' as a multiplier. In general, we denote by $\delta_{n} a, \quad \delta_{n} e, \ldots, \delta_{n} a^{\prime}, \delta_{n} e^{\prime}, \ldots$ terms of the $n$-th power in masses $m$ ard $m^{\prime}$. The expressions $\delta_{n} a, \delta_{n} e, \ldots$ $\delta_{n^{\prime}} a^{\prime}, E_{n} e^{\prime}, \ldots$ are to be called perturbations of the $n^{\text {th }}$ order. Yutting in equations (54) and (54')

$$
m==0, \quad m^{\prime}==0
$$

we obtain

$$
a=a_{0}, \quad e=e_{0}, \ldots, \quad a^{\prime}==a_{0}^{\circ}, .
$$

These values will then be substituted in the right hand side of the same equations, i.e., equations (54) and (54'). Since $R$ and $R^{\prime}$ are functions of $t$ and $a, e, \ldots, a^{\prime}, e^{\prime}, \ldots$, then equations (54) and (54') will have, after this substitution, the following form

$$
\begin{aligned}
& d a=m^{\prime} f\left(t, a_{11}, e_{1}, \ldots ., a_{0}^{\prime}, \ldots\right) \\
& \quad d a^{\prime}=m / 1\left(t, a_{0}, c_{0} \ldots \ldots a_{0}^{+}, \ldots\right)
\end{aligned}
$$

Integrating these equations yfelds

$$
\begin{equation*}
a=a_{n}+m_{!_{11}}^{!} f\left(t, a_{t}, c_{0}, \ldots\right) d t=a_{0}+i_{1} a_{1} \tag{57}
\end{equation*}
$$

and, hence, determines the first order perturbations.
is obtain the second-order perturbations, we substitute into the right-hand sides of equations (54) and (54') the values $a_{0}+6 a$, $e_{0}+\delta_{1} e, \ldots, a_{0}^{\prime}+\delta_{1} a^{\prime}, \ldots$ which have been obtained for the elements $a, e, \ldots, a^{\prime}, \ldots$ This substitution yields

After integration, we obtain

$$
a=: a_{0}+i_{1} a \cdot f i_{2} a
$$

These equations are accurate up to the second order in masses. Repeating the same procedure, we may obtain as many terms in the series-expansion of the elements as we desire.

Let us now investigate the analytic form of the expansions given by equations (56) and (56'). The coordinates of the eai 'the planets are periodic functions of the corresponding mean $M$ or $M^{\prime}$. Consequently, the perturbation functions $R$ and $R^{\prime}$ are also periodic functions of $M$ and $M^{\prime}$. Hence, thoy can be expanded in a double Fourler series as follows
where $j$ and $j^{\prime}$ take all the integral values from $-\infty$ to $+\infty$
It is easy to see from equations (56) and (56') and from the expressions of the coordinates in the elliptic motion ( $S\left(\begin{array}{l}\text { 77-82) } \text {, }\end{array}\right.$
that the functions $R$ and $R^{\prime}$ and consequently the coefficients $N$ and $N^{\prime}$ can be expanded in powers of the eccentrisities $e$ and $e^{\prime}$ and the mutual' slope of the orbits J . (We have already pointed out that the functions $R$ and $R^{\prime}$ depend only on the mutual slope of the orbits). Noting that
we finally obtain both the functions $R$ and $R^{\prime}$ in the form

$$
\unlhd A c^{c} e^{\prime 2^{\prime} J^{9}} \cos D
$$

where

$$
D=j(n t+\varepsilon)+j^{\prime}\left(n^{\prime} t-\varepsilon^{\prime}\right)+C .
$$

Substituting these expressions for $R$ and $R$ ' into equations (54) and (54'), we obtain similar summations in the right-hand side of these equations.

Equation (57) shows that, in order to evaluate the second order perturbations, it is necessary to replace in the right-hand sides of equations (54) and (54') a.ll the elements a, c, ... by their initial elements and integrate the resulting trigonometric series. This yields expressions for the first order perturbations in the form of turms having one of the following two types:

$$
A_{0} e_{0}^{x} e_{0}^{\prime z^{\prime}} J_{0 j u_{0}-1-J^{\prime} n_{0}^{\prime}}^{\sin D_{11}}
$$

$j f n_{0}+j^{\prime} n_{0}^{\prime} \neq 0$, and

$$
t A_{0} e_{i 1}^{\prime} e_{v}^{\prime x^{\prime} J_{j}^{\prime} \cos C}
$$

If $j n_{0}+j ' n_{0}=0$.
Evaluating the second-, third-, ... order terms in the above way, we obtain terms of the type

$$
\begin{align*}
& t^{p} A_{0} c_{0}^{\alpha} r_{0}^{r a} f_{0}^{\prime} \cos (v t+C) \\
& \left(j_{1} n_{0}-i j_{1}^{\prime \prime} n_{0}^{\prime}\right)^{k_{1}}\left(i_{4} n_{4}-1 \cdot j_{2}^{\prime} n_{2}^{\prime}\right)^{k_{2}} \ldots . \tag{5,9}
\end{align*}
$$

where

$$
v=j n_{0}+j^{\prime} n_{x}^{\prime} .
$$

If $\mathcal{S}=0$, Such a term is then called a periodic perturbation in the case of $p \geqslant 0$, we will have a secular perturbation if $\nu=0$ and a mixed perturbation if $\boldsymbol{Y}=0$. The sum

$$
\therefore \vdots x!1
$$

is called the degree of perturbation. The higher this degree is, the smaller is the perturbation of the order under consideration.

Particular attention should be paid to the small terms in the denominator of equations (59). These terms are calied the small subgroup. There are responsible for increasing the value of the perturbations. If q is the sum of the orders $\mathrm{k} \lambda, \mathrm{k} \mu, \ldots$ of all the small sub-groups of the term given by equation (59), then the larger $q$ is the larger is the correspunding perturbation, as long as the other conditions are not altered.

Poincare called the difference $n-p$ for the $n-t h$ order perturbation the rank and the difference $n-\frac{1}{2} \mathrm{~F}-\frac{1}{2} q$, the class. Once we know the order, rank and class ofa given perturbation, we have an estimate of the general character of this perturbation. For small intervals of time, most important will. be the lowest order perturiations, and in particular the first order terms. On the other hand, for long intervals of time, the value of a perturbation is mainly determined by its class. For very long intervals of time, the contribution of the perturbation is best of all juaged by its rank.

## 16. Long-periodic perturbations

Amongst the periodic perturbations, given by equation (59), Particular attention should be paid to those, for which the coefficients $\nu$ involved in the arguments of the trigonometric functions are small. The periods of these perturbations, which are equal to $\frac{360^{\circ}}{\gamma}$, can be considerably longer than the periods of inversion of the planets under consideration. These perturbations are called long-periodic perturbations.

Long-periodic perturbations play an important role in the theory of the motion of planets. It happens that the amplitudes of some of these perturbations are very large even when their degrees are large. Let us consider a perturbation of the first order. If the term

$$
A_{1} e_{0}^{a} e_{0}^{\prime a j j_{0}^{3}} \stackrel{\sin \left[\left(j n_{1} \mid j n_{0}^{\prime}\right) t ; C \mid\right.}{j n_{v} \mid j n_{0}^{\prime-}}
$$

corresponds to a long-periodic perturbation, then the quantity $j n_{0}+j$ ' $n_{0}^{\prime}$ appearing in the denominator is small and the amplitude is much larger than what is expected for a perturbation of such a degree.

The average longitude of the planet has a strong influence on the long-periodic perturbations. The average longitude is given by the following relation ( $\$ 13$ )

$$
A .=\varepsilon+!
$$

where


To obtain $\mathcal{S}$, let us integrate twice each term on the right-hand side of this equation. This yields

$$
1, \ldots, \quad \begin{array}{llll}
1 / 11 & 1111 \\
11 n_{.1} & 111
\end{array}
$$

Thus, long-periodic perturbations contribute to the average longitude by terms of the first-order in mass, i.e. of class zero, devided by the squares of small quantities.

The detailed theory of sereis-expansion of perturbation functions will be given in one of the following chapters. This theory shows that the following relation holds ORIGRNL PAS: OF

$$
x: a^{\prime} \cdot \beta=j: j+\text { even integer }
$$

Thus, the long-periodic term can $b$ ive a considerable amplitude only when then numbers j and j ' have small absolute values.

In order to find the values of $j$ and $j^{\prime}$, for which the perturbation becomes long-periodic, it is most convenient to expand the ratio $n_{0} / n_{0}^{\prime}$ in a continued fraction. For example, for Jupiter and Saturn

$$
n_{0}=299^{\prime \prime} .12 x^{\prime} ?
$$

when the initial moment is chosen to be January 1.0, 1900.
Accordingly,

$$
\frac{n_{0}}{n_{0}}=2+\frac{1}{2+\frac{1}{14+} \ldots}
$$

The appropriate fraction may be the following

$$
\begin{array}{ccc}
2 & 5 & 72 \\
1 & 2 & \frac{72}{29}
\end{array}
$$

If we choose $j=1$ and $j^{\prime}=-2$, then

$$
j n_{0}+j^{\prime} n_{u}^{\prime}:-=58^{\prime \prime} .67 .36
$$

which approximately equals $1 / 5 n_{0}$ or $-\frac{1}{2} n_{0}^{\prime}$. Such $f$ divider cannot $b=$ considered as small. On the contrary, when $j=2$ and $j^{\prime}=-5$, we obtain

$$
j n_{0}-\frac{j^{\prime} n_{v}^{\prime}}{} \quad 4^{\prime \prime} .1110^{\prime} 9,
$$

which approximately equals $\frac{1}{74} n_{0}$ or $\frac{1}{30} n_{0}^{\prime}$. The corresponding longperiodic inequality, whose period is approximately 900 years, has in the longitude of Satum an amplitude of the order of 50'. Finally, if we consider the next fraction, we obtain

$$
j n_{0}+j n_{0}^{\prime} \cdots 29 n_{0}-72 n_{0}^{\prime}-1 " 9 \times 23 .
$$

The corresponding inequality, the degree of which is not less than $/ 29-72 /=$ 43, is completely insensible.

A large inequality in the motion of Jupiter and Saturn, depending on the subgroup $2 \mathrm{n}-2 \mathrm{n}$, was discov $\cdots$ ed empirically. Several unsuccessful attempts to interpret this ineoualıiy led Euler and Lagrange to assume the existance of an unknown type of gravitation in addition to the gravitation influenced by the Sun. The correct interpretation was given by Laplace who evaluated all the first order inequalities for the motion of Jupiter and Saturn up to the third degree.

## Annotation:

In practical studies of the motion of planets, one is rarely met with more than one small subgroup. Actually, let the zatio of mean durnal motions be expanded in a continuous fraction, so that

$$
n_{n_{1}}^{\prime \cdots}=\frac{1}{a_{1}+\frac{1}{\alpha_{2}}}
$$

Let the first appropriate fraction leading to a small subgroup be

$$
P_{k}^{p_{k}-x} \dot{a}_{a_{1}}+\cdots+\frac{1}{z_{k-1}}
$$

The next incomplete quotient $\alpha_{k}$ will be a large number. Thus, the next approprlate fraction

$$
\begin{aligned}
& P_{1,1}=P_{k} a_{k}+P_{k}, \\
& O_{k+1}=O_{k} 2_{1}-O_{k},
\end{aligned}
$$

will have a very large numerator and denominator. The perturbations that correspond to this fraction, as well as to all subsequent appropriate fractions, will be insensible.

## 17. Secular perturbations

It has been pointed out that the secular perturbations are obtained from the terms of the perturbation functions, in which the arguments of the trigonometric functions do not depend on time. The totality of these terms are called the secular parts of the perturbation function.

Any term in the perturbation function depends on and only, through the mean anomalies $M$ and $M^{\prime}$. Hence, terms in equation (58) in which $j=j^{\prime}=0$ do not depend on $E$ and $\mathcal{C}^{\prime}$. If $R_{0}$ denotes the secrilar part of the perturbation function, then

$$
\begin{aligned}
& d R_{11} \\
& d \varepsilon
\end{aligned}
$$

Referring to the first ofequations (41), we note that the expression of $\frac{d a}{d t}$ does not involve constant terms. In other words, a does not have a secular perturbation of the first order. The last of equations (41) indicates that the average durnal motion will also not have a secular perturbation of thefirst order. This result holds for the mutual perturbations of any arbitrary number of planets. It leads to the following fundamental theorem.

The semimajor axes of the orbits of plamets and their average durnal motions do not have secular perturbations of the first order relative to masses.

Laplace (1773) proved this theorem for terms of degree not higher than the second. The general proof 0 ? this theorem was given by Lagrange (1776). In the year 1809. Poisson showed that there are no pure secular terms in the perturbations of the semimajor axes and among the second
order terms ${ }^{(1)}$. In the year 1878, Spiru C. Haretu was able to find third-order secular terms.

The other elements $e, 1, \Omega, \ldots$ have secular perturbations.
For exmple, Leverrie found that for Jupiter
where $T$ denotes time in centuries ( 36525 days) counted starting from the mean value Paris midday time on January 1.0, 1900

Secular perturbations have always been connected with the stability of the solar system. However, it is necessary to point out that, even if the convergence of the series (56) could not be proved for an arbitrary time $t$, the presence of secular terms in various-order perturbations would not be sufficient for concluding that the solar system would be unstable. In fact, the expansion of periodic functions of time In powers of the mass can involve an infinite number of secular terms. For example, let us consider the function sin (mat), where $m$ is the mass of the perturbing planet and a is an arbitrary constant. Expanding this function in powers of $m$ yields

$$
\text { Mnmul } m!!-\frac{1}{1} m^{s} \|^{\prime} B ;
$$

Thus, any method of integration of equations (54) and (54'), based on the expansion of the solution in powers of the perturbing masses, will lead to secular terms, even if the solution is expressed in terms of periodic functions of time.
(1)

This result is known as the Poisson theorem.

## 18. Poisson's method

When the secular perturbations of the angular elements $\Omega, \pi$ and $\in$ are large, it is sufficient to apply the method of integration suggested in $\$ 15$ because of the slow convergence of the successive approximations. This situation is met with in the theory of 1 unar motion. There, the secular perturbations produce variations in the perinelion and node longitudes. These variations are given by (2)
where $T$ denotes time in centrueis of the average Paris astronomic time, starting from January 1.0, 1900.

Poisson (1835) su gested a special method for the integration of these equations. This method is to include in the first approximation the contribution of thesecular perturbations of the angular elements to the periodic inequalities. Denoting the average longitudes of the planets by

$$
{ }^{\prime}=i p . \quad \lambda^{\prime} \varepsilon^{\prime} \cdot s^{\prime} \text {. }
$$

where

$$
\because \cdots \dot{j}^{\prime} n d \| . \quad o^{\prime} \cdot j^{\prime} n^{\prime} d!
$$

we rewrite equations (54) $\mathrm{ar}^{\prime}$ ( $54^{\prime}$ ) in the following general form
(1) The numerical coefficients in these equations are taken from the table of Radon (see 117). Heisen's (Gajzen) tables yield



$$
\begin{aligned}
& \frac{d d}{d!}=m^{\prime} F\left(:+i, \Omega, r, \ldots, s^{\prime}+p^{\prime}, Q^{\prime} . \quad .\right)
\end{aligned}
$$

. Let

$$
0-h+a m^{\prime} t, \quad r=\pi: \quad!m^{\prime} t . \quad \varepsilon-e_{1}+m^{\prime} t
$$

where $\alpha, \beta$ and $\gamma$ are arbitrary constants and $\theta, \pi$ and $\in$ are the new unknown. Substituting these expressions in the previous ones, we obtain

As we have done in $S 15$, let us put
where $\varepsilon_{n}$ denotes terms proportional to the $n$-th power of masses $m, m^{\prime}, \ldots$ We substitute these expressions in equations (60) and expand the functions $\Phi, \ldots$ in series in the following way

This means that we are going to take as the incremencs of the arguments on1y the periodic terms $\delta_{1} G, \delta_{2} \in, \ldots \delta_{1} \Theta_{,} \ldots$ and not the secular elements m't, m't ans m't. We obtain in the first approximation

$$
(d)
$$



Integrating these equations, and equating to zero the secular perturbations of $\theta, \pi$ and $\in$ yields three equations for the determination of the quantities $\alpha, \beta$ and $\gamma$.

We note that the integration of equations (60) has been made in a way as simple as the method of succe sive approximation. The reason is that the elements $\leq, \Omega$ and $\pi$ appear in the expansion of the perturbaiion function only in the arguments of the trigonometric functions. Actual1y, since the perturbation function is a periodic function not only of $\lambda$ and ' $\lambda$ ', but also of $\Omega, \Omega$, $\Pi$ and $\pi$ ', then its $\cdots{ }^{\prime}:$ nsion is given by

In the Poisson method, one is not very strict on the expansion in powers of the masses, since in the arguments of the periodic perturbations there will be terms like $<m^{\prime} t, \beta m^{\prime} t, \ldots$ having masr m'as a multiplier. In other words, part of the second-order terms will be taken into account in the fir: $=$ approrimation.

## CHAPTER IV

## THE CANONICAL ELEMENTS

19. The canonical equations

Let us consider the motion of a system of $n$ material points. $W^{*}$ denote their masses by $m_{i}$ and coordinates by $x_{i}, y_{i}$ and $z_{i}$. We assume that the interaction between tne particles is described by the force function $U$. The motion of the system will then be described by the following equations

$$
\begin{aligned}
& \text { (i . : 0.1. . . . . } n-1 \text { ) }
\end{aligned}
$$

There are several methods to replace these $3 n$ second-order differential equations by $6 n$ first-order equations. The following method is of particular importance. Let us introduce the following notations

$$
x-m_{1} x_{1}, \quad y_{1}-m_{1} y_{1}, \quad z_{1} \quad m_{6} x_{1}
$$

The 'ine*ic ...ergy of the system will then be

$$
\begin{aligned}
T & \left.=1 \sum_{m_{1}\left(x_{1}:\right.}!y_{1}^{:} \ldots:\right)= \\
& =1 \sum_{m_{1}}^{1}\left(x_{1}^{\prime} \cdots_{i}-y_{1}^{\prime}+z_{1}\right.
\end{aligned}
$$

It is sasy to see that equations (1) are equivalent to the following equations

$$
\begin{aligned}
& d x_{1}=d H \quad d x_{1}=-d H
\end{aligned}
$$

where

$$
H-T-U
$$

is the total energy of the system, since the force function is equal to the negative of the potential energy. These equations are known as the canonical equations. The function $H$ is called the Hamiltonian of the system.

Let us now consider the more general case, when the positions of the points of the system are defined by the $S$ parameters $q_{1}, q_{2}, \ldots q_{s}$, which may be subject to a number of holonomic constraints. These parameters are called the generalized coordinates. The value of $S$ defines the number of degrees of freedom of the system. In this case, the equations of motion (1) are transformed into the following Lagrange equations


where
is the Lagrangian of the system.

We note that when we express the rectangular coordinates in terms of the goneralized coordinates, the kinetic energy becomes

$$
\begin{equation*}
T:=\frac{1}{2} \sum_{1}^{N} \dot{U}_{1 k} u_{1} q_{k}: \sum_{n}^{N} i_{1} \varphi_{n}-i-A, \tag{.2}
\end{equation*}
$$

where $A_{i k}, A_{k}$ and $A$ are functions of $q_{1}, r_{2}, \ldots, q_{8}$ and $t$. The summations over all of the indices are carried from 1 to $S$.

We can now show that equations (1) can be replaced by firot-order equations having a canonical form. We introduce the subsidiary unknowns

$$
p_{1} \therefore \frac{v /}{d \dot{q}_{1}} .
$$

which will be called the generalized momenta. Since $U$ does not depend on the derivatives, then

$$
\begin{equation*}
D_{1}=\sum A_{14} Q_{L}: A_{1} . \tag{4}
\end{equation*}
$$

Equations (2) may then be replaced by

$$
d \nu_{1}=\frac{d L}{d q_{1}} \quad p_{i}:=\frac{d t}{d q_{1}}
$$

To eliminate $q_{i}$, we use equations (4). We are able to solve these equations relative to $\dot{q}_{k}$ since the determinant formed by the coefficients $A_{i k}$ cannot be equal to zero. Indeed, in the case we are interested in, the halomomic constraints do not involve time. Therefore, in equations (3) $A_{k}=0$ and $A=0$. If the determinants formed by the coefficients $A_{i k}$ were equal to zero, then there would be nonvanishing values of $q_{k}$, for which ${ }^{(1)}$
and consequently $T=0$. Evidently, this cannot take place.
Let us introduce the following quantity

$$
/ / \quad \therefore \because . \quad l .
$$

which is a function of $p_{k}, q_{k}$ and $t$, since $q_{i}$ c.r. be expressed in terms of $p_{i}$ ancording to the above arguments.
(1) To simplif; the formulae, we shall not indicate the limits of summation whenever all of the indices run through the same values $1,2, \ldots . s$, as in the present case. It is also possible not to indicate the summation indices if we introduce the "rule of dummy indices": Summation is always carried out over the indices which are repeated in the sumnand at least twice. For example

Such indices are called dummy since chey disappear after summation.

Varying $p_{k}$ and $q_{k}$ will lead on the one hand to
and on the other hand to

Using the second of equations (5), we obtain

Comparing the two expressions of $\delta \mathrm{H}$, we obtain

$$
\begin{aligned}
& \text { Okilia. } \\
& \text { OFPOOK \& }
\end{aligned}
$$

Taking into consideration the first of equations (5), we finally obtain

$$
\begin{align*}
& d \psi_{k} \quad 1 / 1, d p_{1} \quad i / /  \tag{ti}\\
& \text { vt } \quad \text { dp, db } \quad d \boldsymbol{l} \\
& 1 h^{3} \text { - 1.2......s) }
\end{align*}
$$

If the Lagrangian $L$, and consequently the function $H$, do not explicity depend on time, then equations (5) will have the following first integral
// canst,
winch is nothing else but the kinetic energy integral. In fact, this equation can be rewritten as

$$
\sum_{n}^{1} q_{k}^{d t} \frac{d \eta_{k}}{\partial \|^{2}}-1=\text { con },
$$

or

In our case, $T$ is a homogeneous function of the first order in $q_{k}$, and thus

Noting that $L=T+U$, we write the complete integral obtained above $a s$

$$
r-11 \text { ro:st. }
$$

This is nothing else but the law of conservation of energy.
The canonical equations have several remarkable properties. We
are interested in the most elementary properties of these equations
which can easily be deduced from the following theorem.

## Theorem I

If the general solution of equations (6) is given by the following

## equalities

where $X_{1}, X_{2} \cdots X_{5}$ are constants of integration, then equations
are equivalent to the following equations

$$
\begin{aligned}
& (i=1, \therefore, . \quad ., 2 s)
\end{aligned}
$$

First, we show that equations (8) can be easily inferred fron equations
(6). Indeed, when equations (6) hold, the evident identity
leads to
which are equivalent to equations (8). Conversly, transforming equations (8) into the form (10) and using identity (9), we replace these equations by

$$
\begin{aligned}
& (i=1, \therefore .
\end{aligned}
$$

These equations ran be regarded as a system of 25 linear equations in which expressions

$$
\left(\begin{array}{ccc}
d p_{k} & c^{\prime} H \\
d \prime & c^{\prime} \varphi
\end{array}\right), \quad-\left(\begin{array}{cc}
d q_{k} & d / f \\
d y & d p_{h}
\end{array}\right)
$$

play the role of the unknown quantities. The determinant of this system does not equal zero, since equations (7) are solvable relative to $\gamma_{1}, X_{2} \ldots X_{25}$. Therefore, the expressions just writmen must vanish. Namely this is what equations (6) state.

## Annotation

The canonical equations (6) will not be altered if we interchange the positions of $p_{k}$ and $q_{k}$ and at the same time replace $t$ by $-t$. By means of such interchanges, we can obtain from each property of the canonical systems, a new. property. For example doing such permutations in theorem $I$, we find that the equations are also equivalent to the following equations
20. The canonical transformations

Let us replace in the following canonical equations
the variables $q_{k}$ and $p_{k}$ by new variables, $Q_{k}$ and $p_{k}$, as defined by

$$
\begin{align*}
& 0_{2} \cdot U_{1} 11, q_{1}, \ldots, q_{1}, p_{1} \ldots ., p_{1} \\
& f_{k}=\varphi_{4}\left(1, \varphi_{1}, \ldots \ldots, p_{1} \ldots . . p_{3}\right) \tag{11}
\end{align*}
$$

The canonical system will then be transformed into a new system. The following theorem specifies the condition for the resulting system to also have a canonical form. The corresponding transformations will be called canonical trnasformations.

## Theorem II:

If the relation between the new and old variables is such, that the expremsion

$$
\begin{equation*}
\unlhd_{n_{1}} d d_{k}-\check{\prime_{k}} d l_{G} \quad d!V \tag{12}
\end{equation*}
$$

is a complete differential of some function $W$, then, after the transfromation, equations (6) may be represented in the following way
where

$$
\dot{H} \quad \|+\underset{\partial!}{J i}
$$

Here, it is assumed that functions $H$ and $W$ have been expressed in terms of the new variabies $Q_{k}$ and $P_{k}$.

To prove this theorem, we note that equations (11) and (7) allow us to express the variables $Q_{k}$ and $P_{k}$, and consequently the function $W$, in terms of $\gamma_{1}, \gamma_{2}, \ldots . \gamma_{2 \delta}$ and $t$. In equation (12) we understand that dW is the complete cifferential of the function $W$ only with respect to the variables $Q_{k}$ and $P_{k}$, where we consider that $p_{k}$ and $q_{k}$ are expreased
in terms of $Q_{k}$ and $P_{k}$, although the function $w$ may also depend on time $t$. Therefore, it foilnws from equation (12), that
where

Now, if $W$ is expressed in terms of $\gamma_{1}, \gamma_{2}, \gamma_{3}, \cdots \gamma_{25}$ and $t$, then

$$
\begin{array}{lll}
\therefore & \prime 11 \\
\therefore & \ddots & \therefore \\
! & \because! \\
\therefore!
\end{array}
$$

Differentiating the first of equations (14) with respect to $X_{i}$ and the second with respect to $t$ and subtracting them term by term, we obtain

Applying theorem $I$ to the left-hand side of this equation, and noting that
where

$$
\kappa^{\prime}-H+\frac{d W}{d!} .
$$

On the basis of theorein $I$, it follows from these equations that

$$
d \varphi_{k}=\begin{array}{ccc}
d K^{\prime} & d J_{k} & d K \\
d I_{k}^{\prime} & d l & --d)_{k} .
\end{array}
$$

This is what we wanted to prove.

## Annotation I

If the relation between $q_{k}, P_{k}$ and $Q_{k}, P_{k}$ are such, that

$$
\because y_{k} d p_{k}-\unlhd Q_{k} d J_{k}=d W^{\prime}
$$

then, after the transformation of equations (6), we obtain
where

$$
k^{\prime}=H \ldots \quad \partial W^{\prime}
$$

In order to prove this, it is sufficient to vese the substitution indicated at the end of the previous section.

Annotation II
The conditions of theorem II are often expressed in a slightly
different way. Let us add equation

$$
\therefore \because u^{\prime} \quad \therefore y^{\prime}, \therefore \quad \therefore i^{\prime} U_{1}
$$

term by term to equation (12). This yields

$$
\because \mu A, \quad \because \because u^{\prime \prime}, ~ d,
$$

where

$$
\text { , } \because \because ハ,
$$

or

$$
\begin{array}{llll}
p_{k} & a_{S} \\
d_{i_{4}}
\end{array}, \quad u_{k} \quad \begin{aligned}
& i l_{2}
\end{aligned}
$$

Therefore, starting with the arbitrary function

$$
S\left(:, q_{1}, \quad . \quad \therefore l^{\prime}, . . . P_{1}\right)
$$

and using equations (*), we obtain a canonical transformation.
In conclusion, we give some examples for the applications of
theorem II.

## Example I

Let

Since
then

Hence, the conditions of theorem II are fulfilled here. The canonical system obtained as a result of this change of variables may be written immediately.

## Example II

At the beginning of the pr, ious section, we have seen that the equistions of motion in rectangular coordinates (1) may be easily represented in a canonical form. Theorem II allows us to show that the canonical form of the equations is not violated by the transition from the rectangular coordinates to any curvilinear coordinates. We have already obtained this result in the previous section by another method.

We leave it to the reader to prove this result using theorem II.

## Example III

In practice, it oftenly happens that $\mathbf{P}_{\mathbf{k}}$ are linear functions of $\mathrm{p}_{\mathbf{k}}$ and $Q_{k}$ are linear functions of $q_{k}$. In this case, it is sufficient for the transformation to be canonical, that

$$
\Xi O_{:}^{\prime} O_{k}: \quad \Xi p_{k} \varphi_{k}
$$

Indeed, $d Q_{k}$ will be related to $d_{k}$ by the same relations that related OF POOR QUAL.;.... $Q_{k}$ to $q_{k}$. Consequently, it follows from the previous equality that

$$
\left.\because r, r_{1}^{\prime}\right)_{*} \quad \because p_{1}, \psi_{k} \quad 0 .
$$

21. Jacobi's method for solving canonical systems.

Let us consider the canonical system


Introduaing the new variables $Q_{k}$ and $P_{k}$, related to the old ones by

$$
\because(d) \quad-i \prime \quad(4) \quad 1 .
$$

we obtain, on the basis of theorem II, that
where

System (15) will be resolved once we find a function $W$ for which $K=0$. Actually, equations (17) yield
where $\alpha_{k}$ and $\beta_{k}$ are constants of integration. On the other han we obtain from condition (16) that

$$
\rho_{n}==\begin{array}{lll}
d q_{i} & \because & \ldots \\
a w^{\prime}
\end{array}
$$

where we suppose that $W$ is expressed in terms of $q_{k}$ and $Q_{k}$ by means of equations (11). Replacing in $W$ the quantities $Q_{k}$ by $\chi_{k}$, we obtain the function $W\left(t, q_{1}, q_{2}, \ldots, q_{s}, \ldots, \alpha_{1}, \ldots, \alpha_{s}\right.$ ' that satisfies the following relations

These relations permit us to express $p_{k}$ and $q_{k}$ in terms of $t$ and the $2 s$ arbitrary constants $\alpha_{k}$ and $\beta_{k}$ and thus give the general solution of system (15).

We know the expression of the function $H$ in terms of $p_{k}$ and $q_{k}$. Let this expression be $H\left(t, q_{1}, q_{2}, \ldots, q_{s}, p_{1}, p_{2}, \ldots, p_{s}\right)$. We may write the equation $K=0$ that defines $W$ in the following way


Thus, if we ':onsider $W$ as a function of the $S+1$ independent variables $t$ and $q_{1}, q_{2}, \ldots q_{s}$, then this function will satisfy a first-order partial lifferential equation. Any solution of this euqation which will involve $S+1$ unknown arbitrary constants, will be called a complete integral.

In the present case, the unknown function $W$ enters equation (19) only by its derivatives. The solution of this equation will then involve $S$ arbitrary cusstants, among which there is no additive constants. The complete integral is simply obtained by the introduction of the $(S+1)$-th addetive constant and will have the form

$$
\begin{equation*}
\because: y_{1}, q_{3} \cdots, \square \cdot \because, \cdots, \ldots \tag{20}
\end{equation*}
$$

Thus, we are lead to the following result, which has been represcated by the well-known Jacobi theorem.

## Theorem III

In order to solve the canonical system (15), it is sufficient to find a complete integral of the type (20) for equation (19). The general solution of system (15) is obtained by finding $q_{k}$ and $p_{k}$ from the following equations

$$
\begin{array}{lll}
\because! & \because & 11! \\
\text { i; } & 11 ; &
\end{array}
$$

where $\beta_{1}, \beta_{2}, \ldots ., \beta_{5}$ are new arbitrary constants.
The Jacobi method consists in applying this theorem by integrating the canonical systens. The constants $\alpha_{k}$ and $\beta_{r}$ that appear as a result of integrating the system by this method are called canonical constants or canonical elements..

It is easy to obtain the complete integral of equation (19) when the function $H$ does not depend on $t$. Indeed, substituting in this equation

$$
\begin{equation*}
11 \quad \pi \quad \mathbb{H}^{\circ} . \tag{121}
\end{equation*}
$$

we obtain for the new unknown function $W^{\prime}$ the following equation
where $\alpha$ is an arbitrary parameter. The solution of this equation involves s-1 arbitrary constants $\alpha_{1}, \chi_{2}, \cdots, \alpha_{s-1}$ among which there are no additive constants. Once this solutjon is iound, equation (21) can be used to obtain a solution for equation (19) involving the $s$ constants $\alpha_{p} \alpha_{1}, \alpha_{2}, \ldots . ., \alpha_{-1}$. The general solution of the canorical system
(15) is then given by

$$
\begin{aligned}
& \text { (i. 1. } \because . \quad . \quad .-11 \text {. }
\end{aligned}
$$

where $\beta^{\prime}, \beta_{1}, \beta_{2}, \ldots, \beta_{s-1}$ are new constants.
It is interesting to note that the simplificaion of the integration of system (15), achieved thif way, is a consequence of the existance of the first integial

$$
11 \text { const. }
$$

22. Application for method of variation of arbity:arv constants to the canonical elemento

Let the following canonical sys*em

$$
\begin{array}{cccc}
d l_{1} & d i & d i & d / l \\
d & d p_{i} & d! & d \|
\end{array}
$$

be solved by the Jacobi method. Wa have already seen thai the solution
is obtained from equation (18), i.e.

where $W$ is the complete integral of the equation

Suppose that se rave to solve a new canonical system
where $R$ is a function of $t, q_{1}, \ldots q_{s}, p_{1}, \ldots p_{s}$. To apply the method
of veriation of arbitrary constants, we try tc satisfy these equatiane again by expressions (23) considering the quantities $\alpha_{i}$ and $\beta_{i}$ as functions of time and not as constants. For this purpose, we replace the variables $q_{k}$ and $p_{k}$ in equations (15) by the new varisbles $\alpha_{k}$ and $\beta_{k}$ defined by equations (23). Using equations (24), we get

Applying theorem $I I$, we write the tranaformed equations in the form

$$
\begin{equation*}
\frac{d x_{k}}{d t}=\frac{\partial R}{d_{i} k_{k}}, \quad \frac{d \prime \prime}{d t}=-\frac{\partial R}{\partial x_{k}}, \tag{27}
\end{equation*}
$$

since equation (25) in the present rase yields

$$
K=(H-K) ; \quad \begin{aligned}
& d W \\
& d t
\end{aligned} \cdots
$$

Hence, the application of the method of variation of arbitrary constants to the canonical elements iumediately allows us to write the differential equations for these elaments in a simple form.

## 23. Canonicai elements of elliptic motion

We shall now apply the Jacobi method to the solution of the twobody problem. We denote by $x, y$ and $z$ the coordinates of the planet in the hellocentric courdinate system, and by m its mass. In order to write the equations of motion of the planet,
where

$$
1: \quad: 11: i .1
$$

in a canonical form, it is sufficient to assume that

$$
\begin{array}{ccc:c}
1 & 1 & n & n \\
: & n
\end{array}
$$

Let us introduce the spherical coordinates
and agree to write $k^{2}$ instead of $k^{2}(1+m)$. The Hami_conian will then be written as

In this case, the Hamilton-Jacobi equation reads

This equaticn explicitly involves neither $t$ nor $\theta$. We substitute (see $\$ 21$ ).

$$
W^{\prime} \cdot \cdots x_{1} l: x_{2}^{\prime \prime}: W_{1}
$$

into this equation. We obtain

It is sufficient for us to find a solution of this equation that involves one arbitrary constant. Therefore, we assume
which allows us to write the previous equation as

$$
\binom{d r^{\prime} 1}{-\ddot{d r}}^{z}+z_{j}^{\prime \prime} r: \quad: k^{\prime \prime} r \quad{ }^{\prime} \quad r_{2} z_{1}
$$

so tha: the variables $r$ and $\varphi$ become separated. In other words we assume that

$$
W_{1}-W^{\prime}: \|^{\prime \prime}
$$

where $W^{\prime}$ is a function only of $\varnothing$ and $W^{\prime \prime}$ only of $r$. These functions can be found from the above-equations by means of quadratures. We thus
obtain

We fix the lower limits of integration in order not to introduce unnecessary arbitrary constants. We take $\varphi=0$ as a lower limit for the first integral and the smaller of the two roots of the expression inside the square root as a lower limit for the second integral.

According to theorem III, the general solution of the equations is given by

$$
\begin{array}{llllll}
\therefore!! & \therefore! \\
\therefore & \because & \vdots & \therefore & \because, \\
\vdots & \therefore
\end{array}
$$

In the present case, this solution will have the form

These relations define the coordinates $r, \theta$ and $\varphi$ as functions of $t$ and the six arbitrary constants $\alpha_{1}, \alpha_{2}, \alpha_{3}, \beta_{1}, \beta_{2}$ and $\beta_{3}$. The latter constants are the canonical elements of the elliptic motion.

Let us now find the relation between the canonical elements and the conventional ones. Equation (28) indicates that $r$ should vary in the interval

$$
r_{10} \cdot r \cdot r_{1}
$$

where $r_{0}$ and $r_{1}$ are the roots or the equation

$$
2 x_{1} r^{2}+2 h_{r} r-x_{1}^{2}-: 0 .
$$

In the elliptic motion, the limits of the radius vector are a(1-e) and a(1+e). Therefore,

$$
\left.2 u=r_{1}\right) \vdots r_{1}=--\frac{k^{2}}{x_{1}}, \quad u:(1-c) \quad r_{0} r_{1}=-\frac{x_{1}^{2}}{l_{x_{1}}} .
$$

Hence,

$$
x_{1}=-\frac{k}{2 a} \cdot \quad x_{i}=-k \sqrt{ } a(1 \quad i=) \quad \& v^{\prime} p
$$

On the other hand, denoting by $T$ the time at which the planet passes by the perihelion, we obtain from equation (28)

$$
i_{1}=-7
$$

because $r=r_{0}$ at the moment of passing by the perihelion.
It follows from equation (29) that

$$
\therefore \quad-x_{i}^{2} \sec ^{\circ} \div-11,
$$

or

$$
\cos =\begin{gathered}
\pi \\
a, ~
\end{gathered}
$$

Noting that the lower value of cos $\varphi$ takes place when $\varphi= \pm 1$, where is is the slope of the orbit, we obtain

$$
\text { ! } i_{1}: \operatorname{lin}_{1}
$$

When $p=0$, the planet is at one of the nodes of its orbit. Therefore, we can consider, on the basis of equation (29), that

$$
\because \quad \because
$$

Now, we consider again equation (30). Instead of the latitude, we inc:oduce the argument of the latitude $u$. Since

$$
\text { , in: } \because \quad \therefore \quad i \leq 1 \pi t \text {, }
$$

then, substituting the values obtained for $\alpha_{2}$ and $\alpha_{3}$, we obtain

Hence

ORIGINAL PAGE IS OF POOR QUALITY:

At the moment of passing by the perinelion, $u=\omega$ and $r=r_{0}$, $s 0$ that

$$
p_{3} \quad \omega \quad=\pi-4 .
$$

Thus, we obtain the following system of canonical elements

We shall evaluate the average longitude for the perturbed as well as the unperturbed motion by ( $S(12$ )

$$
\text { A. }=: j_{i}^{!} n d t
$$

On the other hand, we have for the unperturbed motion

$$
i=\pi f \cdot n(t-\eta)=-\cdot \cdot n t \text {. }
$$

Hence,

$$
\beta=\ddot{\square}=
$$

or

To conclude with, we express the elliptic elements in terms of the conventional ones. We obtain

$$
\begin{aligned}
& a=-\frac{k^{2}}{2 a_{1}} ; \quad \because \cdot y_{2}
\end{aligned}
$$

$$
\begin{align*}
& \text { cosi- }{ }^{a_{n}}: \quad \beta_{n}: \beta_{1} \mid \dot{i}_{1} \beta^{:}\left(-2 a_{1}\right)^{j} \tag{3.5}
\end{align*}
$$

## 24. Application of the canonical elements to the derivation of the

## Lagrange Equations

The canonicai elements have the advantiage over the conventional ones in that the equations for their variation during the perturbed motion, are very nimple ( $(22)$. These equations are

$$
d x_{k}-d R \quad d y_{k}=-\begin{align*}
& d k  \tag{k-1,2,3}\\
& d l \\
& d y_{k}
\end{align*} \quad d t=-\quad .
$$

However, the conventional elements are used in the actual calculation of the positions of the planets. Therefore, it is useful to derive equations for the derivatives of the elliptic elements. For this purpose, we differentiate equations (33), and taking into account equations (34), we obtain

$$
\begin{aligned}
& d a=+\begin{array}{cc}
2 a^{2} & d i^{2} \\
R_{2}^{2} & d \beta_{1}
\end{array} \\
& \begin{array}{cc}
d e \\
d t & a\left(l-c^{2}\right) \\
k^{2} e & \partial R \\
d_{i 1}
\end{array}-\begin{array}{cc}
n a \sqrt{ } 1-e^{2} & \partial R \\
k^{2} e & d \beta_{1}
\end{array}
\end{aligned}
$$

$$
\begin{aligned}
& d 2=d R \\
& d t=\cdot d x_{2} \\
& \frac{d \pi}{d f}=-\frac{d R}{d x_{2}}-\frac{\partial R}{d x_{3}}
\end{aligned}
$$

These equations enables us to express the derivatives of the perturbation function with respect to the canonical elements in terms of its derivative
with respect to the elliptic elements. It is easy to see that

$$
\begin{aligned}
& \frac{\partial R}{\partial Y_{1}}=n \frac{\partial R}{O i} \\
& \frac{\partial R}{\partial \beta_{2}}=\frac{\partial R}{\partial S}: \frac{\partial R}{\partial=-1} \frac{\partial R}{\partial z} \\
& \begin{array}{l}
\partial \xi_{2} \\
\partial \xi_{2} \\
=\frac{\partial R}{d \pi}+\frac{\partial R}{d i}, ~
\end{array}
\end{aligned}
$$

$$
\begin{aligned}
& \partial R \text {-custici } \partial R \\
& \text { da: Rfa(l-ex) di }
\end{aligned}
$$

Thus we obtain the following equations, which we have already obtained by another method in chapter III,

$$
\begin{aligned}
& d t=\underset{n a^{2} \backslash 1}{d t}
\end{aligned}
$$

We remind the reader that in these formulae the quantity $\mathrm{ka}^{-3 / 2}$ is denoted by $n$. We note that the last of these formulae can be understond in two ways. On the one hand, if the average longitude of the planet is evaluated by the formula

$$
\begin{equation*}
\lambda=\varepsilon+n\left(t \quad t_{0}\right)=z+k a \quad-\left(t-t_{0}\right) . \tag{36}
\end{equation*}
$$

then $R$ will depend on "a" explicity and through $\boldsymbol{\lambda}$. Hence
where the derivative on the left-hand sides corresponds to the total var: lation of "a", which $\left(\frac{\partial R}{\partial a}\right)$ is the derivative evaluated when is kept constant. On the other hand, if the average longitude is evaluated by

$$
\begin{equation*}
x \cdots \varepsilon+j_{i_{0}}^{j} n d t . \tag{37}
\end{equation*}
$$

where n denotes a function of time defined by

$$
\begin{gather*}
d n  \tag{38}\\
d t=-\begin{array}{l}
3 d R \\
a^{2} d s
\end{array} .
\end{gather*}
$$

then, it is necessary to evaluate the derivatives of $R$ with respect to " a " in the last of formulae (35) fixing the value of $\lambda$, i.e. to take instead of $\frac{\partial R}{\partial a}$. In this way we avoid obtaining a term, that ficreases with increasing time, on the right-hand side of the last of equations (35).
25. The canonical elements of Deiaunay and Poincaré

The canonical elements, defined by equations (31) suffer from a shortcoming, closely related to what we have seen in the end of the previcus section. The element $\alpha_{1}$ enters the perturbation function explicity and through $n$. Thus, we will have on the right-hand side of the following equation

$$
\begin{gathered}
-94- \\
\frac{d i_{1}}{d} \ldots \\
d t
\end{gathered} \begin{gathered}
d x_{1}
\end{gathered}
$$

a term proportional to time. The canonical form of the equations will be lost if we try to bypass this shortcoming by tile method suggested in the previous section, i.e. replacing the element $\epsilon$, given by equation (36), by the element $\mathcal{E}$, given by equation (37).

Delaunay suggested the introduction of the elements

$$
\begin{gathered}
L=k V a \because k^{2}\left(-2 x_{1}\right) \quad: \\
l=-n(t-r)=k a^{\prime} \quad\left(t+p_{1}\right)-k^{2}\left(-2 x_{1}\right)^{\prime \prime}\left(t+\beta_{1}\right) .
\end{gathered}
$$

instead of the elements $\alpha_{1}$ and $\beta_{1}$. Since the difference

$$
i_{1} d x_{1}-l d l l-l d a_{1}
$$

Is a complete differential of the function $W=-t \alpha_{1}$, then on the basis of theorem II (§20) we obtain again a canonical form after the transformation. Introducing the following notation
we obtain

| $\begin{array}{cc} d L & d R^{\prime} \\ d! & d l^{\prime} \end{array}$ | $d /$ $d t$ | $O R^{\prime}$ $d l .$ |
| :---: | :---: | :---: |
| d(j) dR | $d y$ | $d R^{\prime}$ |
|  | $d$ | $d i$ |
| d/I $\\|^{\prime \prime}$ | $d / h$ | $d R^{\prime}$ |
| $d l$ | $d t$ | d// |

where

$$
\therefore \quad R-a_{1} \cdots R+\frac{R^{4}}{2!}
$$

$$
\begin{aligned}
& \text { andill TAGETB }
\end{aligned}
$$

Delaunay elements are expressed in terms of the elliptic elements by

$$
\left.\begin{array}{lll}
I=k V^{\prime} a, & G=k V^{\prime} a\left(1-r^{2}\right), & H=k V a\left(1-r^{2}\right) \cos i  \tag{40}\\
l=n(t-\cdots T), & k-r \cdot \underline{U}, & h=\underline{l}
\end{array}\right\}
$$

Hence, in this system, one of the elements is the mean anomaly $\ell$. The elements $L, G$ and $H$ have the dimension of areal velocity, while the other elements, $\ell, g$ and $h$, are angles.

It is possible to find other homogeneous canonical elements, which have some advantages over Delaunay elements. First of all, following Poincaré, we consider the following system of elements
where $\lambda$ is the mean longitude. This system has the advantage that at small eccentricities and slopes, the elements $\rho_{1}$ and $\rho_{2}$ are also small.

We prove that, in the transition to the elements (41), the differential equations (39) preserve the canonical form. We consider the expression

$$
d d l . \mid \cdot g u\left(j \cdot \mid-h d l l-\lambda d L-w_{1} d l_{1}-w_{2} d d_{2}\right.
$$

which evidently equals

$$
l d L+g \cdot g(i-h d h-(l+g+h) d L+(f ; h)(d L \quad d(i)+h(d)-d l) \quad 0
$$

so that the conditions of theoren II are fulfilled.
In addition to system (41), Poincaré also introduced the following system

# orfanal page is <br> OF POOR QUALITY 

which is canonical in the light of the arguments given in $\$ 20$ (example I). The elements $\xi_{1}$ and $\eta_{1}$ are of order of magnitude of the eccentricity, while $\mathcal{E}_{2}$ and $\eta_{2}$ are of the same order as the slope of the orbit.

The characteristic property of the canonical elements (40), (41) and (42) is the choice of the mean anomaly or the mean longitude as one of the variables, Levi-civita and Hill were able to find other canonical systems, in which one of the elements is the eccentrisity or the true anomaly. De Sitter and Ardoyer developed the general approach for obtaining such systems of elements ${ }^{(1)}$.
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G.W. Hill, Motion of a system of material points under the action of gravitation, Astron Journal, 27 (1913), 171.
W. de Sitter, On canonical elements, Koninklijk Academie van

Wetenschappen te Amsterdam, 16 (1913), 279.
H. Andoyer, Sur 1'anomalie excentrique et l'anomalie varie comme elements canoniques du mouvement elliptique, d'apres M.M. LeviCivita et G.W. Hill, Bull. estr., 30, 1913.

## APPLICATION OF THE CANONICAL VARIABLES IN THE STUDY

## OF PERTURBATIONS

26. The canonical form of the equations of relative motion

The transformation of the equations of motion of a system, for which a force function exists, into a canonical form is carried out in $\delta 19$. In the present section, we consider the equations of relative motion. When the motion of points $\left.m_{1}\left(x_{1}, y_{1}, z_{1}\right), \ldots, m_{n-1} i_{n-1}, y_{n-1}, z_{n-1}\right)$ are related to point $\mathrm{m}_{0}\left(\mathrm{x}_{0}, y_{0}, z_{0}\right)$, taken as the coordinate origin, the equations of motion are ( § 3)

$$
\left.\begin{array}{cc:cc}
d=x_{1} & k^{2}\left(m_{n}\right. & \left.m_{b}\right)_{1} & d k_{1}  \tag{1}\\
d l^{2} & & & \\
r_{1}^{2} & d x_{1} \\
\cdot & \cdot & \cdot & \cdot \\
(i & i, 2, & \cdot & n-
\end{array} \right\rvert\,
$$

where $R_{i}$ denotes the perturbation function that corresponds to point $m_{i}$. To each point $m_{i}$ there corresponds a force function

$$
U_{i}=\frac{k^{2}\left(m_{0}+m_{1}\right)}{r_{1}}+l_{1}
$$

Thus, equations (1) may be transformed into the following forms

| $\begin{gathered} d x_{1} \\ d f \end{gathered}=\begin{aligned} & \partial H_{1} \\ & d x_{1}^{\prime} \end{aligned}$ | $\begin{array}{cc} d x_{1}^{\prime} & d l_{1} \\ d t & d x_{1} \end{array}$ |
| :---: | :---: |
| $d y_{1} \quad d H_{1}$ | $d y_{i}^{\prime} \ldots \quad \because \pi$ |
| dt dy', | $d l-d y_{1}$ |
| $d z_{i}=\partial H_{i}$ | $d z_{1}^{\prime}=-\quad d H_{i}$ |
| $d t^{-} d z^{\prime}$ | $d \mathrm{l}-\mathrm{dz}{ }^{\prime}$ |

where

$$
H_{i}=r-U_{i}
$$

 are different. Poincaré called these equations semicanonical.

It is useful to obtain the equations of relative motion in a canonical form. For this purpose, it is necessary to choose the relative coordinates in a different way. A convenient choice of relative coordinates is given in $\varsigma 4$. In these coordinates, the equations of motion are given by
where

$$
\begin{aligned}
& U=\begin{array}{l}
1 \\
2
\end{array} k_{i} V_{1} m_{1} m_{1} .
\end{aligned}
$$

These equations have the same structure as the equations of the absolute motion in the presence of a force function. They cant be transformed into a canonical form in the usual way ( $£ 19$ ). Assuming that

$$
\begin{gathered}
r \cdots \frac{1}{2} y_{1} n_{1} x_{1} \cdots y_{i}: z_{1} \\
\cdot\|\quad r \cdot\| \\
x_{1}^{\prime}=r_{1} x_{1}, \quad y_{1}^{\prime} \quad \mu_{1} y_{1}, \quad \because_{1} \quad \mu_{1} z_{1},
\end{gathered}
$$

If the masses $m_{1}, m_{2}, \ldots, m_{n-1}$ are small comparef to $m_{0}$, then their mutual gravitation may be neglected. We may then take instead of U the function

$$
U_{0}=k_{i}^{n} \sum_{r_{1}}^{n} m_{0} m_{1}
$$

where

$$
r_{1}^{2}=x_{1}^{2}-y_{1}^{2}+z_{1}^{2},
$$

Consequently, the equations of metion are devided inco $n-1$ separate systems. For example, equations (2) will have the form

Hence, in the first arproximation, we only have to solve a two-body problem to obtain the courdinates $x_{i}, y_{i}$ and $z_{i}$ as functions of time and six constant orbital elements. We shall take as orbital. elements the second set of canonical elements of Poincaré (S 25). We denote the elements that correspond to point $m_{i}$ by

$$
1 /, i_{1,}, \vdots_{1,1}, \vdots_{1,2}, r_{1,1}, r_{1,2}=
$$

In section 25 , it was convenient to avoid double indices. Here, we shall adopt the notation

Thus, we obtain in the first approximation

$$
\begin{aligned}
& x_{1} \quad f\left(l, L_{1}, \lambda_{1}, \xi_{i, 1}, i_{2}, \eta_{2}, \lambda_{12}\right) \\
& y_{1}==\ddot{(t, L,}, \text {. . . . . . . . ) } \\
& z_{1}=\cdots\left(t, L_{1},\right.
\end{aligned}
$$

as well as similar expressions for $x_{i}^{\prime}, y_{i}^{\prime}$ and $z_{i}^{\prime}$.
We apply the method of variation of arbitrary constants to obtain the general solution of equations (3). We replace the variables $x_{i} \quad y_{i}, z_{i}, x_{i}^{\prime}, y_{i}^{\prime}$ and $z_{i}^{\prime}$ by the variables $L_{j}, \lambda_{j}, \xi_{j}$ and $\gamma_{j}$ using the equations just written. We choose as a perturbation function the quantity

$$
R=-11-1 i_{0},
$$

since, in this case,

$$
H \cdot T-U_{0}-\cdots
$$

We obtain the transformed equations in the following form

$$
\begin{aligned}
& \text { (i=1, 2, . . ., } n-1 ; \quad j=1,2 . . . .2 n-2 \text { ), }
\end{aligned}
$$

where ( $\delta^{\circ} 25$ )
since, in the present case, the quantity $k^{2}$ for point $m_{1}$ should be replaced by

It is useful to note that the replacement of $R$ by $R^{\prime}$ is connected with the choice of the mean longitude $\lambda_{\text {ias }}$ one of the variables.

We finally apply equations (4) to the motion of three bodies having masses $m_{0}, m_{1}$ and $m_{2}$. In this case,
where

Fquation (16) of Chapter I2yields

$$
\begin{aligned}
& y_{i 1}^{\prime \prime} \quad x_{1}^{2}: y_{1}^{\prime \prime}: \because_{1} \quad r_{1}
\end{aligned}
$$

In this way, we obtain

$$
R \cdot k \cdot m_{0} m_{:}\left(\begin{array}{cc}
: & 1  \tag{i}\\
د_{20} & - \\
r_{2}
\end{array}\right): k: m_{1} m_{:} \begin{gathered}
1 \\
د_{1}:
\end{gathered}
$$

## 27. The integrals of area

In $\$ 4$, it was claimed that the form of equations (2) was similar to the form of the equations of the absolute motion of $n$ bodies, and for this reason, both sets of equation should have similar integrals of area, Thus, in the three-body problem $(i=1,2)$, equations ( 2 ) will have the following integrals, which correspond to integrals (6) of $\$ 1$,

$$
\begin{aligned}
& \mu_{1}\left(y_{1} \dot{n}_{1}-z_{1} y_{1}\right): \mu_{2}\left(y_{2}=\therefore-y_{2}\right)=C_{1} \\
& x_{1}\left(=\dot{x}_{1}-x_{1} \dot{x}_{1}\right): n_{2}\left(\dot{z}_{2}-x_{2} \dot{x}_{2}\right)=\left(i_{0}\right. \\
& \because\left(x_{1} \dot{y}_{1}-y_{1} \dot{x}_{1}\right)+1_{2}\left(x_{2} y_{2}-y_{2} \dot{x}_{2}\right) \quad C_{1} \text {, }
\end{aligned}
$$

where

$$
\begin{array}{cc}
m_{1} m_{10} \\
\vdots & m_{1} \cdot 1-m_{1}
\end{array} \quad \because=-\begin{gathered}
m_{1}\left(m_{4} \cdot j-m_{1}\right) \\
m_{n}+m_{1} \vdots m_{:}
\end{gathered}
$$

We shall use the elliptic coordinates as an intermediate step to express these integrals in terms of the canonical elements. In the case of unperturbed motion, the areal velocities are expressed in terns of the elliptic elements by

$$
\begin{array}{ll}
x z-z \dot{y} & \because(i \sin \sin ! \\
z \dot{x}-x \dot{z} & r \sin 1 \operatorname{cov} \because \\
x y-v \dot{x} & \because G \cos 1 .
\end{array}
$$

where

$$
\text { (; } \cdot k y^{\prime} d(: \quad \rho:)
$$

These relations hold also for the perturbed motion because we are using osculating elements. The relations between the elliptic and canonical elements ( $£ 25$ ) yield
and thus

$$
(i: a+11 \quad \mid \because \because 1 / \quad \therefore 1) \quad \therefore
$$

Consequently

We thus write the integrals of area in terms of the canonical elements in the following way

$$
\begin{aligned}
& i_{1}\left(L_{1} \cdots i_{1,1}-\beta_{1,!}\right)!\mu_{i}\left(I_{2} \quad \rho_{1} \cdots \mu_{2}\right) \quad=C_{3} .
\end{aligned}
$$

If we take the invariable plane as the $x y$ plane, then $C_{1}=C_{2}=0$. In this case, the first two equalities yield

$$
\begin{array}{llll}
r_{1}^{1}: & \ddots & r_{1} ., \\
r_{1,2} & \vdots & \vdots
\end{array}
$$

or
( $\leqslant 25$ )

$$
t_{1} \mathscr{S}_{1}-t_{1} \mathscr{S}_{2}
$$

In other worde, the line of intersection of two relative osculating orbits is parallel to the invariable plane. This property permits us to introduce one node instead of two. This was the reason why Jacobi salled this property the elimination of nodes. It was shown in section 2 that the elimination of nodes is a particular case of a more general property of dynamic systems.
28. Expression of rectangular coordinates in terms of canonical elements

Before integrating equations (4), we have to express the perturbation function $R$ in terms of the canonical elements. The perturbation function is easily expressed in terms of the rectangular coordinates. Thus we start by expressing rectangular coordinates $X_{i}, y_{i}$ and $z_{i}$ in terms of the elements $L_{i}, \lambda_{i}$, ... of this point. We recall the formulae that connect the canonical elements of Poincaré with the elliptic elements (§25). Introducing the angle of eccentricity $\varphi$, we write these formulae in the following manner

$$
\begin{aligned}
& \text { 1. ' } 1 \text { l' } 1 \text { : } \quad \because \quad \text { : } \because!=1 \\
& \therefore 1 / .:= \\
& \therefore \quad \because / \mathrm{CW} \cdot \boldsymbol{\square} \quad 1 .
\end{aligned}
$$

The well-known relations between the rectangular and alliptic coordinates lead to
$x \quad r \cos u \cos u-r \sin u \sin \because \cos t$
$y: r \cos u \sin \because, r \sin u \cos \because \cos i$
$z \cdot r \sin u \sin i$,
where

$$
u-v ;-\cdots
$$

is the argument of the latitude. These formulae can be represented in t:he following way
if

$$
x \cdots r \cos (0 \cdots m), \quad y \quad r, m(i n-d)
$$

and $M=\lambda-\pi$ denotes the average anomaly.
The result which we are trying to obtain can be expressed in the form of the following theorem:

Theorem
Each of the rectangular coordinates can be expanded in a series of the type
where $\alpha_{1}, \alpha_{1}, \beta_{1}, \beta_{2} \quad$ and $k$ are positive integers or zeros, $H$ is a constant and A depends on $L$.

We first show that the coordinates can be expanded in series of positive integral powers of $\xi_{1}, \sum_{2}, \eta_{1}$ and $\eta_{2}$, such that the expansion coefficients depend on $L$ and are periodic functions of $\lambda$. Indeed, the expressions of $X$ and $Y$ in equations (7) consist of sin and $\cos \lambda$ multiplied by the quantities

$$
\begin{equation*}
\cos ^{2} 2^{1} \quad \sin ^{2} \because \cos 29 . \quad \text { an } i i^{\cos } \sin 4 \tag{'}
\end{equation*}
$$

However, on the basis of expressions ( $6^{\prime}$ ),

Consequently,

We see that $\sin ^{2} \frac{i}{2}$ and $\cos ^{2} \frac{1}{2}$ car be expanded in series of the required type. Moreover, we have

On the other hand,

$$
\text { !y - } \quad \begin{aligned}
& ! \\
& \vdots
\end{aligned}
$$

Hence

Comparing these expressions with the previous nes, we see that the five quantities (9) are expandable in series of positive integral powers of
$\xi_{1}, \zeta_{2}, \eta_{1}$ and $\eta_{2}$. Let us now consider the quantities $X$ and $Y$.
Since

$$
\text { riosu } a(c o s t-i), \quad r \sin v \quad a y i-a \sin t ;
$$

then

Each of the quantities
can be expanded in positive integral powers of $e \sin M$ and $e \cos M$. This is evident as far as the expansion of the first two quantities in powers of $e^{\prime}$ is concerned. On the other hand, the Kepler equation

$$
11 \quad 1 \quad \therefore!!
$$

leads to

Assuming that

$$
u^{\prime} \quad i \quad i H, \quad z_{1} \cdot \sin i t, \quad \therefore \quad r i o s i
$$

this equation may be rewritten as

$$
f\left(1 \prime, x_{1}, \therefore\right) \quad 0
$$

where the left-hand side is a holomorphic function of $w, z$, and $z_{2}$ at the point $w=z_{1}=z_{2}=0$. According to a well-known theorem on implicit functions, if

$$
\frac{d f}{d u}+0 \quad 2.24 \quad w \quad r_{1}=z_{3} \cdots 0
$$

then thisequation has one and only one solution $w=\varphi\left(z_{1}, z_{2}\right)$ being holomorphic in the vecinity of point $z_{1}=0, z_{2}=0$. This property is satisfied in the present case since

$$
\binom{\partial j}{\partial w}_{0}=1 .
$$

Therefore, we obtain from equation (10) the quantity $w=E-M$ in the form of a series expansion in positive integral powers of $z_{1}$ and $z_{2}$. We conclude that $\sin (E-M)$ and $\cos (E-M)$ are also expandable in similar series.

Finally, the equalities

$$
\begin{aligned}
& C^{2} \cos (E-f-M)=C^{2} \cos 2 M \cos (E-N) \quad C^{2} \sin 2 M \sin (E-N)
\end{aligned}
$$

$$
\begin{aligned}
& e^{2} \text { cos } 2.17 . .(e \cos .17)^{-}-\left(e^{\circ} \sin \right. \text { M) }
\end{aligned}
$$

Indicate that the expression

$$
0_{i n}^{3}(t \mid, 1 f)
$$

has the required property. In this manner, the possibility of expansion of $X$ and $Y$ in positive integral powers of $e \sin M$ and $e \cos M$ is proved.

Since $M=\lambda_{-} 7$, then.

$$
\begin{aligned}
& e \sin A f e \cos \pi \sin A-e \sin \pi \cos \lambda \\
& e \cos A 1 e \cos \pi \cos A+e \sin \pi \operatorname{cin} \lambda
\end{aligned}
$$

Thus, $X$ and $Y$ can be expanded in positive integral powers of $e \cos \pi$ and $e \sin \pi$, and the expansion coefficients will be functions of time.

The theorem will be completely proved when we prove the possibility of expanding e $\cos \pi$ and $e \sin \pi$ in positive integral powers of $\varepsilon_{1}$ and $7_{1}$. Firat of all, if follows from equations ( $6^{\circ}$ ) that

$$
1!r_{1}^{2}-1 l . n m: \because \quad \because l .(1-\cdots, r) \quad \because l .\left(1-\sqrt{1} \quad e^{2}\right)
$$

From this equation, it follows that

Consequent1y,

$$
\left.\therefore \quad \begin{gathered}
l i_{i} \\
V / r_{i 1}^{\prime \prime}
\end{gathered} \right\rvert\, 1 \quad 1 L^{1}\left(i_{i}^{\prime} \mid r_{i}\right) \|^{\prime}
$$

On the other hand, the same formulae ( $6^{\prime}$ ) yields

$$
\stackrel{c \cos \pi}{\vdots_{1}}:=\frac{e \sin \pi}{-r_{12}}=\frac{e}{1^{\prime} ;{ }_{1}^{2}+r_{1}^{2}}
$$

Hence

$$
\begin{aligned}
& \left.e \sin \pi=\frac{-\eta_{1}}{V / l} \quad 11\left(l_{1}^{2}+1+\eta_{1}^{2}\right)\right)^{\prime},
\end{aligned}
$$

We have proved that each of the coordinates $x, y$ and $z$ can be represented by a series of the type
where $B$ is a function of $L$ and $\psi(\lambda)$ a periodic function of $\lambda$, the period of which equals $2 \pi$. The function $\psi(\lambda)$ can be expanded in a Fuurier series

$$
F(i)=\Sigma\left(C_{k} \cos k A \cdot+-D_{k} \sin n_{n}^{\prime} \lambda\right)
$$

Since

$$
\sin k \lambda=\cos \left(k i+F-\frac{\pi}{2}\right),
$$

then, this proves our theorem.
The series (8) obtained above converge for small values of $\varepsilon_{1}, \xi_{2}$, $\eta_{1}$ and $\eta_{2}$, A more exact determination of the region of convergence of these series will not be given here.

## 29. Expression of the perturbation function in terms of the canonical

## variables

The perturbation function $R$, which we are going to study now, is equal to the difference $U-U_{0}$ (see section 26). Consequently
where

$$
\begin{array}{ll}
-109- & \text { ORIGINAL PAGE IS } \\
r_{1} x_{i}^{:} y_{i}^{\prime} z_{i} & \text { OF POOR QUALITY }
\end{array}
$$

If netther the quantity $r_{i}$ nor $\Delta_{i f}$ equals zero at any time, when

$$
\vdots=:=0, \quad i_{1}=11, \quad \cup \quad 12, \quad, \quad \therefore
$$

then both of them, and consequently the perturbation function will be holomurphic functions in the vicinity of point $\xi_{j}=0, r_{j}=0$. Hence, the function $R$ can be expanded in positive powers of $\mathcal{E}_{j}$ and $\bigcap_{j}$ in the vecinity of this point. The expansion coefficients willobe finite and continuous for all values of $t$. They will be periodic functions of $\lambda_{i}$ having a period of $2 \pi$. Hence, these coefficients may be expanded in multiple Fourler series of the type.
where the summation is carried over the indices $k_{1}$ run over all the positive as well as the negative integral values. This result can be stated in the follcwing way:

## Theorem I

If points $m_{o}, m_{1}, \ldots, m_{n-1}$ move in such a was, that their mutual separations $\Delta_{k 1}$ and radius vectors $r_{i}$ are never equal to zero, then the corresponding perturbation functions $R$ can be expanded in a series of the type
where $H$ is a constant, the coefficients $A$ depend only on $L_{i}$ and $m$ is a product of positive powers of $\mathcal{E}$, and $n_{1}$, i.e. $m=\xi_{1}^{\alpha_{1}} \xi_{2}^{\alpha_{2}} \ldots \eta_{1}^{\beta_{1}} \eta_{2}^{\beta_{2}} \ldots$ The summation $\pm$ swer the indices $\alpha_{1}, \alpha_{2}, \ldots, \beta_{1}, \beta_{2} \cdots$ and also over $k_{i}$ which all run over all possible positive and negative

## ORIGINAL PAGE 2

 OF POOR QUALITYintegral values. The series (12) converges for sufficiently small values of $\xi_{1}$ and $\eta_{1}$.

We now investigate how the function $R$ can be expressed in terms of the variables $\rho_{j}$ and $\omega_{j}$, related to $\varepsilon_{j}$ and $\eta_{j}$ by

$$
\begin{equation*}
i_{1}=V^{\prime} 2 \theta_{1} \cos \omega_{1}, \quad \eta_{1}=\sqrt{2} \theta_{1}, \sin \omega_{1} . \tag{1;3}
\end{equation*}
$$

We prove the following theorem:

## Theorem II

Under the same conditions, as those which hold in theorem $I$, the function $R$ can be expanded in the series
which converges for sufficiently small values of $S i$. Each of the quantities $2 q_{i}$ takes the values $0,1,2$, $\ldots$, while the indices $k_{i}$ and $p_{i}$ run over all the integral values from $-\infty$ to $+\infty$. For each term of this expansion, the following condition holds

$$
\begin{equation*}
2 q, \cdot 1 p, i ; \quad 2 \varphi, \quad p_{1}(\bmod 2) \tag{1}
\end{equation*}
$$

where the coefficients $A$ depend only on $L_{i}$ while the quantities $H$ are constants.

In order to prove this theorem, we transform each term of the series (12) by introducing the new variables $\rho$ and $\omega$ instead of the variables $\xi$ and $\eta$ by means of equations (13). We start by those terms for which the product $m=1$. Evidently, these terms will not be changed, their form will remain to be

$$
d \cos (こ \nleftarrow, 1,111)
$$

Accordingly, they satisfy the condtions of the theorem.
We now consider the following expansion:
mitre $q$ and $p$ satisfy the conditions requisit by the theorem. We show tras the multiplication of these terms $\tilde{H} \cdot *$ and $\mathcal{F}_{\mathrm{h}}$ produces a Dun erms havirg the same type. For in.: urpose, we consider expressicn

It is better tonsider the gen ${ }^{\text {a }}$. : ession
which is evidently equal to

Let us consider the particular case in which $h=3$. After the transformation, the coefficients $2 q_{1}, 2 q_{2}, 2 q_{4}, \ldots$ and $2 p_{1}, p_{2}, p_{4}, \ldots$ will not be altered and tinerefore satisfy condition (15). The coefficient $2 q_{3}$ will be replaced by $2 q_{3}+1$ while the coefficient $p_{3}$ will be replaced by $p_{3}-1$. Since, by condition,

$$
\because q . \quad D_{1}(\bmod \because), \quad \because 43 \quad \therefore p_{3},
$$

then

$$
\begin{aligned}
& \because n_{1}: 1 \quad p, \quad 1 \text { mon!! } \\
& 2 u_{1}+1-1 p,+1, \quad \because u_{1} \mid \quad p_{3}-1
\end{aligned}
$$

In this way, starting with a term of type (16), and progressing successively to other terms of expansion (12) by means of multiplications by $\mathcal{\xi}^{\boldsymbol{F}}$ and $\mathcal{T}$, we will only obtain terms satisfying condition (15).

## 30. Poincare's theorem on the rank

In the previous sections, we have studied the forms of the expansions of the perturbation function $R$. Now, we consider again the integration of equations (4), which may be rewriten in terms of the new coorinates as follows

$$
\begin{aligned}
& \text { (i=1,2, . . } 1-1 ; \quad, \quad 1, \cdots 2, \ldots, n-2)
\end{aligned}
$$

where

$$
R^{\prime} R_{0}: R
$$

を $\eta_{1}=$
while the function $R$ is defined by equation (5). On the be is of theorem I of section 29 , the perturbation function can be expanded in a series of the type
where $m$ is a product of positive powers of $\varepsilon_{1}$ and $\varkappa_{i}$.
We remind the reader that the coefficients $A$ depend only on elements $L_{i}$ and can have as a multiplying factor only one of the masses $m_{i}$. This can easily be seen from equations (11). Accordingly, in the first approximation in which all $m_{1}=0$, we may write

$$
1_{1}-1 \therefore, \quad \vdots \quad \xi_{1}^{0}, \quad \eta_{1}=\eta_{1}^{0}, \lambda_{1} \quad i_{1} 1 \quad \lambda_{1}^{0}
$$

where the upper index zero denotes constant values,
and

$$
i_{1} \quad!: \quad \quad \because, \quad y_{i} \cdots m_{v} \begin{gathered}
m_{1} \cdot m_{1}: m_{0} \\
m_{0}
\end{gathered} m_{1}
$$

We substitute the values (18) into the right-hand side of equations (17).
We obtain expressions of the following type:
where

$$
\Sigma^{\prime} n_{1}
$$

In the second approximation, we obtain
 OF POOR QUALTE:
in which we denote by $\left.\delta_{1} L_{i}, \ldots, \delta_{1}\right\}_{1}$ the sums of the type

$$
r_{0} t-\underline{v}_{y}^{\beta} \sin \left(\mathrm{v} \mid:-H^{\prime}\right)
$$

The secular elements $B_{o} t$ are obtained as a result of the integration of the three terms of series (19), in which $v=0$.

It is easy to see that $\delta_{1} L_{i}$ does not involve any secular elements. This result is equivalent to the Laplace-Legendre theorem, given in section 17, and which states that the semimajor axes of the orbits are invariable.

Substituting expressions (20) into the right-hand side of equations (17), we obtain the third approximation, and so on. In this way, we obtain, after an arbitrary number of approximations,
where each of the quantities $\delta L_{i}, \ldots, \delta \eta_{i}$ is represented by a series having the fism

$$
\because A t^{r} \cdot W_{i} \cos \left(\cdot l \vdots \mid t^{\prime}\right) \text {, }
$$

in which $m$ is a product of nonegative powers of $\xi_{1}$, and $\mathcal{\gamma}_{2}^{0}$, while the coefficient $A$ depends only on $L_{i}^{0}$ and has a multiplying factor of $m_{1}^{m}$ $m_{2}, \ldots$ where $m^{\prime}, m^{\prime \prime}, \ldots$ are integers satisfying the relation

$$
m^{\prime} \cdot 11, \quad m^{\prime \prime} \quad 11, . . . . m^{\prime} \cdot m^{\prime \prime} \text { : . . . } 1 .
$$

We remind the reader that the sum $m^{\prime}+m^{\prime \prime}+\ldots$ is called the order of the corresponding series while the expression $m^{\prime}+m^{\prime \prime}+\ldots-p$ is its rank. Poincare proved the following theorem:

Theorem
If the mean motion of $n$ planets is such that the following relation

$$
\nu=\Sigma_{i}^{\prime} \because_{\iota}=
$$

where $k_{i}$ are integers, then
(1) The rank of each term of the expansions of $\zeta_{L_{i}}, \delta \lambda_{i}, \delta \xi_{j}$ and $\delta \eta_{i}$ is more than or equal to zero.
(2) The rank of each mixed term equals at least unity, and
(3) The expansion of $\delta L_{i}$ does not involve zero-rank terms.

We have just seen that this theorem is valid for first-order
terms. Indeed, the quantifis $\delta_{1} L_{i}, \delta_{1} \lambda_{i}, \delta_{1} \varepsilon_{j}$ and $\delta_{1} \eta_{j}$ do not involve mixed terms. They can involve secular terms only of the type At, i.e. having a rank of more than or equal to zero. Finally, the expression $\delta_{1} l_{i}$ involves no secular terms. We shall now prove that once the theorem is valid for all terms having orde: $\leqslant \mathrm{m}$, it will also be valid for the $(m+1)$-order terms. We divide our proof into three parts. First, we deduce the expressions required for the calculation of the ( $m+1$ ) order term. We substitute equations (21), in which $\delta L_{1}, \ldots, \delta \eta_{1}$ are understood as the aggregate ofterms having orders $\leqslant m$, into the righthand side of equations (17). Beforehand, we write these equations in the following way

Integrating the three equations that do not involve $R_{o}$, we obtain

The quantity $R$ is of the first order relative to the masses $m_{1}, m_{2}, \ldots$ The substitution of expressions (21) into the right-hand side of these
equations, the error of which is of the $(m+1)$ order, will thus yield right-hand sides having an error of the $(m+2)$ - order. This allows us to evaluate all of the $(m+1)$-order terms in $\delta L_{i}, \delta \varepsilon_{j} j$ and $\left.\delta\right\rceil_{j}$ In order to evaluate the $(m+1)$ - order terms in the expressions of the : mean longitudes $\lambda_{i}$, we first consider the expansion of $\partial_{R} / \partial_{i} i$, Since

$$
\begin{aligned}
& R_{6}=K_{0}\left(L_{1}^{\prime \prime}-1-i L_{1}, \quad L_{z}^{u}+i L_{2}, . . .1=\right.
\end{aligned}
$$

where
and by $\Phi$ is denoted the aggregate of the third- and higher-order terms reiative to $\delta L_{i}$, then

Noting that
we obtain, after integrating,

We want to be sure whether the substitution of expression (21) into the right-hand side of this equation is done within an error of the $(m+2)$ order relative to the masses. This is evidently correct as far as the first and the last terms are concerned. The reason is that the perturbation function $R$ is a first-order quantity. Only the second term is required to be considered. First of all, we note that the partial derivative $\partial \Phi / \partial L_{i} \quad$ is a sum of terms having at least the second order relative to the perturbations $\delta_{L_{k}}$. Each of these perturbations
is of the first order relative to the masses. We replace the quantities $\delta L_{k}$ by their approximate values $\delta^{\prime} L_{k}$ which involve errors of the order $m+1$. Identities of the type

$$
\begin{aligned}
& A B \cdots A^{\prime} \cdot A(B-N) \cdots\left(A-A^{\prime}\right) .
\end{aligned}
$$

show that this replacement produces the partial derivative $\partial \Phi / \partial L$ to within an error of the $(m+2)$-order. Thence, if we replace the perturbation involved in the right-hand side of equations (22) and (23) by their approximate values, which are correct to within m-order terms, we obtain all of the $(m+1)$-order terms in the left-hand side.

We now prove the theorem as far as it concerns quantities $\delta L_{i}$, $\delta \tau_{j}$ and $\delta \eta_{j}$. We first note that the multiplication of two terms of positive ranks yields a sum of positive-rank terms. Similarly. the multiplication of terms of negative rank yields terms of negative rank. Hence, when we substitute expressions (21), which consist of terms of negative rank, into the right-hand side of equations (22), we obicain a sum of terms of non-negative rank in the expression of the integr . Moreover, since each term of the perturbation function $R$ is multiplied by $m_{1}$ or $m_{2}$, or .... the ranks of all of these teims will be greater than or equal to unity. In integrating the secular terms, their ranks are decreaser by a unity as the following formula indicates

$$
\int_{0}^{0} d t^{\prime} d t=\frac{A r^{n+1}}{f^{\prime}+1}
$$

Hence, expressions (22) consist of terms whose rank is not less than zero. Iue above-mentioned reduction of the rank takes place only for purely secular terms. Hence, equations (22) cannot involve mixed terms having zero-ranks.

We still have to prove that the expr :ssion

$$
\text { ?1. } \int_{0}^{t} N_{1} d t
$$

involves no terms of the zero rank. We substitute expression (21) into that of the partial derivative $\partial_{R} / \partial \lambda_{i}$. Expanding this partial derivative in a series, we obtain
where $D_{0}$ denotes those partical derivatives of $\partial k / \partial \lambda_{i}$ with respect to the elements, in which the values of the elements are replaced by the following initial values:
and $R$ denotes a product of non-negative powers of $\delta_{i}, \delta \lambda_{i}, \delta \varepsilon_{j}$ and $\delta_{j} \eta_{j}$ On account of theorem $I$ in section 29, each of the partial derivatives $\partial R / \partial \lambda_{i}$ may be expanded in a series of the type

$$
\Sigma d \cdots i \cos \left(\Xi h_{1} \lambda_{2}+H\right.
$$

where all of the indices $k_{i}$ may evidently be assumed not to equal zero. Replacing the elements by their above-mentioned initial values, we obtain
where

$$
v \quad \because k_{t}^{n} n_{r}
$$

The quantity $\mathcal{V}$ cannot be equal to zero since neither of the indices $k_{i}$ equal zero. The partial derivatives $D_{0}$ will thus consist entirely of periodic terms. The rank of each of these terms is $\geqslant 1$, because the function $R$ is a first-order quantity relative to the masses.

Now, considering product $R$, it is easy to see that each zero-rank term of this expression can only be a result of the multiplication of zero-order terms relative to $\delta L_{i}, \delta \lambda_{i}, \delta \xi_{j}$ and $\delta \eta_{j}$. We are assuming here that these latter quantities can only have secular terms of
zero rank. Hence, each zero-rank term in $R$ must be secular. When we multiply quantity $D_{0}$ by product $R$, we will obtain 'only terms the ranks of which are mor? than or equal to unity. The first-rank terms will be either periodic or mixed secular. In both cases, the irt egration cannot reduce the rank as can be seen from the following well-known formula:

The theorem is already proved as far as it concerns $\delta L_{i}, \delta \varepsilon_{i j}$ and $\delta \%_{j}$. It remains for us to consider the expression, given by equation (23). This expression consists of three terms. Everything that is applicable to equation (22) holds true for the last of these three terms. This term can thus lead to terms having neither negative nor a zero rank. We now consider the second term $\therefore$
$i$
$i$
We have shown that the derivative $\partial \mathscr{J} / \partial L_{i}$ consists of terms at least of the second order relative to $\delta \mathrm{L}$. Since the quantity $\delta \mathrm{L}$ is equal to a sum of terms all of which have rank $\geqslant 1$, then the rank of each term of the partial derivative $\delta \Phi / \partial L_{i}$; will be $\geqslant 2$. The integration reduces the rank of each term by a unity, yet the ranks will still be $\geqslant 1$ as the theorem implies.

It now remains to consider the first term

$$
\mathbf{V}^{\prime} \because \int_{0}^{\prime} \int_{0}^{\prime} i^{\prime \prime} \quad \therefore
$$

It follows from the above arguments that there are no zero-order terms in the expression of the partial derivative $\partial R / \partial \lambda_{k}$ First orde terms are either periodic or mixed. In both cases the rank does not change after the double integration. The rmaining terms will lead either to periodic or co mixed terms having ranks $\geqslant 2$, or to secular terms having zedandepar.


Hence, among the ( $m-1$ ) order terms of the expression of the quantity there will be no terms having negative ranks and no mixed terms of zero rank.

The theorem is thus completely proven.

## 31. Poisson's theorem

Poincaré's theorem, proved in the previous section, is a generalization of the well-known Laplace-Lagrange theorem on the absence of secular perturbations in semimajor axes. Poisson's theorem, mentioned in section 17, gives a generalization of the Laplace-Lagrange theorem in ancther direction.

The semimajor axis $a_{i}$ of an orbit is related (section 25) to the element $L_{i}$ by

$$
H_{1} \ldots H_{1}^{\prime} \quad \prime!
$$

where $M_{i}$ is a factor, which depends en masses and is slightly different from $a$ unity when the mass of the sun is chosen as the mass unit. Denoting by $\delta_{m} a_{i}$ and $\delta_{m} L_{i}$ the m-order perturbation of the elements $a_{i}$ and $L_{i}$ and by $a_{i}^{c}$ and $L_{i}^{o}$ the osculating elements for the moment $t=0$, we obtain
where

We have already proved that $\delta_{1} L_{i}$ consists only of periodic terms. Accordingly, secular terms will be present in $\mathcal{E}_{2}{ }_{i}$ only if they are present in $\delta_{2} L_{i}$. Hence, Poisson's theorem may be reformulated as follows:

## Poisson's theorem:

If the mean motions $n_{1}$ of a planet are such that the relation

$$
\begin{equation*}
\because \because \leq: \because \tag{11}
\end{equation*}
$$

does not hold for any integral value of $k_{1}$, then, no secular terms can be present among the terms of $\delta L_{i}$ that have a second order relative to the masses.

In order to prove this theorem, we consider equations (22), which yield

$$
\begin{array}{cc}
d \\
\left.d l^{2} L_{1}\right) & d R \\
d!
\end{array}
$$

## Putting

$$
l_{4}=l_{i}^{\prime \prime}-i_{1} L_{i}, \quad i_{1}=n_{1} t: i_{1}^{\prime \prime}: i_{1} ; \quad \vdots_{1} \cdot \vdots,-i_{1} \bar{i}_{1}, \quad r_{1}=r_{1,}^{\prime \prime}+i_{1} r_{1,},
$$

in the right-hand side of this equation, we obtain

According to equations (22), the first order perturbations $\mathcal{E}_{1} L_{k}$, $\delta_{1} \Sigma_{j}$ and $\sigma_{1} \eta_{j}$ are equal to

We divide $\delta_{1} \lambda_{k}$ according to equation (23) into two parts, such that

$$
i_{1} 1_{1} \quad \ddot{i_{1}} \quad i \quad i_{1}^{\prime}
$$

where

The last term in equation (23) produces perturbations of orders not lower than the second and therefore, can be neglected in the present discussions. In this way,

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It was shown in the preceding section that each of the second derivatives involved in the previous expression can be expanded in a series of the type

$$
\sum A>\cos (:+N)
$$

where $\nu=\sum k_{i} n_{i}$ cannot be equal to zero otherwise the corresponding term vanishes after differentiating with respect to $\lambda_{i}$. Secular terms can appear in the expression of the quantity $\delta_{2} L_{i}$ only when constant terms are present in the expansions (25). It is easy to see that the first of the sums involved in equation (25) cannot produce a term. Indeed, each term

$$
\eta-A \cdots i \cdot c o \cdots k \lambda_{i}: \pi
$$

of series (12) in which $R$ is expanded, gives the following part of this sum

These expressions equal a sum of periodic terms, none of which are constant. Similarly, we can show that no term in the series expansion of $R$ can give a constant term in the second sum of equation (25). The third term can contribute only with periodic and secular terms since

$$
\begin{aligned}
& \text { かいい! ! は! }
\end{aligned}
$$

and

$$
\begin{aligned}
& \because!\cos \left(\underline{\prime} k+ \pm 2 \because^{\prime}:!!\right.
\end{aligned}
$$

This completely proves Poisson＇s theorem，we note that the third sum in equation（25）yields a mixed term of the second rank in the expression of the second rank in $\delta_{2} L_{1}$ ．

Poincaré generalized Poisson＇s theorem by proving that the expression of $\delta \mathrm{L}$ cannot involve secular terms not only of the zero rank（section 29）but also of the first rank ${ }^{(1)}$ ．
32．Poincaré＇s theorem on the class
In section 30 ，we began to study the expressions of the perturbations $\sigma L_{i}, \delta \lambda_{i}, \delta \xi_{j}$ and $\delta \eta_{j}$ obtained as a result of applying the method of successive approximations to equation（17）．Each of these quantities is obtained in the form of a series of terms having the following structure：
．．．．：© ！！！！！！
where

$$
v=\Sigma R n, \quad \text { it } 1, \therefore
$$

are factors introduced by the integration．If the coefficients $A_{0}$ are of the m－order relative to the perturbing masses，then（section 15 ）．
（1）M．Poincare，Lecon de Mechanique Celeste，t．1，Paris 1905， 294.

$$
\begin{array}{cc}
-1.23- & \text { ORIGINAL PAGN } 2 \mathrm{I} \\
\because-1, p-\frac{1}{2} & \text { OF POOR QUALITY }
\end{array}
$$

is called the class of the term under considerations relative to the corresponding divisor $\nu_{i}$.

## Theorem

If the mean motions $n_{i}$ is such that, for the arbitrary integer, $k_{i}$ the following relation holds

$$
\because h_{i} \prime_{i}: 11 .
$$

then, the class of each term in the expansion of $\delta L_{i}, \delta \xi_{j}$ and $\delta \eta_{j}$ relative to any divisor is not less than $\frac{1}{2}$. In the expansion of $\delta \eta_{1}$, the class of each term is not less than zero.

In order to prove this theoren, we first note that the theorem is valid for all the first order perturations. In this case, $m=1$ and $p+q \leqslant 1$ for the expansions of $\delta_{1} L_{1}, \delta_{i}$ and $\delta_{1} \eta_{1}$, and $m=10$, $p=0$ and $q \leqslant 2$ for the expansion of $\delta_{1} \lambda_{i}$.

Let us assume that the theorem is valid for all of the perturbations that have orders less than or equal to $m$. We then show that the theorem will also be valid for all the $(m+1)$ - order perturbations. We make use of equations (22) and (23) to calculate the ( $m+1$ ) - order perturbations in terms of the m-order ones. We first of all obtain

Each of the partial derivative of the function $R$; involved in this expression, can be expanded in a series of the type:

$$
\sum D_{0} n,
$$

In which we dencte by $D_{0}$ the second, third, ... derivatives in which the elements are replaced by their initial values: It has allready been pointed out in section 30 that the quantity qUAdafly be expanded in a series consisting entirely of periodic terms. For these terms, $m=1, p=0$ and $q=0$ since they have not been obtained as a result of an integration. In other words, the class of each term of the expansion of $D_{0}$ equals to unity. The quantity $R$ is a product of positive integral powers of $\delta L_{i}, \delta \lambda_{i}, \delta \mathcal{Z}_{j}$ and $\delta \eta_{j}$ evaluated up to terms having masses of m-order inclusively. Since the product of each two terms can yield only terms having classes less than the original two, then the expansion of each of the derivatives

will involve only terms having classes $\geqslant 1$.
In the integration, the class of a term relative to the mass does not change. However, the value of one of the coefficients $p$ or $q$ may be increased by one unit. Therefore, the classes of the terms of expansion (26) will be $\geqslant \frac{1}{2}$. Hence, the theorem holds for forms of the $m+1$ order.

Now, we investigate the class of the $(m+1)$-order terms in the expansion of the perturbation

$$
\begin{equation*}
i_{i}=-\underline{v} C_{i+} \int_{i}^{t} d s \int_{i}^{t} d R_{i} d t-\int_{i}^{!} d R_{i} d t-\int_{i}^{0} d \| d \tag{28}
\end{equation*}
$$

The integrands of the first two terms in the right-hand side may be expanded in series of the type just considered. Noting that the double Integration increases the sum $p+q$ by two units, we conclude that these two terms can only lead to terms having classes $\geqslant 0$. The partial derivative $\partial \Phi / \partial L_{i}$, involved in the third term of equation (27), consists of terms at least of the second order in $\delta L_{i}$. The product of two or more qua.tities $\delta L_{i}$ consists only of terms having classes $\geqslant 1$, since the class of each term in $\delta L_{1}$ is $\geqslant \frac{1}{2}$ as shown above. The
integration with increase the sum $p+q$ by one unit, yet the terms of the expansion will still have a class $\geqslant \frac{1}{2}$.

The theorem is thus proven. We note that the double integration yields the least-class term in $\delta \lambda_{i}$.

## 33. The least-class perturbations

Let us consider the structure of those terms, the class of which relicive to some given divisor

$$
r_{0} \quad \Xi x_{n} n
$$

is the least. We show that all of the terms of class $\frac{1}{2}$ in the expression of $\delta L, \delta \xi$ and $\delta \eta$ as well as terms of class zero in the expansion of $\delta \lambda$. are of the following form

$$
\begin{equation*}
N^{\prime \prime} \cdot 0 \cdot(1+1,1 \neq \|) \tag{29}
\end{equation*}
$$

where $\beta$ is an integer. We assume that this holds for perturbations evaluated up to the m-order inclusively, and then show that the same form can represent the least-class terms of the $(m+1)$ - order. We refer to equations (26) and see under which conditions can terms of class $\frac{1}{2}$ be obtained in the right-hand side. First of all, it is necessary that the term under consideration

$$
\begin{equation*}
A l^{p} \cos \left(\|: H^{\prime}\right. \tag{30}
\end{equation*}
$$

involved in the expression of the corresponding derivative $\partial R / \partial \lambda_{i}$, $\partial R \backslash \partial \eta_{j}$ or $\partial R / \partial \xi_{j}$, has a class equal to unity. Indeed, each of these derivatives may be exprnded in a series of the type (27), where the factor $D_{0}$ consists of terms of class unity as we have already seen. Henne, the class of the term (30) cannot be less than unity.

Furthermore, it is necessary that the integration of the term (30) decreases its class by $\frac{1 / 2}{2}$. This can only take place in two cases; 1) if $v=0$, then the integration increases by one unit the exponent $p$, and if $\nu=\beta \nu_{s}$, where $\beta$ is an integer, then the power $q$ of the devisor $\nu_{c}$
increases by one unit after integration.
Hence, all terms of class $\frac{1}{2}$ in the expressions of the perturtations $\delta L_{i}, \delta Z_{j}$ and $\delta \eta_{j}$ must have the form (29). We now show that in order to obtain all the terms of class $\frac{1}{2}$ in the expressions of $\delta L_{i}$, $\delta \eta_{j}$ and $\delta \Sigma_{j}$, it ig sufficient to consider those terms of the perturbation function $R$. the arguments of which are 'multiplies of

$$
1=v_{1}^{\prime \prime} a_{1} .
$$

We again consider expression (27), in which we have denoted:by $D_{0}$ those pari-al derivatives of the perturbation function $R$, in which the lements $L_{i}, \lambda_{i}, \varepsilon_{j}$ and $\eta_{j}$ are replaced by their initial values $L_{i}^{0}, \lambda_{i}^{0}+\because:-$ $n_{i} t, \varepsilon_{j}^{o}$ and $\mathcal{Z}_{j}^{o}$. In other words, the quantity $D_{o}$ consists of terms of the fcllowing type:

$$
\begin{equation*}
H_{0} \text { cus }\left(v / ; H_{u}\right) \text {, } \tag{31}
\end{equation*}
$$

where

$$
v: \because A_{t} n_{i}
$$

Here, $H_{o}$ is a constant while $B_{o}$ is a function of $L_{i}^{O}, \sum_{j}^{o}$ and $\eta_{j}^{0}$. This term is obtained by the term

$$
\begin{equation*}
\operatorname{li} \cos (\underline{2} k, i, \mid H) \tag{32}
\end{equation*}
$$

involved in the expansion of che function $R$ and, subsequently, ;ubstituting the above-mentioned initial values.

The factor $K$ in the expansion (27) is a product of non-negative powers of $\delta L_{i}, \delta \lambda_{i}, \delta \Sigma_{j}$ and $\delta \eta_{j}$ evaluated inclusively up to the m-order terms relative to the masses. In the frame of our assumftions, the terms of factor $R$ that have the least class (zero) will assume form (29) since they are obtained as a product of the least class terms in the expressions of $\delta L_{i}, E_{1}, \ldots$

Hence, the least-class terms in the partial derivatives $\partial \mathrm{R} / \partial \mathcal{E}_{1}$, $\partial R / \partial \eta_{i}$ and $\partial R / \partial \lambda_{i}$ qre obtained as the product of expressions
(29) and (30). Their arguments will thus have the form $\therefore$

$$
\left(\xi v_{0} \cdot v\right) t+\text { const. }
$$

We have already seen that in order to obtain terms having class $\frac{1}{2}$ after integration, each such argument should be of the form

$$
i^{2}, t+\text { corst. }
$$

Consequently,

$$
v= \pm(y-3) v
$$

or

$$
v=\sigma_{v} .
$$

where $G$ is an integer. Therefore

$$
A_{i}=i k_{1}^{k_{1}},
$$

Arcurdingly, terms (32) of the function R which lead to least-class perturbations. will have arguments of the form

$$
\underline{V}_{1} i_{1}+H=\Sigma_{1}-H .
$$

i.e. all multiplied by $\theta$. This is what has been required to prove.

It remains for us to investigate the structure of the least-order terms of the expansion of $\delta \lambda_{i}$. The easiest manner by which these terms are obtained is also required to be shown. We start by considering formula (28). In the previous section, we have seen that the least-class (zero) terms in the expression of $\delta \lambda_{i}$ may be only obtained from the first term of formuld (28). Ia order to obtain these terms, wa take

$$
i i_{1} \cdots-\searrow C_{i} \int_{0}^{1} d t \int_{1}^{1} d k_{h} d t,
$$

or, considering equation (26),

$$
\begin{equation*}
i_{1}=-\left.\underline{v} C_{1}\right|_{v} ^{i} \dot{L_{4}} d t . \tag{3.3}
\end{equation*}
$$

We again use expression (27) for considering the partial derivative $\partial \mathrm{R} / \partial \lambda_{\mathrm{k}}$. Similar arguments show that the least possible class for
terms of this type is unity. Hence, in crder to obtain terms of class zero in the expression of $\delta \lambda_{i}$, it jis necessary chat the double integration increases the sum $p+q$ by two units. This is only possible if the arguments of these terms have the form $\beta \nu_{0} t+H$

Finally, we combine the differential equations which allow us to obtain the least-class terms of $\delta \mathrm{L}, \delta \xi, \delta \eta$ and $\delta \lambda$. We start by $\delta \lambda$. In order to evaluate the zero-class terms in $\delta \lambda$, we use formula (33). From this formula, it follows that

$$
\therefore \quad-V: I_{1}=\quad \dot{A}+-i
$$

Noting that $\lambda_{i}=n_{i} t+\lambda_{i}^{0}+\delta \lambda_{i}$, we obtain

$$
\therefore=:-\perp \sum_{i}, \quad \therefore,-L
$$

Assuming that (cf. section 30 )

$$
H_{y}=-C_{3}-N(L-1, N C, i \quad \therefore \quad L,-i
$$

where $C_{0}$ is a constant, we obtain

$$
\begin{aligned}
& : \quad \therefore: \\
& 6:=-\quad \because
\end{aligned}
$$

This equation yields only zero-class terms if the $L-L_{c}$ involved in the expression of the function $\Psi_{0}$ is understood as the aggregate of terms of class $\frac{1}{2}$.

We have already seen that in order to obtain terms of class $\frac{1 / 2}{2}$ in the expressions of $\delta L_{i}, \delta \sum_{j}$ and $\delta \eta_{j}$, it is sufficient to keep only terms of the function $R$, the arguments of which are of the type $\sigma \theta$. Let us denote the aggregate of such terms $t_{y} \Psi$. Using equations (26), we obtain the following equations

$$
\begin{aligned}
& \because \quad \because \quad \because \\
& \therefore=\therefore=- \\
& \begin{array}{ll}
\because & \therefore \\
\because & \therefore=-\therefore
\end{array}
\end{aligned}
$$

which define the terms of class $\frac{1}{2}$. The right-hand side of these equations can be further simplified. We have seen that, in order to obtain terms of class $\frac{1}{2}$, it is necessary to keep only the zero-rank terms of $R$ in the expressions (27) of the derivatives $\frac{\partial R}{\partial j}$, ... However the zero-class terms of $R$ can be obtained only by replacing the quantity $\delta \lambda_{i}$ by its zero-class terms, and putting

$$
\because-\quad \because=\quad \vdots=
$$

that is to say

$$
\therefore-\therefore \quad \vdots \ldots \cdot \quad=
$$

Since such substitution is already done in $D_{0}$, then denoting by
the results of this substitution into

$$
\begin{array}{lll}
: & \cdots & : \\
\therefore & : & 1
\end{array}
$$

we obtain

Noting that
we formulate our conclusions in the form of the following theorem

## Theorem:

In order to obtain the least-class percurbations, it is necessary to integrate the following equations:
34. The Delaunay-Hill method for calculating long-periodic perturbations

We assume that the main motion $n_{i}$ of the system of material points under consideration is such that the quantity

$$
v_{u}=\underline{V} R_{i}^{0} n_{1}
$$

is small. In this case, the perturbation of least class relative' to the divisor $v_{0}$ will be of particular interest, since the amplitudes of these perturbations will be particularly large. The theorem, given in the previous section, enables us to determine these perturbations independently for the others. Putting again

$$
g=\Sigma K_{0}^{\prime \prime} i_{i}
$$

and denoting by $\mathcal{W}$ the aggregate of the cerms of the expansion of the cuntion $R$, the arguments of which are multiples; of $\theta$, we can conclude that $\Psi$ depends only on $\theta, L_{i}, \sum_{j}$ and $\hat{\gamma}_{j}$. Consequently

$$
\Psi_{0,}\binom{d I_{j}}{d j}_{v},\binom{d 川^{\prime}}{d x_{i}}_{0}
$$

are functions only of $\theta$. This situation enables us to obtain the solution of the system (36) and (37), which defines the least class term, by means of quadratures. The first of equations (36) yields

$$
d L_{1} \cdots \|_{0}^{d_{0}}=d_{t}^{\prime}{ }_{d \|_{1}}^{k_{1}}
$$

Introduc $i g$ the auxiliary function $U$ by means of the relation

$$
d U=-\begin{aligned}
& d y_{0} \\
& d l
\end{aligned}
$$

we obtain from the previous equations

$$
\frac{d l_{1}}{d i} \quad k_{1}^{\prime \prime} \quad d!=0 .
$$

Let $U=0$ for $t=0$. Then, integrating the previous equation from $t=0$ to $t=t$ yields

$$
1.1: 1 . \quad 1:-
$$

Substituting these values of $\mathrm{L}_{1}$ finto equation (34), we obtain

$$
\because=\because-n t,-\cdots i!
$$

where $A, B$ and $C$ are constants. On the other hand, equations (36)
have the evident integral
defining the dependence of $U$ on $\theta$ in a closed form. This integral may be represented in the following way
from which it follows by expressing $U$ in terms of $\theta$ that

$$
A U^{\prime} \cdot b=1 \dot{B}: \cdots
$$

Since
or,
then we obtain the following relation between $\theta$ and t :

$$
t=\int^{\dot{A}+i+\cdots \cdots+i_{0}}
$$

We now consider the second of equations (36) which yields

$$
\begin{align*}
& a^{\prime}=\begin{array}{l}
\pi \\
i
\end{array} \quad \because . . \pm, i,-1,- \\
& =-\quad \text { V!'. }
\end{align*}
$$

Replacing the quantity $U$ by its value given by equation (39) and integrating, we obtain element $\lambda_{1}$ as a function of $\theta$. Similarl ${ }^{\prime}$, integrating equations (37) leads to expressions for the coordinates $\mathcal{E}_{j}$ and $\mathcal{F}_{j}$ in terms of functions of $\theta$. Equation (40) allows us to express the coordinates $\lambda_{i}, \Sigma_{j}$ and $\eta_{j}$ in terms of time $t$. This gives the complete solution of the three-body problem under consideration.

In this way, in order to obtain the perturbations of least class relative to the argument $\theta$, it is necessary to pick all of the terms, the arguments of which are multiples of $\theta$, out of the peiturbation function R. Replacing thequantities $L_{i}, \mathcal{E}_{1}$ and $\mathscr{Z}_{i}$ in the function $\Psi$, obtained in this way, by their initial values we obtain the function $\Psi_{0}$.

Equations (38) and (39) allow us to express the quantity $L_{i}$ in terms of the argument $\theta$. Integrating in the same way equations (41) and (37), we obtain $\lambda_{i}, \varepsilon_{j}$ and $\gamma_{j}$ as functions of $\theta$.

Finally, equation (41) defines the dependence of the argument $\theta$ on time $t$.

Delaunay was the first to note that it is possible to obtain all periodic perturbation by integrating those equations of motion, in which the perturbation function is replaced by some of its separate terms. We applied this method to construct the most complete analytical theory of lunar motion ${ }^{(1)}$.

Tisserand considerably simplified the Delaunay method by relating
(1) C. Delaunfy, Theorie du mouvement de la Lune, Memoires de 1'Academie des Science de Paris, 28 (1860), 29 (1867).
it to the general theory of canonical transformation ${ }^{(1)}$. On the other hand, Hill significantly generalized this method by showing how the terms of the perturbation function, the arguments of which were multiplea, of a given argument, could be taken into account ${ }^{(2)}$

The merhod obtained in this ray is known as the Delaunay - Hill Method. Finally, the work of Poincare ${ }^{(3)}$ helped in clarifying the main mathematical points of this method.
(1) F. Tisserand, Traite de Mathematique Celeste, 3, Ch XI, 1894.
(2) G.W. Hill, On the Extension of Delaunay's Method in the Lunar Theory to the General Problem of Planetary Motion, Transations of the American Mathem. Soc. 1, 1900, 205-242 = The collected Mathem. Works, 4, 1907.
(3) H. Poincare, Le methodes nouvelles de la Mecanique Celeste, 2, Ch. XIX, Paris 1893;
H. Poincare, Lecon de Mechanique Celeste, 1, Ch. XIII, Paris 1907.

CHAPTER VI

## SOME PARTICULAR CASES OF THE THREE

BODY PROBLEM

## 35. Introduction

In 1772, Lagrange was awarded a prize by the Paris Academy for his well-known memoir "Essai sur le problême des trois corps (Oeuvres, 6, 229-324)". Lagrange pointed out in the preface of this work, that this included a method for the solution of the three-iody problem, which was very different from all previouscontributions. This method was shown by Lagrange to consist in reducing the determination of the relative coordinates of the three bodies, which requires the integration of a twelve-order system, to the determination of the sides of the triangle formed by the three bodies. This requires the integration of a 7-order system consisting of two second-order equations and one third-order equation. These equations involve two arbitrary constants introduced by the kinetic energy integral and the integral of areas. Accordingly, the mutual distances of the three bodies will depend on nine arbitrary constants. When the mutual distances are known, the determination of the relative coordinates which introduces another three arbitrary constants is quite simple.

Eliminating time from theabove-mentioned 7-order system, we finally reduce the solution of the problem to the integration of a 6-order system. . The reduction, perfermed by Lagrange is essentially identical to that indicated in section 2. However, the special form in which Lagrange obtained the equations of motion enabled him to formulate and solve the problem of finding all the three body types of motion when their mutical distances always keep constant ratios. These types of motion are called Lagrangean. We shall see that a Lagrangean motion will
necessarily be coplanar.
If we have to proceed along the same path as Lagrange, when studying the lagrangean motion, we then have to initially deduce the differential equations that define the mutual distances of the three bodies. However, as Lagrange himself pointed out (loc. cit., page 431 ), the particular case under consideration can be resolved in a much simpler way if we, beforehand, assume that the three bodies are moving in an invariable plane. Indeed, restricting the problem by this subsidiary condition, Laplace was able to derive a very simple deduction for the aqations of the lagrangian motion ${ }^{\text {(1) }}$.

Lagrange assumed that the solution of the general problem, i.e. without restricting it to coplaner motions alone, is indispensibly connected to several difficulties. Andoyer and Caratheodory ${ }^{(2)}$ proved that this was not true. They developed a simple method for obtaining the general solution of the problem suggested by Lagrange. We shall give the details of this method in the next sections. This is very interesting since it can very easily be extended to the n-body problem.

By this method, it is easy to show that the lagrangean motion of n-bodies will also take place only in an invariable plane, except in some almost trivial_cases,_when the_motion proceeds along straight lines
(1) 'Laplace, Mechaniqưe céle'ste; Sécóñe partie, Livee X, Ch. VI (Oeuvres, 4). Laplace': method is explained in: Charlier, Die Mechanik des Himmels, 2, 89-102, 1907; A simple geomet fical method for obtaining the results of Laplace is given by C.D. Cernyj in the paper: Geometrische Losung zweier spezieller Falle des problems der drèi Körper, Astr. Nachr. 171, 1906, 129-136.
(2) H. Andoyer, Sur l'équilibre relatif de $n$ corps, Bulletin Astr., 23, 50-59, 1906.
C. Caratheodory, Uber die strenge Losungen des Dreikorproblems, Sitzungs berichte der math. naturwiss. Abtellung der I yerschen Akademie der Wiss. zu Munchen, 1933, 257-276.
passing through a common centre of gravity.
36. Equations of Lagrangean Motion. The Case of Moncollinear Motion

We shall construct the equation of relative motion of three bodies, starting with the assumption that the ratios of the mutual distances of these bodies remain constant. We denote by $m_{1}, m_{2}$ and $m_{3}$ the masses of points $P_{1}, P_{2}$ and $P_{3}$ and by $x_{i}, y_{i}$ and $z_{i}$ the coordinates of point $P_{i}$. We take the origin of the coordinate at the centre of gravity 0 , and the $x y$ plane as the $p$ ane of the tri-angle $P_{1} P_{2} P_{3}$.

The distance from the centre of gravity to the vertices $P_{1}, P_{2}$ and $P_{3}$ are proportional to the dimensions of the triangle. Hence we can use in the case of a lagrangian motion, a rotating coordinate system and put

$$
\begin{gather*}
x_{1}=a_{1}: y_{1}=b_{1}, b_{1} \quad z_{1}==0  \tag{1}\\
(i=1,2,3,
\end{gather*}
$$

where $\rho=\varphi(t)$ is a properly defined function of time, and $a_{i}$ and $b_{i}$ are constants. We denote by $\mathcal{p}, \underline{q}$ and $\underline{r}$ the components of the angular velocity of the system along the axes $x, y$ and $z$. The components of the velocity of a point, whose coordinates are $x, y$ and 2 , are well-known and equal to

$$
\dot{x}-y \mathrm{r}, \mathrm{zq}, \quad \dot{y}-z \mathrm{p} \mid x \mathrm{r}, \quad z-x q+y \mathrm{p} .
$$

Consequently, the components of acceleration are given by

$$
\begin{aligned}
& { }_{d f}^{d}(x-y r+z q) \cdots r(y-z p \mid-x r) \mid q(z-x q: y p) \\
& \underset{d i}{d}(\dot{y}-z p: x r)-p(z \cdots x q \cdot y p)+r(x \quad y r \cdot z q) \\
& { }_{d f}^{d}(z-x q+y p) \quad q(x \quad y r+z q): p(y \quad z p ; x p) .
\end{aligned}
$$

Substituting the values (1) of the coordinates under consideration Into these expressions, and noting that the components of acceleration of point $P_{1}$ caused by the attraction of the other two points, are equal to

$$
A_{i}{ }^{2}, \quad H, p^{2}=0 .
$$

where $A_{i}$ and $B_{i}$ are constant factors depending on $a_{1}, a_{2}, \ldots, b_{3}, m_{1}$, $m_{2}, m_{3}$ and the constant of gravitation, we obtain the following equation of motion of the point

$$
\begin{align*}
& u_{1}\left|-4: \div(q-p r)_{i}\right|+b_{1}\left|2 p:+(p \mid-4 r)_{i}\right| 0 . \tag{3}
\end{align*}
$$

In the case of a collinear motion, in which the three points $P_{1}, P_{2}$ and $P_{3}$ are always on one straight line, or in other words, when

$$
\begin{aligned}
& a_{1} \\
& b_{1}
\end{aligned}=\begin{aligned}
& a_{2}=\begin{array}{l}
a_{3} \\
b_{2} \\
b_{3}
\end{array}, ~, ~
\end{aligned}
$$

will not be considered in this section. Hence, equations (2), (3) and (4) lead to the following relations

$$
\begin{align*}
\dot{p-\left(q^{2}+f-r^{2}\right)-A^{\prime} q^{-2} ;} & 2 r \dot{p}+(\mathbf{r} \quad p q) p=B^{\prime} p^{-}  \tag{2'}\\
\dot{p}\left(p^{2}+r^{\prime}\right) \quad A^{\prime \prime} p^{-z} ; & 2 r \dot{p}+(r+p q) p=B^{\prime \prime} p^{-z} \\
2 q \dot{p}+(\dot{q}-p r) p=0 ; & 2 p \dot{p}+(p+q r) p=0,
\end{align*}
$$

where $A^{\prime}, B^{\prime}, A^{\prime \prime}$ and $B^{\prime}$ are new constants though they can be expressed In terms $A_{i}, B_{i}, a_{i}$ and $b_{i}$. The term-by-term substration of equation (2') and (3') yields
from which it follows that

$$
\begin{equation*}
\rho=q p^{2}, \quad q=\beta p^{-3}, \tag{5}
\end{equation*}
$$

where $\alpha$ and $\beta$ are constants. Substituting these values into equation
(4), we obtain

$$
\begin{equation*}
4 p-2 a r p=0, \quad a+2 \dot{p} r b=0, \tag{6}
\end{equation*}
$$

from which it folluws that

$$
\begin{equation*}
\left(a^{2}+p^{2}\right) p=0 . \tag{7}
\end{equation*}
$$

It is now easy to show that the moticn proceeds in the invariable plane. For this purpose, we show that we can obtain $p=q=0$ by means of an appropriate choice of the axes $0 x$ and $0 y$. If $\alpha^{2}+\beta^{\alpha}=0$, then $\because=\beta=0$, and hence the relation (5) yields

$$
p-q=0 \text {, }
$$

which proves that the plane $P_{1} P_{2} P_{3}$ is invariable.
We now investigate whether the sum $\alpha^{2}+\beta^{2}$ cannot be equal to zero. If this can take place, we may then conclude from equation (7) that $\rho^{\circ}=0$. Since $\rho \neq 0$, then the relations (6) yield $\underline{r}=0$. Once $\dot{\rho}=0$ and $\underline{r}=0$, it then follows from equations (4') that

$$
p=c o n i s 1, \quad q \text { const. }
$$

We add the vectors $p$ and $q$ which are directed along the $x$ - and $y$-axes, and take the direction $\checkmark f$ the resulting vector as a new $x$-axis. In this case, we get $q=0$ in the new coordinate system and, hence, equations (2) yield $A_{i}=0$. In other words, the projections of all of the forces on the axis $0 x$ vanish, so that all of the forces will be parallel to the axis Oy. This is impossiole since we agreed to only consider the case in..which the three bodies are not located along one straight line.

We take the plane $0 x y$ of the fixed coordinate system as the invariable plane in which the motion takes place and draw the perpendicular axis Oz rrom the centre of gravity of the system. The motion of the points $P_{1}$ in the case under consideration will then consist of the rotation of triangle $P_{1} P_{2} P_{3}$ as a whole, around axis $O z$ with an angular velocity $\underline{r}$
and with the motion of each point $P_{i}$ along the ray $\cap P_{i}$. Hence, the areal velocity of point $P_{i}$ will differ only by a constant factor from the quantity $r^{2} \Phi$, so that the integral of area for the axis $\mathrm{O}_{2}$ will give

$$
r_{1}^{2}=\text { cont. }
$$

from which it follows that

$$
2 r_{p}+r_{p}=0 .
$$

In this way, equations (2) and (3) become
which yield

These equations show that the resultant of all the forces that act on each of the points $P_{i}$ passes through the centre of gravity of the system. It is now already not difficult to determine the form of the triangle $P_{1} P_{2} P_{3}$ (figure 1). Denoting, as previously, the sides of this triangle by

li. . 1

$$
\begin{array}{r}
P_{1} P_{2}-1_{1: 0} \\
P_{1} P_{1} P_{3}-I_{2} \\
S_{23}
\end{array}
$$

We obtain the following expressions for the three accelerations that points $P_{2}$ and $P_{3}$ produce on point $P_{1}$.

$$
\begin{aligned}
& \overrightarrow{M_{1} h} k: M_{1} \text { 。 }
\end{aligned}
$$

We determine the angle $\alpha$, which is formed between the geometrical sum $P_{1} R$ of these axcelerations and the straight 1 in $P_{1} P_{2}$. For this perpose, we project the acceleration $P_{1} B$ on the axes $\left.P_{1}\right\}$ and $P_{2} \eta$ into the components

$$
\overrightarrow{P_{1} A} \because \overrightarrow{P_{1} A} \cos { }_{F 1} \quad \overrightarrow{P_{1}} A \cdot \overrightarrow{P_{1} H}, m F_{1} \quad \text { ORIGINAL PAGE E }
$$ OF POOR QUA

where $\mathscr{\varphi}_{1}$ is the angle $F_{2} P_{1} P_{3}$. Then

On the other hand, denoting by $\xi_{0}$ and $\xi_{0}$ the coordinates of the centre of gravity of the system, we obtain

On account of equations (8), the scraight line $P_{1} R$ passes through point 0. We therefore evidently obtain the following equation

Taking into account that $\mathrm{m}_{2} \neq \mathrm{D}, \mathrm{m}_{3} \neq 0$ and $\sin \varphi_{1} \neq 0$, we obtain from the previous equation

$$
\lambda_{1}^{3} \quad \lambda_{1}
$$

or

$$
د_{1:}=د_{1} .
$$

If $m_{1} \neq 0$, we can then prove the equality of the other two sides of the triangle in an exactly similar way. We will then finally obtain

$$
J_{12} \cdot J_{1:} \cdot J_{13}
$$

The case in which two of the masses $m_{1}, m_{2}$ and $m_{3}$ are infinitesimal will not be considered here, since it is trivial.

It remains for $u$ to only
consider the case in which one of the


Figz.
masses, say $m_{1}$, is infinitesimal. In this case, the previous approach will be valid so far as it concerns the vertices $P_{2}$ and $P_{3}$. We can only conclude that the sides $\Delta_{12}$ and $\Delta_{13}$ are equal. It is however pasy to see that the triangle $P_{1} P_{2} P_{3}$ will be equilateral.

Thus let $\mathrm{m}_{1}=0$. We draw the coordinate axes as shown in figure 2, taking into account that the contre of gravity in this case is located on the straight line $\mathrm{P}_{2} \mathrm{P}_{3}$. The coordinates of points $\mathrm{P}_{1}, \mathrm{P}_{2}$ and $\mathrm{P}_{3}$ will be equal to

$$
\left(a_{1}, b_{1}^{\prime \prime}\right), \quad\left(u_{2} \therefore, 0\right) \|\left(a_{2} \therefore 0\right) .
$$

Since the origin of the coordinates is at the centre of gravity, then

$$
\begin{array}{cc}
-a_{2} & m \\
a_{1} & m \\
m
\end{array}
$$

On the other hand, since the sides $P_{1} P_{2}$ and $P_{1} P_{3}$ are equal, then

$$
\because a_{1} \quad a \quad a .
$$

Hence denoting by $\varphi$ each of the angles at the base of the triangle, we easily find

$$
b_{1} \quad\left(a_{1}-a_{2}\right) 4_{2} \because \cdots-u_{i} m_{2}+1 m_{3} \text { li: }
$$

It folows from equation (8) that

$$
d: \quad i
$$

In the pre sent case,
as it can be easily seen from figure 2. Hence

$$
1, د_{1}, \cos :=A_{2}
$$

Since

$$
\Delta_{: \because}=\Delta_{1}: \cos \beta .
$$

we finally then obtain

$$
\cos \cdot \bar{r}=\frac{1}{y} \quad \because=(0)^{\prime}
$$

Hence, when three material points move under the action of their mutual attraction in such a way, that the distances between them keep constant ratios, then if these points are not on a straight line, they will always form an equilateral triangle. The plane of this triangle will keep an invariable position in space.

Liet us assume that the initial positions of two of the three points, say $P_{1}$ and $P_{2}$, are fixed, similarly as the plane in which the motion is taking place. In this case, in order to obtain a motion of the type under consideration, we have to place the third point at the vertex of one of the two equilateral triangles that can be formed at both sides of $P_{1} P_{2}$,
i.e. at one of pints $\mathrm{L}_{4}$ and $\mathrm{L}_{5}$ of


Fig. 3.
figure 3, in which the points $P_{1}$ and $P_{2}$ are denoted by $m$ and $m^{\prime}$.

These points are called the triangular points of libration ${ }^{(1)}$.

In conclusion, we show that, in in the case under consideration, the motion of each of the points $P_{i}$ relative to the common centre of inertia 0 proceeds in such a way, as if each of these
points was attracted by armass equal to the masses of the two other points,
(1) These points are also called the equilateral points of libration. Guilder cailed these points, as well as the points that we shall. Consider later, by the centres of libration.
located at 0 . In other words, the motion proceeds according to the generalized laws of Kepler.

We now turn to equations (23) of section 6 which define the motion relative to the centre of gravity. Dencting by $\Delta$ the common value of the presently equal distances $\Delta_{i}$, and taking into account that

$$
m_{1} x_{1}+m_{1} x_{2}+m_{1} x_{2} \quad 0 .
$$

we obtain the following equations of motion for the point $P_{i}$ :

where

$$
A: \quad C_{1}\left(x_{1} \cdot y, \therefore \because\right) .
$$

in which $c_{1}$ is constant. These equacions are identical in form to the equations of motion of the two-body problem. This proves the validity of our assumption.
37. The case of collinear lagcangian motion

We now consider the case in which the three bodies $\mathrm{P}_{1}, \mathrm{P}_{2}$ and $\mathrm{P}_{3}$ are always on one straight line. Taking this line as the x-axis, we obtain the following expression for the coordinates of point $\mathrm{P}_{\mathbf{i}}$

$$
x_{1} \quad a_{1} p(0) \quad y_{1}=0, \quad z=\cdots 0 .
$$

so that the problem is reduced to the determination of the function $\rho(t)$ and the constants $a_{1}, a_{2}$ and $a_{3}$.

Since all of the forces are along the $0 x$, axis we can choose a coordinate system which does not rotate around this axis, i.e. we can
consider $p=0$. Taking into account that in the present case $b_{i}=0$ and $B_{1}=0$, we obtain, from equations (3) and (4)

$$
\begin{equation*}
2 \mathbf{r} p+\dot{r}!=0, \quad 2 \mathbf{q} ;+\mathbf{q} p^{-=0} . \tag{9}
\end{equation*}
$$

We assume that $q \neq 0$. Multiplying these equations by $q$ and $\underline{r}$ and substracting them term-by-term from one another, we obtain

$$
\mathbf{r q}-\dot{q} \mathbf{r}=0
$$

Consequently,

$$
r=-\lambda q
$$

where $A$ is a constant. Hence, taking the direction of the geometrical sum of the vectors $\underline{r}$ and $q$ as the new $0 z$ axis, we will obtain $q=0$. Thus, we can always consider that the $O y$ and $0 z$ axes are chosen in such a way that $g=0$.

Consequently, the motion of the straight line $P_{1} P_{2} P_{3}$ in space will consist of a rotation of this straight line around the Oz axis with an angular velocity equal to $\underline{r}$. Integrating the first of relation (9), we obtain the integral of area

$$
\begin{equation*}
\mathbf{r}_{\ddot{r}^{\prime \prime}}=\text { const. } \tag{10}
\end{equation*}
$$

In che present case, equations (2) read

$$
j-r^{2} p \quad \frac{A_{1}}{a_{1}} p:
$$

and hence yield

$$
\begin{aligned}
& A_{1} \\
& d_{1}
\end{aligned} \quad \begin{aligned}
& A_{1} \\
& d_{1}
\end{aligned}=\frac{A_{1}}{a_{1}} .
$$

Assumin; that
and putting

$$
\because-u
$$

we easily obtain
since

$$
\begin{aligned}
& A \quad-m_{1}\left(a_{3}-a .1-2 m_{1}\left(a_{3}-a_{1}\right)\right.
\end{aligned}
$$

Since the origin of the coordinates is at the centre of gravity of the system, then

$$
m_{:} a_{1}+m_{2} c_{1}-m_{3} a_{3}=0 .
$$

If follows from equations (11) that $a_{1}, a_{2}$ and $a_{3}$ are proportional to each other. Taking this into account, we obtain the following equation for the determination of $z$ :

This equation has at least one positive root since the left-hand side has different signs at $z=0$ and $z=+\infty$. On the other hand, according to a theorem by Descartes, equation (13) can have no more than one positive root since its coefficients change sign only once. Hence, whatever the values of the masses are, we only obtain one positive value for 2 . Equations (11) enable us to obtain the ratios $a_{1}: a_{2}: a_{3}$ which correspond to this value of $z$.

The three different masses can be located on a straight lines by three different manners. This leads to three collinear lagrangean motions.

In astronomical applications, we denote the masses by $m, m^{\prime}$ and $m^{\prime}$ and assume that the mass $m$ is very large (the mass of the sum), the mass $m^{\prime}$ is small (the mass of a planet) and the mass $m^{\prime \prime}$ is very small (the mass of a planetcaid, comet, metenrite, etc.). Putting into euqation (13) $m_{1}=m, m_{2}=m^{\prime}$ and $m_{3}=m^{\prime \prime}$, we obtain an equation, the positive root of which is very small. Keeping the most important terms in this equation, we obtain

$$
\left(3 m: m^{\prime}\right) z^{2} \quad\left(n a^{\prime} \cdot m^{\prime}\right)=0
$$

from which it follows that

$$
z=\binom{\left.m^{\prime}+m^{\prime \prime}\right)^{\prime}}{3 m+m^{\prime}}^{\prime}
$$

In this manner, denoting the distance between the planet and the sum by $r$, and the distance between the astroid and planet by $r$ ', we obtain

$$
\begin{equation*}
r^{\prime}=r\binom{m^{\prime}+m^{\prime \prime}}{\therefore m^{\prime}+m^{\prime}}^{\prime} \tag{14}
\end{equation*}
$$

In the case, when $m_{1}=m, m_{2}=m^{\prime}$ and $m_{3}=m^{\prime}$, i.e. when the planetoid is between the planet and the sum, we ortain

$$
\cdots, \left.\begin{array}{cc}
m & m \\
\ldots m & m
\end{array} \right\rvert\,
$$

Finally, if the planet and planetoid are at different sides of the sun, $s$ that $m_{1}=m^{\prime}, m_{2}=m$ and $m_{3}=m^{\prime}$, then equation (13) reads

$$
f(:) \cdot 11
$$

where

Evidently, the positive root of this equation slightly Giffers from unity. Hence we may approximate the value of this root by

$$
2 \quad 1-\mu(1)=\begin{gathered}
7-m^{\prime} \cdot m^{\prime} \\
12 m-2(1)
\end{gathered}
$$

## Consequent1y

$$
r^{\prime \prime}=r\left(\begin{array}{cc}
1- & 7 m^{\prime}-m^{\prime \prime}  \tag{16}\\
1!m+\cdot+20 m^{\prime} & 3 m^{\prime \prime}
\end{array}\right),
$$

where $r^{\prime \prime}$ is the distance from the planetoid to the sun.
Each of the three positions in which the third mass could be located on the straight line joining the other two masses $m$ and $m^{\prime}$, will be called a collinear point of libration. These positions are denoted by $\mathrm{L}_{1}, \mathrm{~L}_{2}$ and $\mathrm{L}_{3}$ in figure 3 .

When the magnitudes of the masses satisfy the above condition, the positions of the collinear points of librations will be defined by formulae (14), (15) and (16). If the mass $\mathrm{m}^{\prime \prime}$ is negligibly small in comparison with the two others, ther the position of the five points of libration will be given in the first approximation by

$$
\begin{aligned}
& \text { for } l, 1 \quad r^{\prime \prime} \quad r \quad r\binom{m^{\prime}}{\vdots m}^{\prime \prime} \\
& \text { • } 1.3 \quad r^{\prime \prime} \quad r \quad i r\left(\begin{array}{c}
m^{\prime} \\
. m \\
i m
\end{array}\right)^{\prime} \\
& \text { - l. } \quad r^{\prime}=r-1 \quad \begin{array}{c}
7!m^{\prime} \\
12 m+2 u m
\end{array} \\
& \text { - l., and l., } r^{\prime} \text {.. } r^{\prime} 1 \text {. }
\end{aligned}
$$

The following table gives the positions of the first three libration points for the different planets of the solar system. The table gives the ralues of the distances $r^{\prime \prime}$ of the points of libration from the sum, expressed in fractions of the radius vectors of the planets

| Mercury | 0.9966 | 1.0034 | $1-0.000 .000 .07$ |
| :--- | :--- | :--- | :--- |
| Venus | 0.9907 | 1.0093 | $1-0.000 .001 .43$ |
| Earth | 0.9899 | 1.0101 | $1-0.000 .001 .78$ |
| Mars | 0.9952 | $1 . \underbrace{}_{\text {\& }}$ | $1-0.000 .000 .19$ |
| Jupiter | 0.9332 | 1.0698 | $1-0.000 .557$ |
| Saturn | 0.9550 | 1.0164 | $1-0.000 .167$ |
| Uranus | 0.9758 | 1.0216 | $1-0.000 .026$ |
| Neptune | 0.9743 | 1.0261 | $1-0.000 .030$ |

It is interesting to note that all of the planets have sattelites at distances much smaller than the distances to the libration points $L_{1}$ and $L_{2}$. For example, the distance from the earth to the moon is approximately four times smaller than the distance to any of these points.

After obtaining the value of $z$, and from equations (11) the values of the ratios $a_{1}: a_{2}: a_{3}$, we can start to study the motion of point $p_{1}$. For this purpose, we can apply equations (10) and (10'). In order to obtain the function $\mathcal{S}(t)$, we give one of the constants $a_{i}$ an arbitrary nonvanishing value. We put, for example, $a_{1}=0$. Then, the above mentioned equations yield.

$$
\mathbf{r} p^{2} \quad C_{,} \quad \because \because \cdots \mathbf{r}^{\prime} \because \quad A_{1} .
$$

where $C$ is an arbitrary constant, while $A_{1}$ is defined by formulae (12). We denote by $u$ the angle between the straight line $P_{1} P_{2} P_{3}$ and an arbitrarily given direction in the plane $x 0 y$. Since $\underline{r}$ is equal to $d u / d t$, then the equations of motion will have the final form

$$
f^{2} \frac{d u}{d t}=C, \quad f^{2} \frac{d^{2}}{d t^{2}}-\binom{d u}{d t}^{3} p^{\prime}==A_{1} .
$$

from which it follows that

$$
\frac{d: p}{d / t}=C^{\prime} p^{+}-A_{1}^{\prime}:
$$

Multiplying by $2 \frac{d \rho}{d t}$ and integrating, we obtain

$$
\binom{d t}{d t}^{2}=h \quad 2 A_{1 p^{-1}}-C^{2} p^{-2}
$$

where $h$ is a new constant.

In order to obtain an equation for the orbit, we eliminate dit by means of the integral of area. We obtain

$$
\rho^{\prime}\binom{d_{1}}{d!}^{2}=h++^{A}\left(\begin{array}{lll}
1 \\
C
\end{array}-\left(\begin{array}{cc}
C & A_{1} \\
k & \\
C
\end{array}\right)^{2}\right.
$$

or

$$
\binom{d_{i}}{d u}^{2} \cdots h+\frac{A_{i}}{\vdots} s^{3}
$$

where

$$
s=\frac{0}{6}+\frac{A_{1}}{i}
$$

The integration of the latter term yields
where

$$
\text { e. } v^{\prime} 1+h C^{-} A_{1}^{-}, \quad p \cdots-A_{1}^{\prime} C^{\prime}
$$

and $\omega$ is a new arbitrary constant.

Thus, the motion of any of the points $P_{i}$ around the common centre of gravity proceeds by a conic section which satisfies the law of areas. In other words, this motion proceeds according to the laws of Kepler.
38. The Equations of Motion, The Jacobi Integral
where $a_{1}+a_{2}$ is the seminajor axis of the orbit along which one of the bodies $S$ and $J$ moves under the action of the mutual gravitation.

We choose the positive direction of the axis $0 y$ so that $n$ is always positive. Let $x, y$ and $z$ be the coordinates of point $P$. Since the coordinate system rotates with angular velocity $n$ around the $z$ axis, then the components of the absolute velocity of this point are

$$
\dot{x}-n y, \quad \dot{y}+n x, \quad \dot{z}
$$

If we denote by $m_{0}$ the mass of point $P$, the kinetic energy of this point is given by

$$
\therefore \quad!\quad \because \mid 11 \quad .1, \quad 11: \cdots \quad: 1
$$

Applying the Lagrange equations ( $S$ 19), we obtain
where $U$ is the force fun $f=n$ acting on point $P$ divided by $m_{0}$. In the present case, point $P$ moves under the action of the gravitation of points $S$ and J. Therefore

$$
\text { ll. } \begin{gathered}
k-m_{1}, \\
r_{1} \\
\\
r_{2} \\
r_{2}
\end{gathered}
$$

Assuming that

$$
\because \quad 1, n:\left(x^{2} \quad j\right) ; k \cdot\left(\begin{array}{cc}
m_{1} & m \\
r_{1} & r_{1}
\end{array}\right)
$$

$$
1 \therefore
$$

the equations of motion in the restricted problem of three bodics will be given by

$$
\begin{aligned}
& \therefore==\frac{111}{1.2}
\end{aligned}
$$

| 1-2'y |  | 19 |
| :---: | :---: | :---: |
|  |  | ${ }^{+} \cdot x$ |
| $\boldsymbol{y}$; 2m |  | 010 |
|  |  | dy |
| $\varepsilon$ |  | O! |
|  |  | d 2 |

Multiplying these equations by $x, y$ and $z$, adding and integrating, we obtain

$$
x+j:+2 \quad 2 \because \quad 0
$$

where $C$ is an arbitrary constant. This relation is known as the Jacobi integral. The constant $C$ will be called the Jacobi constant.

The Jacobi integral enables us to draw many important conclusions on the character $n f$ motion of point $P$. This will be now investigated. 39. The Surface of Zero-Velocity

Let us denote by $v$ the velncity of point $P$ relative to the moving coordinate system. We then write the Jacobi integral, given by equation (4), as follows

$$
\prime=: \because \because \quad \text { C.. }
$$

Using this relation, we are able to determine the relative velocity $v$ in each point in the rotating space, for all motions characterized by a given value of the Jacobi constant $C$. Inversely, if the constant $C$ and velocity $v$ are given, then this relation defines the locus of points of the rotating space, in which body $P$ can exist.

We consider the totality of motion of point $P$, for which the constant $C$ has a given value. Evidently, these motion are possible in the space region, in which $2 \Omega \cdots c \geqslant 0$, otherwise the velocity $v$ of body $P$ is imaginary. The surface

```
ごい!!!!!
```

defines the boundary between the regions in space，in which the motions corresponding to a given value of $C$ are possible and the regions in which these motions are impossible．This surface is called the surface of zero－velocity，since $v=0$ at each of its points．

In the following，we study the form of the surface of zero－ velocity for different values of $C$ ．We choose the units of length and time such that

$$
\text { SJ. } a_{1} \div a, \quad 1, k=1
$$

Using equation（1），we obtain

$$
n^{?}=m_{1} \cdot m_{2}
$$

Taking into account expression（2），we write equation（5）in the following way
where

$$
r_{1} \quad \|^{\prime}\left(x ; u_{1}\right): y: \quad: \quad r_{0} \quad \left\lvert\,\left(\begin{array}{ll}
1 & d_{0}
\end{array}\right)\right.: y^{\prime 2}+z^{2}
$$

The surface，represented by equation（6），evidently lies inside the cylinder

$$
\left(m_{1}-m_{1}\right)\left(x^{:} \vdots y^{2}\right) \quad(:
$$

and asymptctically approaches the cylinder when $z$ increases to infinity．

The equation of the curve resulting from the intersection of surface（6）with the plane $x 0 y$ ，is obtained by subu＋ituting $Z=0$ In equation（6）．This substitution yields

Let us assume that $C$ is a large number. In this case, equation (7) is satisfied by points of one of the following types:

1- Points, foi which the quantity $\mathrm{x}^{2}+\mathrm{y}^{2}$ is large. For such points, the second and third terms of equation (7) are small, so that this equation reads

$$
\begin{equation*}
x:+y: \quad(-\cdots \tag{8}
\end{equation*}
$$

where $\epsilon_{0}$ is a small positive quantity.
2- Points for which the radius vector.

$$
r: \quad \mid 1: \therefore) \cdot\}
$$

is sma11. For these points, the first and third terms of equation (7) are small, so that this equation reads

$$
\begin{array}{llll}
\therefore & a_{1} 1^{3} & +y & \left(\begin{array}{ll}
\therefore m_{1} \\
\therefore & r_{1}
\end{array}\right)^{\prime}
\end{array}
$$

where $\epsilon_{1}$ is a smaıl positive quantity
3- Points for which the distance to J, i.e.

$$
r_{2}-1\left(r-a_{2}\right)+y_{2}
$$

Is sufficiently small. For these points, equation (7) may be written as

$$
\left(\begin{array}{ll}
x & a_{i} \tag{1111}
\end{array}\right)_{1} y:=\binom{2 m_{2}}{C-i_{2}}^{2}
$$

Hence, for large values of $C$, curve (7) consists of three separate closed parts, each having a form slightly differing from a circle. The larger is mass $m_{1}$ as rollpared to $m_{2}$, the greater are the dimensions of curve (9) as compared to those of curve (10).

As C decreases the dimensions of curves (9) and (10) increase, and their forms become more and more stretched along the axis $0 x$. At some


Pac. 4.

## Figure 4

value $C=C_{1}$, these curves touch each other. At smaller values of $C$ we have two separate ovals and one curve enclosing points $s$ and $J$ (figure 4). On the other hand, when $C$ is decreased, the dimensions of curve (8) decrease. At some values $C=C_{2}$ and $C=C_{3}$, this curve touches the two internal curves, mentfoned just above. Subsequently, these curves are amalgamated.

At large values of $c$, the domain of the plane $x 0 y$, in which the motion of body $P$ is forbidden, consists of points external with respect to curve (9) and (10) and internal with respect to curve (8). At smaller values of $C$, this region consists only of points lying inside curves $C^{\prime \prime}$ (figure 4) which decrease when $C$ decreases and turn into points at some value $C=C_{4}$ and then completely disappear. Thus, at sufficient values of $C$, body $P$ will have the possibility of moving over all plane $x 0 y$.

Figure 4 gives a schematic representation for the curves of these curves for decreasing values of $c$, namely $c^{\prime}>c_{1}>c_{2}>c^{\prime \prime}$. of the curve that results by the intersection of surface (6) with the plane $x 0 z$. The equation of this curve is obtained by the substitution $y=0$ in equation (6). This substitution yields
where

$$
\left.r_{1} \mid \cdots, a_{1}\right):=r, \quad r \quad \cdots\left(i \quad a_{0}\right)-\infty
$$

When the value of $C$ is very large, this equation can be simplified by three different ways; either by making $x^{2}$ very large, or by making any of the quantities $r_{1}$ and $r_{2}$ very small. Accordingly, curve (11) consists of three separate parts, respectively defined by the following equations

$$
\begin{gathered}
\left(m_{1}: m_{1}\right):-i-: \\
r_{1}:=m_{1} \\
C-\varepsilon^{\prime} \\
r_{2} \quad \because-m_{1}
\end{gathered}
$$

This case is represented by curves $C^{\prime}$ in figure 5. On decreasing $C$, we pass again by the critical values $C_{1}, C_{2}$ and $C_{3}$ where different parts of curve (11) get into contact. Finally, when the value of the


Figure 5
Fi45



Jacobi consi:.ınt becomes safficiently small, e.g. C = C", curve (11) does not cut the Ox axis.

The intersection of surface (6) with the plane yOz is defined by

$$
\left(m_{1}+m_{2}\right) \cdot \frac{2 m_{1}}{r_{1}}+\frac{2 m_{2}}{r_{2}}=C .
$$

where

The corresponding curves are represented in figure 6 for the different values of $C$. The curves are ohtained on the assumption that mass $m_{1}$ is considerably larger than mass $m_{2}$.

The comparison between the three cross sections of surface (6), represented by figures 4,5 and 6, enable us to have a clear picture on the shape of this surface for the different values of the Jacobi constant C.

After this qualitative study of the surface (6), we turn into the study of specific points on this surface.
40. Specific Points on the Surfaces of Zero-Velocity

A specific point on the surface

$$
F(x, x, z) \quad 11
$$

is defined by the following equations

$$
\begin{array}{llllll}
d F & d r & d r & d r & \\
d x & -0 & d y & d z & d, & d z
\end{array}
$$

They can be solved in combination with the equation of the surface. For a surface of zero-velocity, defined by an equation of the type

$$
\begin{equation*}
\because \because \quad \therefore \quad . \quad 1, \tag{1111}
\end{equation*}
$$

where
the specific points are given by

$$
\begin{align*}
& \frac{d y}{d y}=\left(m_{1}+m_{3}\right) y-m_{1}^{m_{1} v}-\underset{r}{m_{2} y}=0  \tag{114}\\
& \stackrel{0!}{12}-\frac{m_{1} z}{r_{1}^{\prime}}-\frac{m_{2} z}{r!}-0 .
\end{align*}
$$

Solving equations (14), we find the coordinates of the specific points. Subsequently, we use equation (13) to find the corresponding values of the Jacobi constant C.

It is easy to find the mechanical meaning of the specific points. Comparing equations (14) with the equations of motion of body P , equations (3), we find that at each of the specific points not only

$$
\therefore=3=11
$$

but also

$$
\ddot{i} \ddot{y} \quad \ddot{<} 0 .
$$

Thus, once body $P$ arrives at a specific point and its corresporing value of $C$, its velocity and acceleration vanish. The body then remains' eternally at this point, hence, the specific points are the positions of relative equilibrium of point $P$. In these points the body can remain at rest relative to the moving coordinate system. Wher body $P$ is at a specific point, the ratio of the distances between the three bodies $S, J$ and $P$ remains unchanged. We thus conclude that the specific points are nothing else but the libration points, which we have studied in the previous chapter.

Let us now find the coordinates of the libration points and evaluate the corresponding values of the Jacobi constant. The last of equations (14) yields $z=0$ so that the libration points lay in the plane $x 0 y$.

They may be identified with the specific points of curve (7). We shall make use of this situation in the practical evaluation of the coordinates of the libration points. We first observe that when $z=0$.

This is because the origin of coordinates is taken in the cantre of mass, and hence

$$
-m_{1} a_{1} ;-n_{1} a_{2}=0 .
$$

Equation (7) can then be written as

$$
\begin{equation*}
m_{1}\left(r_{1}^{2} \mid \because r_{1}^{-1}\right)-: m_{;}\left(r_{2}^{2} ; \because r_{3}^{-1}\right)=C^{\prime} . \tag{1j}
\end{equation*}
$$

where

$$
C^{\prime}==C_{i}+m_{1} a_{i}+f \cdot m_{i} u_{i}^{\prime}=C+\begin{gather*}
m_{1} m_{2}  \tag{16}\\
m_{1} m_{2}
\end{gather*} .
$$

since, evidently,

$$
\begin{gathered}
a_{1}=m_{2} \\
m_{1}: m_{2}
\end{gathered} \quad u_{i} \quad \begin{gathered}
m_{1} \\
m_{1}+m_{2}
\end{gathered}
$$

Writing the equation of curve (15) in the form

$$
f(x, y) \quad 0
$$

we obtain the following equations for the specific points of curve (7)

$$
\begin{array}{llll}
1 f & 0 . & d / & 0 . \\
d x & 1 y & d
\end{array}
$$

In the present case, these points may be represented by

$$
\left.\begin{array}{cc:cc}
d \prime & d r_{1} & d / & d r_{2}  \tag{17}\\
d r_{1} & d & d r_{2} & d x \\
d o & 0 \\
d r_{1} & d & d r_{3} & \\
d, & d & d r_{2} d y & 1
\end{array} \right\rvert\,
$$

They may be satisfied in two ways; either to put

$$
\begin{array}{cccc}
\text { af } & 11 \\
i r_{1} & \text { ir } & 11
\end{array}
$$

which yields
ORIGINAL PAGT:
OF POC?
or to put

$$
\begin{gathered}
\prime \cdot ; \quad, \quad, \quad a_{2}=1 . \\
-1(1, b
\end{gathered}=\frac{1}{r_{1} r} \quad . \quad . \quad .
$$

frcm which it follows that $y=0$. In the first case, we obtain two Jibration points, $\mathrm{L}_{4}$ and $\mathrm{L}_{5}$ (figure 3), which form equilateral triangles wich points $S$ and $J$. Thus, points $L_{4}$ and $L_{5}$ are isolated points of curve (15). In other words, they are double points of complex tangents. The corresponding value $C=C_{4}$ are easily obtained from equations (15) and (16) as

$$
\because \quad 1 \mid \cdots \quad m=1 \quad \cdots \cdot \cdots
$$

In the second case, in which the double points are subject to condition (18) and lay on the axis $0 x$, one of the following conditions holds

depending on the situation of the duble point relative to S and J . We shall successively consider each of these cases.

The first case
Let $r_{1}+r_{2}=1$. Then

$$
\begin{array}{ccccc}
r_{1}: l & : u_{i}, & i & & 1 \\
\therefore r_{1} & 1, & י r & ! & \\
r & 1, & \therefore & ! &
\end{array}
$$

so that the first of equations (14) yields

$$
\begin{array}{ccccc}
m: & r_{1} & r_{1} & \quad \vdots i & \therefore r  \tag{:}\\
m_{1} & r & r & 11 & r 11 i \\
i
\end{array}
$$

or

$$
\begin{aligned}
& \text { ORIGINAL Agr:- }
\end{aligned}
$$

If the ratio $m_{1} / m_{2}$ is small, we obtain the required positive root in the form of a series-expansion. Actually, expanding the right-hand side of equation (20) in a power series, we obtain

$$
m_{i}^{m}=: m_{i}\left(1 \cdot r_{i}, j_{i}, \cdot \cdot\right)
$$

Taking the cubic root of each side, and introducing the notation

$$
V=\sqrt{\cdots}
$$

we obtain

$$
1 \vee 311 \vdots \frac{1}{i} \quad r^{r}:
$$

Hence,

$$
\therefore \quad, \quad \begin{gather*}
1 \\
i
\end{gather*}
$$

From equation (15), we obtain the corresponding values of the Jacobi constant;

$$
\begin{aligned}
& m_{1}(2!402-r!. .) \text {. }
\end{aligned}
$$

If we want to obtain a more accurate value than that which the series
(21) yields, we shall find it easier to numerically solve equation (20).

The second case

$$
\text { Let } \begin{aligned}
& r_{1}=1+r_{2} \text {. Then, } r_{1}=x+a_{1}, r_{2}=x-a_{2} \text { and } \\
& \text { (r } \quad \text {, }
\end{aligned}
$$

the first of equations (17) then yields

$$
\begin{array}{lccc:c:l}
m_{2} & r_{1}-r_{1}^{-i} & r_{3}^{\prime}(i) \cdot\left(r_{i}\right) \\
m_{1} & r_{2} & r_{2}^{\prime} & (1 \cdots \cdot!)(1 & \left.r_{2}\right)^{2}
\end{array}
$$

or

$$
\left(m_{1} ; m_{1}\right) r_{3}^{3}+\left(3 m_{1}+2 m_{i}\right) r_{2}^{\prime}:\left(3 m_{1}+m_{2}\right) r_{2}^{s} \cdots m_{0} r_{3}^{\prime} \quad 2 m_{2} r_{2}-m_{2}=0 .
$$

This equation is identical to equation (14) of $S 37$ when $m_{3}=0$ is substituted in the latter equation. Hence, it has one and only one positive root. Let us expand this root into a power series. Expanding the right-hand side of equation (23) in powers of $r_{2}$, we obtain

$$
v_{3}^{3}=r_{3}^{3}\left(1-r_{2}+y_{3}^{4}-\quad .\right)
$$

Raising both sides to the power 1/3 e obtain

$$
\left.\because r_{:} 11-\frac{1}{3} r_{:}: \frac{1}{3} r_{:}+\cdots .\right)
$$

Solving this equation, we obtain

$$
\begin{equation*}
r_{2} \cdot v: \frac{1}{3} v-\frac{1}{4} v+\ldots \tag{24}
\end{equation*}
$$

The corresponding value of the Jacobi constant is

$$
C_{2}-m_{1}+3: 1: 5-5 \cdot \left\lvert\, \cdot . \quad 1-\begin{gather*}
m_{1}, m  \tag{25}\\
m_{1}
\end{gather*}\right.
$$

The third case
Let $r_{1}=r_{2}=1$. Then, $r_{1}=-x-a_{1}, r_{2}=-x+a_{2}$ and

$$
\begin{array}{llll}
\prime \prime \prime & \text { n. } & \\
\cdots & -1, & \cdots & -1 .
\end{array}
$$

In analogy with the previous case, we obtain

$$
\begin{aligned}
& m_{:}=\begin{array}{rl}
r_{1}-r_{1}: \\
m_{1} & 1
\end{array} r_{:}:
\end{aligned}
$$

1.1'

Since $r_{2}>1$, then $r_{1}<1$. We assume that

We obtain the following equation for the quantity $\alpha$
or

Then,

$$
x \cdots\left(\begin{array}{l}
7 \\
1
\end{array}-32: \ldots\right)
$$

Solving this equation by the method of successive approximations, we obtain

$$
\begin{equation*}
a=-\frac{7}{4} v^{3}-21 v^{2} \tag{2i}
\end{equation*}
$$

The Jacobi constant will then be given by

In other words,

$$
C_{3}=m_{1}\left(3+12 v^{3}-\frac{3}{11}+-.\right)
$$

If we assume that the sum of the masses $m_{1}$ and $m_{2}$ equals unity and introduce the following notation

$$
m_{1}==1-!, \quad m_{2}=!
$$

then all the quantities under consideration will be functions of the variable $\mu$ only. It is sufficient for this variable to vary from 0 to $\frac{1}{2}$ in order to cover all of the possible cases ${ }^{(1)}$.
(1) The coordinates of the libration points as well as the corresponding values of the Jacobi constants, $\mathrm{C}_{1}(\mu), \mathrm{C}_{2}(\mu), \ldots$ are studied in the following paper
M.Martin, On the libration points of the restricted problem of three bodies, American journal of Mathematics, 53, 1931, 167-177.
Corrections and addenda to this paper are given in A.A. Markov, irogress in Astronomical Sciences (Uspehi Astronomiceskih nauk) 3, 1933, 75-77.
Tables of the abscissae of points $L_{1}, L_{2}$ and $L_{3}$ are given in the follcwing paper
J. Rosenthal, Tabl $s$ for the libration points of $t$ ? restricted problem of three bodies, Astr. Nachr. 244, 1931, 1969.

In figure 7, the curves that corraspond to the critical values of the Jacobi constant are shown for the ase

$$
m_{1}=10, \quad m_{2}=-1 .
$$

These critical values, as well as the bipolar coordinates of the libration points are given in the following table

For the case

$$
n_{1} \quad 1 . \quad \because \quad 1
$$

The corresponding values are

Curves obtained for these values of the jaccibi constants are shown in Figure 8.


PII. 1.

Figure 7

bic. s.

Figure 8
41. Periodic solutions of the restricled problem of three bodies

In the previous chapter, we studied the Lagrange motions in which bodies simultaneously move in eliiptic orbits. These motions are examples of periodic orbits in the three-body problem. In these motions, the coordinates of al". the three bodies are expressed by periodic functions of time having equal periods.

Hill gave another example of periodic rbits. He developed a method for the independent determination of some inequalities in the motion of the moon, caused by the gravitation $\cap \mathrm{f}$ the sun (Chapter XVIII). Later, Poincaré suggested a method for finding and studying the general classes of periodic solutions of the three-body problem. Periodic solutions are thus the first targets attained in the three body problem that have never been solved analytically. With the start of the periodic solutions, the study oi other interesting types of solutions, such as the asymptotic solution, became possible.

The study of periodic solutions is just in its initial stage. Even in the simple case of the restricted problem, only a few groups of periodic orbits have been more or less perfectly studied.

The most well-studied orbits are plane periodic orbits, passing close to the libration points $\mathrm{L}_{1}, \mathrm{~L}_{2}, \ldots . \mathrm{L}_{5}$. These are the orbits that inclose planet $J$ but not $S$, and the orbits that enclose the sun only at such a distance,...$a t$ the ratios of the period of revolutions along them to the period of revolution of planet J are simple, such as $1: 3,2.3$ and so on.

The orbits of the first type are useful in tr $\geq$ investigation of the motion of the so-called "Trojans". These are the small planets that move nearly along the Jupiter trajectory. The elements of the orbits of the known "Trojans" are shown in the next table. The elements
are given rela＇ive to the ecliptic and equinox 1925．0．The average longitude of epoch $\in$ is given for the moment 1925 January 10 T．U． The elements of Jupiter in that table are the average elements for the above mentioned moment．The elements of the small planets are the osculating elements of different epochs，grouped around 1935.

| y ${ }^{\text {a mamer }}$ | $a$ | $n$ |  | $i$ | ＇： | \％ | c |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
|  |  |  |  |  |  |  |  |
| 11 Juppiter | －2030 | 2．10\％ 124 | 277.8 | 1．31）＇ | ＇H20．9， | 1.123 | 271142 |
| iss Abillic， |  | 2かり尔 | 4511 | 10.319 | 31638 | 527（4） | 35，3\％ |
| 6．1；Butuilu， | 5111 | 201115 | 740 | $22 \times 1$ | $135 \times 1$ | 316，970 | 291011 |
| bet Hectur | －117．3 | 3u：s：4 | 15．2 | $1 \times 201$ | ． 116 L | 101．521 | ［1937 |
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| 4xt Pram＂ | ，272 | 295114 | 1．8：？ | 884. | $3 \times \mathrm{n}: 2 \cdot 1$ | 270 ${ }^{\prime \prime}$ | 11176 |
| 961 is．ancila．ton | $\therefore$ 圆々 | ：19512， | ． 2111 | 2113.1 | 3．3， 31 | 55383 | $\therefore 2$ |
| 11.3 （1）M mell | $\therefore 16,0$ |  | （1）！ | 31 | 29111 | 939\％ | 31007 |
| 1172 \110．． | ．2101 | 4975 ${ }^{16}$ |  | $11.17{ }^{\text {a }}$ | $\therefore 16117$ | 201180 | 20－15．3 |
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| 120）frombin | $\because 1$ | 2，10，1！ | 10．1 |  | 171.5 | is 214 | $\underline{1.750}$ |

The last column of this table indicates that the planets $588,621,659$ ， 911 and 1.143 are close the libration point $L_{4}$ ，while the others are around $\mathrm{L}_{5}$ ．

The theory of the second－type orbits，which enclose Jupiter $J$ at a short distance，are closely related to the theory of satellites．

The orbits of the third type make it possible to construct a theory for the motion of small planets，the average motion or which is commeasurable with the motion of Jupiter．It is often more useful to use these periolic orbits as a first approximation to the orbits of planets， rather than to use the Kepier ellipses．

Poincare divided the periodic elements of the restricted problem of three bodies into three glades．He related the orbits of bodies $S$ and $J$ that lay in the $x 0 y$ plane to the first and second grade，and those
of the other planets to the third one. The difference between pericilc elements of the first and second grade is the following. If the mass of body J tends to zero, the periodic orbirs will tend to Kepler ellipses. The orbits of the first grade will be those, for which the eccentricities of the limiting ellipses are zero. In other words, the orbits of the first grade are slightly different from circles when the value of $\mu$ is small. The second-grade orbits are near to elliptic orbits.

We shall not consider in detail the properties of periodiz orbits. We shall only consider infinitesimal orbits around 1ibration points in the following section ${ }^{(1)}$.

Alongisde the analytical methods of finding periodic solutions, the method of numerical integration of differential equations is applied by the initiative of Darwin and Tile. The numerical integration of the equations of motion has an advantage over the corresponding ar. tlytical methods. The former method is simpler than the latter when one considers a given concrete case. One is then able to obtain the numerical solution using th. $m$, , 1 ementary methods of calcuiation. However, this solution is only us: $\therefore$ - ro. the interval of ime at which the calculation were made. This is the most serious drawback of numerical solutions. Periodic solutions are evidently free from this deficiency. It is sufficient to obtain an analytical solution for one perioa in order to
(1) Apart frum the classical work:
H. Pr,incaré, Le methodes nouvelles de la Mechanique Celeste, t. I, II, III, Paris 1892-1899, the theory of periodic crbits is given in:
F.R. Moulten, Periudic Orbits, Washingtou, 1920. A detailed bibliography is g.ven in the article: E.T. Whittaker, Prinziplen der Storungsiheorie und allgeneine Theorie der Bahnkurven in dynamischen Problemen, Encyklopädie der Math. Wissenschaften, Bd. VI, 2 (1921) 512-556. obtain a complete picture of the motion that corresponds to the given initial conditions.

At present, analytical methods are applied for the study of only periodic orbits in the case when the mass of $J$ is consid $\begin{aligned} & \text { rably smaller }\end{aligned}$ than the mass of $S$. 0.1 the other hand, the numerical methods are easily used for the arbitrary ratios cf masses. Darwin used this method for the case when the mass of $J$ equal to one-tenth that of $S$ and he was able to find a number of periodic elements. E1is Strömgren, as well as Tile, studied the case when the masses of $J$ and $S$ were equal. Such studies wert started by Burrau in the year 1900. Since 1913, research on this was continued by Elis Strömgren, a scientist of the Copenhagen observatory. These investigations gave the possibility not only to ind a large number of periodic elements, but also to follow the transition of some classes of these orbits into others and to observe the disappearing process of some classes of periodic orbits when the initial condit*ons are changed. These observations naturally led to a consinerable s plification of the analytical solutions corresponding $t$. $t_{1}$. oocesses ${ }^{(1)}$.
(1) The resultr of Darwin are given in his classical work: G. Darwin, Periodic Orbits, Acta Math., 21, 1897, 99-216. Some additions are given in Math. Ann., 51, 1399. The results obtained in the Copenhagen observatory are given in a series of memoirs: Publikationer og nindre Middelelser tra Kobenhavns Observatorium. The conclusion are given in Elis Stromgren's Paper "Connaissance actuelle des orbites dans le probleme des trois corps", wlich is published in No. 100 (1936) of this Journal. This paper contains the full kibliography of the work of the Conenhagen School and is also published in: Bull. astr., 2-e serie, 9, 1936.

## 42. Motion near collinear libration points

Let $(a, b, c)$ be an arbitrary point of the uniformly rotating space Sxyz. We investigate whether it is possible that among the motions, defined by

$$
x-y_{n} y=\frac{d u}{d x}
$$

where
there exists such a motion, that body $P$ is always as close as possible to point ( $a, b, c$ ).

We can consider that the function $\Omega$ is holomorphic in the $v:$. inity of point $(a, b, c)$, except in the case when this point coincides with one of bodies $S$ and J. Accordingly
we expand the right-hand side of the equations of motion in powers of the small nuantities $\overline{5}, \eta$ and $\zeta$. Keeping only the first powers of these quantities, we obtain

$$
\begin{aligned}
& \begin{aligned}
\square \\
\square
\end{aligned}
\end{aligned}
$$

When $\xi$ and $\zeta$ are sufficiently small, the motion takes place in the vicinity of the points, the coorinates of which are given by

$$
\frac{4}{d u}=\frac{0}{d y}-\frac{10}{4}
$$

f.e., in the vicinities of the libration points, since these equations are identical with equations (14).

In this section, we consider the case of motions, proceeding infinitely close to collinear libration points. Hence, we put

$$
a=x_{1}, \quad \forall \quad 11, \quad \subset \quad 11 .
$$

where $X_{k}$ denotes the abscissa of the libration point $L_{K}(K=1,2,3)$. In order to find the second derivatives of function $\Omega$, involved in equations (30), we differentiate expression (29) and replace the variables $x, y$ and $z$ by the above mentioned coordinates of the libration point $\mathrm{L}_{\mathrm{K}}$. We then obtain
where

$$
r_{1}-x_{4} i u_{1} \cdot r=x_{k} \cdots c_{2}
$$

Introducing the following notations

$$
\begin{array}{ccc}
1 \\
1 & 1 \\
1 & 1
\end{array}
$$

we write equation (30) as follows

The last of these equations is independent of the others. If immediately gives

$$
\begin{equation*}
\therefore=c_{1} \sin , d_{2} t-f-c_{z} \cos A_{2} t \tag{.3.}
\end{equation*}
$$

where $C_{1}$ and $C_{2}$ are constants of integration. We search for the solutions of the first two equations in the form

$$
\because-G e^{A t}, \quad \mathrm{~T}_{1} \quad H e^{\prime \prime} \text {. }
$$

Substituting these expressions in equations (31) yields the following relations between the parameters $G, H$ and

$$
\begin{align*}
& \mu^{\prime \prime}-1 n:+2 A, 1(j-2 n: 1 H 0 \\
& \because n=i(i-i-1 ;-1 n \quad A, \| / 1=0.1 \tag{3i3}
\end{align*}
$$

we denote by $\lambda_{1}, \lambda_{2}, \lambda_{3}$ and $\lambda_{4}: \ldots, \cdots$ characteristic equaizior

$$
\left.\left|a:-\left(n^{2}+\quad \therefore A_{2}\right)\right|\right|_{i}=\left(n:-l_{2}\right) \mid+i i^{2}=0
$$

or

$$
\begin{equation*}
\therefore 1\left(1 n^{1}-A_{4}-2\right) i:\left(n:-A_{1}\right)\left(n^{:}:-2 A_{1}\right)=-0 \tag{31}
\end{equation*}
$$

and by $q_{1}, q_{2}, q_{3}$ and $q_{4}$ the corresponding values of the ratio $H: G$, defined by equations (33). The following equations
where $G_{1}, G_{2}, G_{3}$ and $G_{4}$ are arbitrary constants, together with equation (34) define the general solution of system (31).

Evidently, the nature of the body $P$, thai has an infinitesiamal mass, depends on the type of roots of equation (34). One easily sees that thic equation has two real and two imaginary roots for the libration
points $\mathrm{L}_{1}, \mathrm{~L}_{2}$ and $\mathrm{L}_{3}$. Indeed, we shall show that the quantity

$$
n-d_{1} \quad m_{1}: \cdot m_{1} \cdots m_{1}-\frac{n_{2}}{r_{1}} \quad \text { (n, }
$$

is negative for $k=1,2,3$. Consequently, there will be two real roots of opposite signs for equation (34), being considered as a second-order equation with respect to

For poin $L_{1}$, expression (36) is negative because, in this case,

$$
r_{1}<1, r_{2}<1
$$

For points $L_{2}$ and $L_{3}$, equations (23) and (26) give

$$
\begin{array}{lll}
\cdots & - \\
\cdots & \bar{r} & \cdots
\end{array}
$$

Eliminating m from equation (36), we obtain

$$
\left.\therefore \quad \therefore \quad \because\left(\begin{array}{ll}
1 & 1 \\
& \ddots
\end{array}\right) \right\rvert\, l \begin{array}{ll}
1 & \ddots \\
\vdots
\end{array}
$$

Since, for point $L_{2}$,
and, for point $L_{3}$,

$$
r_{1} \cdot 1, r: 1
$$

expression (36) is evidently negative. Hence, two roots of the characteristic equation are purely imaginary complex-conjugate quantities. The other two roots are real and have opposite signs. Accordingly, the libration points $L_{1}, L_{2}$ and $L_{3}$ are positions of unstable relative equilibrium. In other words, when body $P$ is disp!aced from any of these points by an arbitrary small distance with an arbitrary small velocity, it may leave for ever the vicinfty of this libration point.

Let us denote the real roots of equation (34) by $\lambda_{3}$ and $\lambda_{4}$. If we choose the initial conditions so that $G_{3}=C_{4}=0$, we obtain a motion, in which body $P$ will remain forever in the vicinity of the
corresponding 1 ibration point. This is because the coordinates $\xi, \mathcal{Y}$ and $\zeta$ relative to this point will be bounded for all values of $t$.

The existance of orbits, arbitrarily close to the libration points $L_{1}, L_{2}$ and $L_{3}$ is a necessary but not sufficient condition for the existance of periodic orbits in the vicinity of these points. If we limit ourselves to the accuracy, that the approximate integration of equation (31) achieves, we easily obtain periodic orbits by the appropriate choice of the inftial values of coordinates, $\varepsilon_{0}, \eta_{0}$ and $\zeta_{0}$, and components of velocity, $\dot{\varepsilon}_{0}, \dot{\xi}_{0}$ and $\zeta_{0}$ We first set $\zeta_{0}$ and $\dot{\zeta}_{0}$ equal to zero. In equation (32) we will have $C_{1}=C_{2}=0$, i.e. the motion of body $P$ is planer. We then choose $\dot{\xi}_{0}$ and $\dot{\eta}_{0}$ using conditions $G_{3}=0$ and $G_{4}=0$. Formulae (35) will subsequently yield
where $\lambda_{1}=\beta_{1}, \lambda_{2}=-\beta$ and $\beta$ is a real number. On the basis of equation (33),

$$
\begin{array}{llll}
4 & 8 & \cdots & -4
\end{array}
$$

where

$$
\begin{gather*}
\dot{r}: n^{\bullet}:-1 .  \tag{3i}\\
2 n \cdot \beta
\end{gather*}
$$

Expressing the exponential functions in terms of trigonometric functions, we obtain
where

We solve these equations for $\cos \beta t$ and $\sin \beta t$, theil square the resulting expressions and add. We obtain the following equation for the trajectory

$$
\text { it, . U } 1:, . . \quad 11,1+1
$$

or

$$
\therefore \quad \therefore \quad 411011-1 .
$$

That is, the motion proceeds by an ellipse, the axes of which coincide with the coordinate exes. Denoting the semiaxes in directions $L_{k} \xi$ and $\left.L_{k}\right\}$ respectively, by $a$ and $b$, we obtain

$$
\begin{aligned}
& b \\
& a
\end{aligned}
$$

It is easy to show that $q>0$ for all three libration points. Hence, the eccentricity of the elliptic orbits obtained is equal to

$$
\rho .1 \because 1
$$

Consequently, the form of these orbits does not depend on the initial position $\left(\xi_{0}, \eta_{0}\right)$ of point $P$, which affects only the dimensions of the orbits.

The values of $q$ and $e$ that currespond to three values of the ratio of masses of $S$ and $J$ are given in the following table. The first two ratios are relative to the work of stromgren and Darwin, discussed in 41 , the third value takes place in the earth-sun system.


We have thus obtained three systems of infinitesimal periodic orbits, each of which depends on two parameters. Naturally, the existance of these oriots is in sufficient to prove the existance of finite periodic
orbits near the libration points $L_{1}, L_{2}$ and $L_{3}$. However, it is possible to show that such periodic orbits actually exist ${ }^{(1)}$.

In conclusion, we mention some words on the work by Hulden (Goulden) and Moulton, in which they applied the theory developed here to explain the antiaurora effect. They assumed that the libration point $L_{2}$ for the earth may be taken as the centre of an accumulation of meteors, occupying the interplanet space. Indeed, meteors for which $G_{3}=G_{4}=0$ always remain in the vicinity of this point. Those, for which $G_{3}$ and $G_{4}$ are small, remain for a long time near this libration point. The light of the sun is reflected by the cluster of such meteors, we thus observe the reflected light as an antiaurora.

The distance of point $L_{2}$ from the earth is equal to 0.0101 astronomic units ( § 37), i.e. about 1.490 .000 kms . If the above mentioned assumption is correct, the antiaurora will have a parallax of the order of 15'. Unfortunately, the antiaurora is of such a diffused effect, that there is no hope to check the validity of this assumption by observing its parallax.
43. Motion Near Triangular Libration Points

We will now consider the motion of body $P$ near the libration points $\mathrm{L}_{2}$ and $\mathrm{L}_{2}$. We adopt that

$$
-1 \mu, m, \mu
$$

where $\mathrm{n}^{2}=\mathrm{m}_{1}+\mathrm{m}_{2}=1$, and assume that

$$
\begin{array}{lll}
1 & 1 & 1 \\
2
\end{array}
$$

The coordinates of bodies $S$ and $J$ are then equal to
(1) F.R. Muviton, Periodic Orbits, Ch. V; and references cited therein.

$$
-a_{1}=-i r_{1} \quad a_{2} \cdot 1-1_{1}
$$

Hence, the coordinates of point $\mathrm{L}_{4}$ are

$$
a=\frac{1}{2}\left(1-2(3), \quad b \quad \begin{array}{c}
\sqrt{3} \\
2
\end{array}, c-=(1 .\right.
$$

The coordinates of point $L_{5}$ are obtained from the coordinates of point $L_{4}$ by changing the sign of $\sqrt{3}$. We can thus study the motion near points $\mathrm{L}_{4}$ and then obtain the corresponding result for point $\mathrm{L}_{5}$ by changing this sign.

Differentiating the function $\Omega$, and replacing $x, y$ and $z$ by the coordinates of point $\mathrm{L}_{4}$, obtained above, we obtain the equations of motion (30) in the form

$$
\begin{aligned}
& \therefore-\text {-. }
\end{aligned}
$$

The general integral of the last of these equations is

$$
\begin{equation*}
: \quad c_{1} \cos t \quad i=\sin t \tag{10}
\end{equation*}
$$

where $C_{1}$ and $C_{2}$ are arbitrary constants. This solution shows that the projection of point $P$ on the axis $0 z$ performs periodic vibrations about the projection of the libration point on the same axis. The period of this vibration equals $2 \pi$, ie. coincides with the period of rotation of the finite masses $S$ and $J$ around their centre of gravity. We should always remember that this, esuit is only valid for linear vibrations about the position of relative equilibrium, i.e. for such a motion, for which we neglect in equations (30) terms involving second and higher powers of $\xi, \eta$ and $\zeta$.

We now solve the first two equations of system (39). We can write these equations as follows

$$
\begin{equation*}
\vdots \quad \because \quad \therefore \quad \therefore-S_{1}: \quad 1 \quad \therefore \quad \therefore \quad T_{1} \tag{10}
\end{equation*}
$$

if we consider that

$$
I \quad a_{4}^{3} \quad 3 \quad i_{i}^{3}(1-\ldots \cdot 1, \quad 7 \quad 1
$$

Substituting
… Cie", y . He',
into equation (41), we obtain the follor ing equa. ons for the unknown constants

$$
\left.\begin{array}{ccc}
(2-R) O & (2 n+S) H & 0  \tag{42}\\
(2 i-S) O & -(i-2-T) H & =0
\end{array}\right\}
$$

These equations yield

Denoting by $\lambda_{1}, \lambda_{2}, \lambda_{3}$ and $\lambda_{4}$ the roots of the latter equation, we obtain the solution of system (41) in the form
where $G_{1}, G_{2}, G_{3}$ and $G_{4}$ may be considered as arbitrary constants. The corresponding quantities $\mathrm{H}_{1}, \mathrm{H}_{2}, \mathrm{H}_{3}$ and $\mathrm{H}_{4}$ are obtained from equations (42).

The nature of the motion, represented by formulae (44), essentially depends on the type of roots of equation (43). These roots are given by the following equations

$$
\left.\begin{array}{ccccc:c}
i_{1} & -i_{2} & \sqrt{1} & \vdots & - & \sqrt{4} \\
i_{3} & =-i_{1} & \sqrt{\prime} & ! & ! & i n
\end{array} \right\rvert\,
$$

where

$$
M==1-27 \mu(1-\mu) .
$$

Increasing the value of $\mu$ from 0 to $\frac{1}{2}$, we find that $M$ is positive in the beginning, then vanishes at $\mu=\mu_{0}$, where

$$
\mu_{0} \cdot 0.0385 .5108110 \cdot=\frac{1}{2 i}
$$

and, subsequently, M. remains negative up to $\mu=\frac{1}{2}$. Hence, we conclude that the values of $\mu$ are in the interval

$$
n \cdot \mu \cdot a_{1,1}
$$

Equation (43) has different purely imaginary roots. Consequently, the general solution (44) may be written as
where

$$
11, \quad+
$$

and $G^{\prime}, G^{\prime \prime}, G^{\prime \prime \prime}$ and $G^{\prime \prime \prime \prime}$ are new arbitrary constants. The coefficients $H^{\prime}, H^{\prime \prime}, \ldots$ are expressed in terms of $G^{\prime}, G^{\prime \prime}, \ldots$ by relations similar to those connecting $G_{1}, G_{2}, \ldots$ with $H_{1}, H_{2}, \ldots$

It is thus clear that, in the case under consideration, the libration points are positions of stable relatire equilibrium of body $P$, whose mass is infinitesimal. Actually, in order that the absolute values of $Z$, and ' $\eta$ remain less than an any given small quantity, it is necessary that the values of $\xi, \eta, \dot{\eta}$ and $\dot{\eta}$ are infinitesimal.

Whatever the initial conditions are the projection of the motion of body $P$ on the plare $\xi^{?} ?$ may be considered as a superposition of two elliptic motions, defined by
and

$$
\begin{gather*}
\vdots=U^{\prime} \cos \beta t+U^{\prime \prime} \sin \beta t, \quad r_{1} \quad t^{\prime} \cos H^{\prime} t-H t^{\prime \prime} \sin H^{\prime} t  \tag{-1i}\\
\vdots=G^{\prime \prime \prime} \cos \gamma t+G^{\prime \prime \prime} \sin \gamma t, \quad r_{1}==H^{\prime \prime \prime} \cos \gamma t-+H^{\prime \prime \prime} \sin \gamma t \tag{177}
\end{gather*}
$$

If we choose the initial conditions in such a way, that the constants $C_{1}$ and $C_{2}$ involved in formula (40) vanish, and that either $G^{\prime \prime \prime}=G^{\prime \prime \prime \prime}=0$ or $G^{\prime}-G^{\prime \prime}=0$, we obtain periodic orbits $h=v i n g$ pe iod equal to $2 \pi / \beta$ in the first case and the $2 \pi / \delta$ in the second one. Each of these periodic orbits depends on two arbitrary constants.

We have considered the case when $0<\mu<\mu_{0}$. If $\mu>\mu_{0}$ then $M<0$ and all of the roots $\lambda, \ldots$ are complex numbers having. . nonvanishing real parts. Hence, the general solution (44) becomes

$$
\begin{aligned}
& \because=e^{\prime t}\left(G^{\prime} \cos \beta t+C^{\prime \prime} \sin \beta t\right)+e^{-t}\left(i^{\prime \prime \prime} \cos \gamma t+G^{\prime \prime \prime} \sin \gamma t\right) \\
& \gamma_{l}=e^{\prime \prime}\left(H^{\prime} \cos \beta t+H^{\prime \prime} \sin \beta t\right)+e^{-t}\left(/ t^{\prime \prime \prime} \cos \gamma t \cdot H^{\prime \prime \prime} \sin \gamma t\right),
\end{aligned}
$$

where $\sigma, \tau, \beta$ and $\gamma$ are nonvanishing real numbers. In this case, $\mathrm{L}_{4}$ and $\mathrm{L}_{5}$ are positions of unstable relative equilibrium of body $P$.

The intermediate case of $\mu=\mu_{0}$ will be considered in the following saction.

## 44. Application of Normal Coordinates

We have expanded the motion that proceeds infinitely close to the libration points $\mathrm{L}_{4}$ and $\mathrm{L}_{5}$ into the elleptic motions, described by equations (46) and (46). These ellittic motions proceed along orbits, the axes of which are inclined to the coordinate axes. In order' to simplify the study of these orbits, we transform the equations of motion (41) into a form, similar to that of thefirst two equations of (41). Equation (41) may be rewritten as

$$
\because-2 \eta \quad \begin{array}{ll}
! & \partial F \\
! & d F
\end{array}, \quad i+2 \vdots=\begin{array}{ll}
1 & d F \\
2 & d x_{1}
\end{array}
$$

where

$$
F=k s^{2}+2 \cdot 2 s \xi \eta+T \eta^{2} .
$$

These equations have the following integral

Thus, the curves of zero velocity, given by the following equation

$$
\therefore \quad \therefore 14: 14 \quad 1 . \quad \text { in: }
$$

are conic sections, the centres of which are located at the libration point $L_{4}$.

In order to transform equation (48) into the canonical form

$$
i_{1} ; \mu_{1}^{\prime \prime} \quad 1 .
$$

it is necessary to rotate the coordinate axes by any angle $\theta$, defined by

The new coordirates will be expressed in terms of the old ones by
where the coefficients $A$ and $B$ will be the roots of the secular equation, given by

$$
\begin{array}{llll}
K_{1} & \prime \prime & & S \\
S & 7 & (1) & =0
\end{array}
$$

or, in an unfolded form,

$$
w-3 w+\frac{\square}{4}:(!-!)=0 .
$$

Consequent 1y,

$$
A=\begin{aligned}
& 3 \\
& 2
\end{aligned}-\frac{3}{2} \sqrt{1}-3 \mu\left(1-\mu \therefore \quad B=\frac{3}{2}+\frac{3}{2} 11-3 n(1-\mu) .\right.
$$

After the abnve transformations, equations (41) w: t become

$$
\left.\begin{gather*}
\xi_{1}-2 r_{1}=\begin{array}{ll}
1 & \partial F \\
2 & \partial \sigma_{1} \\
\ddot{r}_{1}+2 \dot{r}_{1}= & =A \xi_{1} \\
1 & \partial F \\
\underline{2} & \partial r_{11}
\end{array}=B r_{1} . \tag{i9}
\end{gather*} \right\rvert\,
$$

Adopting again that

$$
E_{1}=E e^{i}, \quad y_{1}=H e^{2 t},
$$

we obtain

$$
\begin{aligned}
& \left(\lambda^{2}-A\right) E-2 \lambda F=0 \\
& 2 \lambda E+\left(\lambda^{2}-B\right) F=0 .
\end{aligned}
$$

We thus obtain for $\lambda$, the previous values given by equations (45),
while the corresponding ratios $F: G$ will be different.

$$
\text { For } \lambda_{1}=\beta_{i} \text { and } \lambda_{2}=-\beta_{i, \text { we find }}
$$

$$
F=x p t: \text { and } F:=p / i,
$$

where
similarly, for $\lambda_{3}=\gamma_{i}$ and $\lambda_{4}=-\gamma_{i}$, we obtain

$$
i \quad t 11 i r \quad u t
$$

where

$$
\because \quad i \cdot \begin{gathered}
.1 \\
\therefore \\
i
\end{gathered}
$$

Thence, equations (46) and (47) are replaced by


The first pair of formulae represent a motion along the ellipse

$$
p^{2}:=-r_{1}^{\prime}=p^{2}\left(E^{\prime} \div-j \cdot E^{\prime \prime}=1 ;\right.
$$

while the equation of the trajectory of the motion, represented by the second pair of formulae, is given by

$$
q^{3}=\ddot{1} \cdot+r_{11}^{\prime \prime}-q^{2}\left(E^{\prime \prime \prime} 2+E^{\prime \prime \prime \prime}=2\right) .
$$

It is clear that the ratios of the semjaxes of these ellipses are $p$ and $q$ respectively. Their eccentiricities do not depend on $E^{\prime}, E^{\prime \prime}, \ldots, 1 . e$. on the initial conditions.

The following table includes the values that characterize the motion near the libration points $L_{4}$ and $L_{5}$.


In figure 9 , the elliptic orbits for the case $\mu=0.01$ are shown In a strongly magnified form. The table shows that the smaller $\mu$ is, the more complete is the coincidence of the semimajor axes $0^{-}$the ellipses under consideration with the tangent of the ellipse, along which point $J$ rotates around $S$.

Let us now consider the case, when

$$
\mu=\mu_{9}=0.038 \text { S. }
$$

and, consequently,

$$
\begin{gathered}
\lambda_{1}=\lambda_{3}==_{-i}^{i}, \quad i_{2}^{\prime} \quad \lambda_{2}=\lambda_{1}=-\frac{i}{V 2}, \\
\prime=-28^{\prime} 5 y^{\prime} 10^{\prime \prime}, \quad A=0.085786+4, \quad B=2.91121351 i .
\end{gathered}
$$

Apart from the motion along the ellipse represented by equations (50) or (51), which will now be considered identical, we shall have a particular solution of che type

involving the arbitrary constants $K$ and $t_{0}$. Accordingly, the motion becomes unstable when $\mu$ becomes equal to $\mu_{0}$.

## 45. Tisseran's criterion

In conclusion of this chapter, we will consider one of the applications of the Jacobi integral, which has been suggested by Tisseran. It is well known that the elements of a conet's orbit may be strongly violated when it passes near a planet. Hence, it is of ten difficult to identify two comets only by their elements. Moreover, the appearance and even the brightness of a comet strongly vary certainly, one can evaluate the perturbation of one of the two comets under considerations from the time 1f appears until the time the other comet appears. However, this calculation is quite cumersome. It is only worthwhile doing this calculation if the chances for the accessful identification of a comet is good.

The orbital elements of a comet changes strongly enough to violate the orbit only if the comet approaches very closely a planet. For a planet

38 heavy and as far from the sun as Jupiter, a comet passing at a Sistance of 0.3 will be affected by the Jupiter's gravitation more than by the gravitation of the sun (S $\mathbf{S}$ ). The changes in the orbit, that will occur in the short time in which the comet is in contact with Jupitar, are stronger than the perturbations induced by other planets. As a first approximation, the perturbation induced by other planets are neglected so that the case under consideration may be regarded as a restricted three body problem. In addition, the eccentricity of Jupiter is small and since the interaction of Jupiter with the comet lasts only for a short time, Jupiter's orbit would only slightly deviate from its circuit. Keeping the notation of $S 38$, we see that the coordinates $x, y$ and $z$ of the comet satisfy equation (4), i.e.

$$
r+1+1=r_{1}+1+1+\cdots \prime\left(\begin{array}{cc}
n_{1} & 1  \tag{52}\\
1 & \vdots
\end{array}\right)
$$



- 186 -

This relation leads to the following necessary condition for the identity of two comets: Two comets will appear to us identical if and only if they have the same value of the Jacobi constant $C$.

In order to make use of this condition, it is necessary to evaluate for each comet cine relative coordinates $x, y$ and $z$ and components of velocity $x, y$ and $z$. Then, equation (52) gives the corresponding value of $C$ for each comet.

In order to simplify the application of this criterion, we make a transition into the fixed heliscentric coordinat e system $S \xi \%$ in which the $\mathrm{S}_{5}$ axis is parallel to the Oz axis. Evaluating the time starting from the moment when the axes $S x$ and 0 coincide, we obtain

$$
\begin{aligned}
& \prime \cdot \sin u+\eta \\
& y=-\sin t \quad u
\end{aligned}
$$

From this it follows that

In the new coordinate system, equation (52) becomes

This formula can be used to calculate $C$ when the comet is so far from the perturbing planet, that the comet moves almost entirely under the influence of the gravitation of the sun. In this situation, formula (53) can be considerably simplified. Let us denote by $a, e, 1, \ldots$ the elements of the comet in its motion around the sun. Taking the mass of the sun as unity, we put $m_{1}=1$. Then, the integral of area and the
integral of the kinetic energy of the two-body problem enables us to write the following equations

$$
\begin{gathered}
\vdots i-n: \quad \text { kjincosi } \\
\because: 1: \because \quad l:(\because-i)
\end{gathered}
$$

where $r=r_{1}$ is che radius vector of the comet. Hence, equation (5:3) may be replaced by

$$
u^{-1}:=i^{\prime} k^{-1} \sqrt{i} \text { coli } \quad i_{i}, i \cdot
$$

where the average durnal motion of Jupiter $n$ is replaced by $n$ ' in order to be distinguished from the elements of the comet. Moresver, the fullowing notations have been used

$$
\begin{aligned}
& \therefore=10
\end{aligned}
$$

In the cases that are usually met with in practice, the coordinates $\bar{\xi}$ and $\eta$ are not large. Since $a_{1}$ and $m_{2}$ are small quantities (of the order of 0.001 ), and

$$
18 A^{\prime}=0.1189 .011^{\circ}
$$

then $\delta$ may be dropped.
We sce that, after the comet has passed out of Jupiter's sphere of action, the expression

$$
i^{-1} \quad \because_{1}^{-1} 1: i^{-1} \quad 1 \quad 1 \quad 1
$$

conserves its value. This equation represents Tisseran's criterion that defines the necessary (but not sufficient) condition for the identification of two comets.

Relation (54) becomes more accurate if on the one hand we take into consideration the corrective term $\mathcal{S}$, and on the other hand replace the
average durnal motion of Jupiter rby the angular heliocentric velocity.

Neglecting the square of the eccentricity of Jupiter, we obtain

$$
\begin{array}{cc}
u & \langle 1 ; \\
\cdots i & \vdots \cdot
\end{array}
$$

Hence, the more exact relation $w 1 l l$ be given by

The radius-vector of Jupiter is best of all evaluated at the moment when the comet approaches Jupiter.

As an illustration of the present theory, we shall consider the approach of the Wolf comet to Jupiter in 1922 and which has been studied by M. Kan1enski ${ }^{(1)}$.

The osculating elements of this comet before entering in Jupiter's sphere of action and after leaving it are

1.. $\because$ Decumben, 0
$\therefore \therefore 11 \therefore 1$
$\therefore 1 \cdot 17$
4.: I:"

$\left.\therefore \begin{array}{lll}\therefore 1 & 9 & 1 \\ 1 & 1 & 10\end{array}\right\} \quad 192.0,11$
い11010
(1) :
" ' 1101

The variations in the elements are strong because the minimal distance between the comet and Jupiter reached the value $\Delta=0.1247$ on
(1) The numbers given here are taken from the work:
M. Kamlenski, Recherches sur le mouvement de la comete periodique de Wolf, Bulletin de 1'Academie' Polonaise des Science et des Lettres, Serie, A, 1925.

September 27, 1922.
Adopting that $\mathbf{C a}^{\prime}=0.74624$ and $\mathrm{C}_{\mathrm{g}} \mathrm{r}^{\prime}=0.73604$ and neglecting in equation (55) the small quantity $\delta$, we obtain for the two comets the following values

$$
\therefore \quad 0.1 \because=10 \mathrm{md} \text { di } 141 \because
$$

We thus see how small the change in the quantity $C_{0}$ is, even during so large a variations of the elements.

## PART THO

TRE NUMERICAL INTEGRATION 9F DIFFERENTIAL EQUATIONS
AND ITS APPLICATION TO THE STUDY OF THE MOTION OF STARS

## CRAPTER VIII

THE NUMERICAL INTEGRATION OF DIFFERENTIAL EQJJATIONS
46. Introduction

All the problems of celestial mechanics are reduced to the solution of some differential equations. That is why celestial mechanics is always inseparably linked with the development of the methods for the solution of differential equations.

The integration of differential equations in a closed form is only possible in the most simple cases, such as thetwo-body problem. In general, the solution caunot be obtained in terms of the well-known functions. One then has to try other methods for the solution of differential equations. Amongst these methods, the two most general and effective methods are (1) the method of integration by series of expansions and (2) the method of numerical integration. In this chapter, we study in detail the numerical integration of differential equations.

The first successful application of the numerical method was given by Clero (1813-1765) in a study of the perturbation of Halley's comet. His method was later developed by Dalamber, Euler and in particular by Laplace. The final stage of this method was achieved by Gauss ${ }^{(1)}$
(1) C.E. Gauss, Exposition d'une nouvelle methode de calculer les perturbations planetaires (Nachlass), Werke, 7, 1900, 439-472 Gauss' formulae were published for the first time by Encke (J.E.Encke, Uher mechanische Quadratur, .erliner Astr. Jahoclaich fur 1837, Berlin 1835 , and published again in Gesammelte mathematische und astronomische Abhandlungen, Berlin 1888, 21 60). The application of the so-called "mechanics" of quadratures is not only out of date, but also may cause a lot of misunderstanding.

تitc suggested the sō-calied meinod or quadratures.
The method of quadratures was developed for its use in the solution of the particular problem of evaluating the perturbations of comets and small planets. This explains why the method of quadratures was not always accepted as a general method. It was considered as a particular way of evaluating perturbations, l.e. calculating small corrections to an already known approximate solution.

In 1908; the eighth satellite of Jupiter was discovered. The motion of this satellite could not be interpretted by Kepler's law. It was then necessary to investigate the general character of the corresponding dynamical problem. Cowell suggested that "mechanical quadratures" should be rejected. He proposed a new method for the general integration of the differential equations involved. This method was the origin of the method of quadratures. A great deal of attention was paid to this method, especially after it had succeeded in predicting the return of Halley's comet in $1910^{(1)}$. When the work on the motion of Halley's comet was over, Cowell mate an important conclusion on the basis of his wide experience on the numerical integration of differential equations ${ }^{(2)}$. This conclusion was that the Cowell's method can be signiftcantly improved. Althnugh this conclusion was theoretically evident, it remained unnoticed for a long time. When Cowell's method
(1) P.H. Cowell and A.D. Crommelin, The Orbit of Jupiter's Eighth Satellite. Monthly Notes, 68, 577-581. P.H. Cowell and A.D. Srommelin, Essay on the return of Halley's Comet, Pablikation der Astr. Gesellochaft, 23, 1910.
(2) Appendix to the Volume of Greenwich Observations for the 1909, 81.
was improved, it became identical in form with the method of quadratures
that Gauss had suggested.
In the past decade, the method of numerical integration of differential equations has been widely applied in fields other than celestical mechanics. In ihis connection, Adams, Stormer, Rugge and others, suggested some other methods. These methods are not as perfect as the method of quadratures. Hence, they were not widely used in celestial mechanics ${ }^{(1)}$. However, we shall not only consider here the method of quadratures, but also the other methods. This will give us an idea on the advantages of the method of quadratures.

## 47. Evaluation of derivatives in terms of differences

We shall consider the values of each function that correspond to the values of the independent variable $t$ that form an arithmetic progression, i.e.

The values of a given function, say $f(t)$, will be denoted by

$$
I_{k}=f\left(I_{1}\right) .
$$

where

$$
t_{k}=A_{1}+t_{1}(i) .
$$

(1) These methods, which are of interest to engineers, are give:i in: A.N. Kryllov, Lectures on approximate cal~ilations (Lekcil o priblizennyh vycislenijah) 3d. edition, 1935.

The following saheme shows the syster of

OF POOR PAGE LS QUALITY

notations which we shall subsequently use for the differences. Accordingly

$$
\begin{aligned}
& f_{k+1}^{\prime}=f_{k+1}-f_{k} \\
& f_{k}^{2}=f_{k+\frac{1}{2}}^{\prime}-f_{k}^{\prime}-\frac{1}{2}
\end{aligned}
$$

for arbitxary integral values of $k$. One of the values of the colunn of the sums of the 1 st order may be chosen arbitra:ily. The other values may be obtained using the following relation

$$
\begin{equation*}
f_{i}^{\prime}, f_{k}^{\prime}, 1-f . \tag{1}
\end{equation*}
$$

Similariy, assuming that one of the values of the second-order sums, say $f_{o}^{-2}$ is arbitrary we evaluate the other numbers in this column using the fol iowing equation

$$
\begin{equation*}
\because, l_{1}: f_{1}^{\prime}: \tag{?}
\end{equation*}
$$

aud making $k=\hat{0}, i, \ldots$ and $k=-i,-i, 3, \ldots$
The scheme of differences indicated above is often completed by the semisums of two neighbouring quantities of the same column. For these semisums, the following notation is used

$$
\begin{align*}
& f_{1}^{-1}=\frac{1}{2}\left(1^{-1}, f_{-}^{-1} \frac{1}{2}\right) \quad f_{1}^{-2}, \frac{1}{1^{\prime}}\left(f_{k}^{-2} \cdot f_{k}^{-2}, 1\right), . . \mid
\end{align*}
$$

Let us now try to express derivatives in terms of differences. We here use the stirling formula

Differentiating with respect to $z$ and putting $z=0$, we obtain

$$
\begin{aligned}
& \prime^{\prime \prime \prime}\binom{d ;}{d t_{j}}_{k}=I_{i} \quad, j \nmid \\
& u^{\prime} \cdot\binom{d \prime}{d l^{\prime \prime}}_{k}=f_{i}^{\prime}-\frac{1}{+} f_{k} \cdot \ldots-.
\end{aligned}
$$

Similariy, Bessel's formula
enables us to express the derivatives of the function $f(t)$ in terms of the differences given in the line $n=k+\frac{1}{2}$. We obtain

$$
\begin{aligned}
& \mu^{\prime \prime}\binom{d /}{d l^{4}}_{k}=f_{k}^{\prime}-\frac{1}{\because} f_{i l} \vdots . .
\end{aligned}
$$

The following formulae are used in the method of the numerical integration of equations, suggested by Adams and Stormer,

$$
\begin{aligned}
& w^{\prime \prime}\binom{d!}{d \prime}=f \ldots=\frac{11}{d} f_{i}!
\end{aligned}
$$

They express the derivatives of $f(t)$ in terms of the differences, located In the ascending diagonal. These relations are obtained by the similar interpolation of Newton's formula:

## Annotation

There are several other ways to denote the differences. The quantity $f_{n}^{1}$, where it is an intege: and $n$ a half fnteger (even or odd), will be denoted by $f^{i}(n)$, or $f_{i}(n)$ or by $f^{i}\left(t_{o} i n w\right)$, or finally by $\Delta \frac{1}{n}$.
48. Integration of First-0rier Equations. The Method ofRIGAdAIdPAGE IS

Let us consider the following differential equation

$$
\left.\left.\begin{array}{l}
d \\
d
\end{array}\right) / r, d\right)
$$

6

We want to calculate a tabie for the values of the function $X(t)$, that satisfies this equation as well as the initial condition

$$
x\left(t_{0}\right)=x_{0}
$$

where $t_{0}$ and $x_{0}$ are given numbers. Let us assume that $x_{0}\left(t_{o}+k w\right)=x_{k}$, and evaluate $x_{1}, x_{2}, x_{3}, \ldots$. . Our problem is to find the way to follow in order to calculate the differences

$$
i_{1}, x_{2} J_{k}, \quad \quad(k=0,1, \because, \quad . \quad i
$$

Since
then,

$$
د_{n} \quad 1==\begin{aligned}
& w \\
& 1!
\end{aligned}\binom{d n}{d!},+\begin{aligned}
& w^{i t} \\
& 2!
\end{aligned}\binom{i \quad x}{d!-}
$$

Adopting that

$$
w f(1,1) \cdots /(1), \quad w f(x, 1,) \cdots 1
$$

and taking equation (7) into consideration, we obtain

This formula essentially solves our problem. However, this formula cannot be easily applied ${ }^{(1)}$, since we have to calculate the derivatives
(1) The inethod of integration of equation (7), that is based on the use of formula (8) is called Euler's method. It is only applied in the rase when one can keep in equation (8) only two or three terms, i.e. when the interval $w$ is very small, or when one does not require an accurate solution.

$$
\binom{d f}{d t}_{t},\binom{d: \prime}{d t!}_{t} . \ldots .
$$

We now express the derivatives involved here in terms of differences. Using equation (6), we obtain

This formula represents Adams' method. Equations (5) leads to the following formula

$$
71
$$

which may be written as

$$
11
$$

since

$$
1, \quad 1,\left(f_{1}\right)=\frac{1}{\because}\left(f_{1}+1 \cdot f_{1}, 1\right) f_{2}, f_{1}
$$

This formula leads to the method of integration that may be called Cowell's method, since a similar method of integrating second-order equarions has been suggested by Cowell.

Once $x_{1}, x_{2}, \ldots, x_{k}$ are calculated equation (9) immediately gives $x_{k+1}$. On the other hand, Cowell's formula, given by equation (10), expresses the unknown difference $x_{k+1}-x_{k}$ in terms of the differences $f^{2}, f^{1}, \ldots$ which depend on $f_{k+1}, f_{k+2}, \ldots$ and consequently on $x_{k+1}, x_{k+2}, \ldots$ The difference $f_{k+2}^{2}, f_{k+\frac{1}{2}}^{2}, \ldots$ are found, in the first approximation, by extrapolation. After the evaluation of the corresponding values, $x_{k+1}, x_{k+2}, \ldots$, these differences

$$
\begin{aligned}
& \begin{array}{c}
-197 \\
3025 \\
\hline 2011
\end{array}
\end{aligned}
$$

are calculated in the usual way. If it is necessary, the quantities $x_{k+1}, x_{k+2}, \ldots$ are evaluated once more. Owing to the rapid decrease of the coefficients in equation (10), the above procedure converges so rapidly, that the second approximation might be unnecessary, provided that the interval $w$ is not large.

In the cases, which we are met with in celestial mechanics, namely in evaluating the perturbations of the elements; of orbits, the righthand side of equation (7) slowly varies with the variation of the variable X. When the variable $x$ slowly varies, the application of equation (10) becomes particularly simple because the values of the function $f$ and all its difforences may be evaluated in advance for several intervals using the approximate values of $x$.

Equations (9) and (10) can he used only when some of the initial values of the unknown function, $x_{1}, x_{2}, \ldots$, are given so that the evaluation of the differences $f^{1}, f^{2}, \ldots$, involved in these formulae is possible. The values $x_{1}, x_{2}, \ldots$ (and also $x_{-1}, x_{-2}, \ldots$ ) are usually evaluated by the expansion of the integral in a series, i.e.

$$
\cdots!0)=x_{0} ; x_{0}^{\prime}\left(t-t_{1}\right)+1{\underset{2!}{1} x_{0}^{\prime \prime}\left(1 \quad t_{0}\right)=+\ldots}_{l}^{l}
$$

The coefficients of expansion can be found by the multiple differentiation of equation (7) and subsequently the substitution $t=t_{0}$. Sometimes, the initial values $x_{1}, x_{-1}, x_{2}, x_{-2}, \ldots$ are found by a successive approximation ( $S$ 57). An example of this approach will be given in $S 55$. It is also possible to find some of the initial values, say $x_{1}, x_{2}, x_{3} \ldots$, using Euler's method.
49. The method of quadratures for the first-order equations:

The limitations of the method of differences, considered in the previous section, is the accumulation of errors when $x_{1}, x_{2}, \ldots$ are
evaluated in terms of their differences $\Delta$. Indeed, in order to evaluare $x_{n}$, we use the following equations

$$
x_{1}-x_{11}=\lambda_{\frac{1}{2}}, x_{2} \cdots x_{2}-j_{\frac{1}{2}} \cdot \lambda_{n} \lambda_{n-1} \lambda_{n-1}!
$$

The term by term addition of these equations gives

$$
\begin{equation*}
x_{n}-x_{0} \quad \stackrel{n-1}{\square} \Delta_{4}! \tag{11}
\end{equation*}
$$

If the error in evaluating each of the differences $\Delta$ is within the limits from $-\epsilon$ to $+\in$, then the maximal possible error in $x_{n}$ is $\pm n \in$. it is well known that the probability that the error in $X_{n}$ achieves this maximal value is very small. However, we still have to be very careful on the progressive decrease in accuracy of the evaluation of $x_{n}$ at large values of $n$.

In the following, we show how the above-mentioned accumulation of errors may be reduced. As an example, we will consider eatuion (10), which may be written as

$$
I_{1+1}-f_{k} 1 \vdots \text { R.d. }
$$

where Red stands for the correction that must be added to a given value $f$ in order to obtain the corresponding difference $\Delta$. For simplificyt, we assume that the values $f$ of the function are exact. Even in this case, the correction Red will have a finite error occurring as a result of the rounding off as well as the dropping of terms in equation (10). Hence, the accumbletion of errors in evaluating $x_{n}$ using equation (10) occurs due to two reasons. First, we use a limited number of laws in, evaluatiug the function $f$, and second, we make errors in evaluating Red when we round the numbers off and drop the small terms. The summation of errors of the first type is not important because the accuracy of the calculation of $f$ can always be put under control. The summation of the errors of Red is more harmful, since these errors cannot be easily controled.

It can be avoided by an appropriate change in method.
Let us substitute expression (10) into equation (11). This yields

Fron equations (1) and (3), we obtain

$$
i . \quad!i, \quad 1,1) \quad \because(1 ; \quad, \quad, \quad, \quad 1
$$

However, it follows from equation (3) that

Therefore,

$$
\prime_{1: 1}^{\prime}=J_{4 i} i^{\prime}-\prime^{\prime}
$$

Similarly

$$
\ddot{f}:!\cdots=f_{k+1}^{\prime} \cdots f_{k}^{1}, f_{k}^{\prime} \quad: \quad f_{k, 1} \quad f_{1} \ldots
$$

Using these equations, we can easily be convinced that

$$
\sum_{n}^{n} f_{1}!=f_{n}^{\prime}-f_{0}^{\prime}, \quad \sum_{n}^{\prime} f_{2}^{\prime}, \cdots f_{n}^{\prime} \quad f_{11}^{\prime}, \ldots
$$

Consequently, equation (12) can be written as follows

Since one tern in the column of the first sums may be arbitrarily chosen, then the quantity $f_{0}^{\mathbf{- 1}}$ is usually defined from the following condition

$$
\begin{equation*}
x_{0}-f_{11}^{-1} \quad \frac{1}{12} f_{11}^{1}: 11_{720} f_{n}^{\prime \prime} \quad \ldots \tag{1:3}
\end{equation*}
$$

Then

$$
\begin{equation*}
x_{n}=-f_{n}^{1}-\frac{1}{12} f_{n}^{1}+\frac{11}{720} f_{n}^{\prime}- \tag{1.1}
\end{equation*}
$$

The calculation using this formula is free from the above-mentioned limitation of the accumulation of errors. The error in $x_{n}$ depends on the ruunding off and neglection of the terms in the correction.

$$
\operatorname{Red} \cdots 1_{12}^{1} f_{11}^{1}: 1_{i 10^{\prime} f_{n}^{\prime}}^{11} \ldots
$$

When the interval $w$ is pruperly chosen, this error does not affect the value $f_{n}=w F\left(X_{n}, t_{n}\right)$ and hence does not affect the subsequent values of $x$.

The method of integration of equation (7), based on the application of a formula of the type given by equation (14), is called the method of quadratures since, if the right hand side of equation (7) does not involve $X$, this formula is reduced to the formula of quadratures (Sec. 56).

The method of differences, suggested by Adams and based on equation (9), corresponds to the method of quadratures in which the fo?lowing formula is applied
where the initial values of the column of sums are defined by

$$
f_{-\frac{1}{2}}^{-1}=x_{0} \cdot \frac{1}{f_{-1}}-\frac{1}{1} f_{-}^{\prime}-\frac{3}{2} f_{-2}^{2}-
$$

To distinguish between this method and the previous method suggested by Gauss, we shall call it Adams' method of quadratures.
50. A second form for the method of quadratures of first-order equations

We shall consider the foliowing problem. Let the integral of the equation

$$
\frac{d x}{d f} \cdot f(x, d)
$$

be given by

$$
x\left(\begin{array}{cc}
t_{0} & \prime \prime \\
& \ddots
\end{array}\right)
$$

It is required to calculate a table for the values of this integral that correspond to the following values of the argument

$$
t_{*}=t_{n} \left\lvert\, \begin{gathered}
\\
\hline
\end{gathered}\right.
$$

where $k$ is an arbitrary integer.
We first show how we can calculate the unknown function $X(t)$ for the following values of the argument

Let us define the difference

Expanding this quantity in a Taylor series, we obtain

Using equation (7) and applying formulae (4), we obtain

This formula is more convenient than formula (10). It involves only differences, while formila (10) Involves semisums of differences.

The initial values of the unknown function, e.g.
are obtained either by expanding $x(t)$ in a series, or by means of the method of successive approximations.

Formula (14) leads to a method of integration similar to the method used by Cowell. The version of the method of quadratures that corresponds ts this formula is obtained by summing equation (16) i. $\sim m k=0$ to $k=n-1 . \quad$ This summation yields

If $f_{-\frac{1}{2}}^{-1}$ is defined by
then, we obtain the following simple formula

$$
x\left(\begin{array}{ll}
1 & 1  \tag{1i}\\
\vdots
\end{array}\right) \quad \therefore \quad 1 \quad \begin{gathered}
1 \\
\hdashline
\end{gathered} \quad . \quad \begin{gathered}
i \\
\pi i l n
\end{gathered}
$$

The method based on the application of formulae (16) and (17) cannot be widely used because these formulae give the values $x\left(t_{k} \pm \frac{w}{2}\right)$, whereas we have to know the values $x\left(t_{k}\right)$ in order to evaluate the right-hand sides of these formulie. Therefore, when these formulae are applied we can find the values $x\left(t_{k}\right)$ by integrating in average. Hence, the application of formulae (16) and (17) is useful only when the differences of $x\left(t_{k}\right)$ may be neglected.

In order to avoid this difficulty, we deduce from equation (17)
a formula that yields $X_{n}=X\left(t_{n}\right)$. For this purpose, we use the well-known formula on the integration in average
which is cbtained from Bessel's formula, given in $\delta^{\prime} 47$, by putting $z=\frac{1}{2}$. Adopting in this formula that

$$
:(\because) \quad x\left(1_{1} \cdots \because\right)
$$

and noiling thai
and, consequently,

$$
\varphi_{n+\frac{1}{2}}^{2} j_{n}^{1}: \stackrel{1}{2 f_{1}} \quad i_{i n}^{i n} 1+\ldots
$$

we obtain

$$
\begin{array}{ccc}
:, & \vdots & 1 \cdots i \\
\vdots & 1,1!1
\end{array}
$$

Hence, the evaluation of $x_{n}$ will be carried out using a formula indentical with that given by equation (14). However, the initial term of the column of sums will be given in the presen': case by


## Annotation I

The methods of integration of equation (7), considered above, can be applied without change to systems of equations of the type.

In this case the integration will be carried out in parallel on three separate sheets.

## Annotation II

In the application of the above-mentioned formulae, it is useful to use the following approximate equations

| 11 | $\therefore 11$ | （\％） | 0）1．11：17］ |
| :---: | :---: | :---: | :---: |
| $1 \cdot 1$ | 1，1） 1 | 117 | 010）（H）1 |
| 1 19， |  | 1115 | （111410 |
| 11 | 1.0 | 1． 3 願 | －1月6011］ |
| ： $1 i$ | 「が，が， | 1 2liii | － $6 \cdot(111)$ |

51．An example of integrating first－order eqauations
Let us calculate a table for the values of the integral of the following equation

$$
\begin{array}{cc}
d: & \vdots \\
d! & 2
\end{array}
$$

which satisfies the inftial condition $t_{0}=0, X_{0}=1$ ．
Choosing the interval $w=0.1$ ，we obtain

$$
1(1) \quad 11.11,5.18
$$

$1+1$

In order to determine the first values of the integral，we differentiate the given equation and put $t=0$ ．We obtain

$$
x_{0}^{!}=11, x_{11}^{\prime 1}-\frac{1}{1}, x_{n}^{\prime \prime \prime}-11, x_{11}^{\prime \prime} \quad 1, \ldots
$$

Consequently，

$$
1: 1: \frac{1}{3}, 141 \ldots
$$

This series－expansion enables us to find the values of $x$ for $t= \pm 0.1$ ， $\pm 0.2$ ．Furthermore the calculation will be carried by the following scheme，in which the values of $x$ obtained by the series－expansion as well as the corresponding values $f, f^{\prime}, \ldots$ are printed in bold type．The semisums of the values $f, \mathbf{f}^{\prime}, \ldots$ are typed in the spaces between the corresponding lines．


In table $A$, the initial values are given, and the first approximation is obtained for $X_{3}=X(0.3)$ and $X_{4}=(0.4)$.

For the line $t=0$, we evaluate

and obtain Red $=-417$ (expressed in units of the 6th digit, which also applies to all quantities $f, f^{1}$, ...). We substitute ; this value in the corresponding column, and find tre principal term in the column of sums

$$
1_{1}^{-1} \therefore 1 \quad K 11 \quad 110110
$$

In order to find the next terms in the column of sums, we construct the semisums in the column of $f$ and use the following relations

$$
I_{1}^{\prime} \quad y_{0} \quad f_{1}, \therefore \quad 1, \quad!\quad . \quad \ldots
$$

We then extrapolate the quantities required for evaluating $x_{3}$ and $x_{4}$. Assuming that the forth difference is zero, we obtain

$$
f \quad j_{0} \quad,=-\prime \quad: 7+
$$

Then, by successive addition, we obtain

$$
\because \quad 111 . \quad \vdots \quad 18
$$

and so on, until we obtain $f_{3}^{-1}$ and $f_{1}^{-1}$. Finally, we evaluate Red using equation (**), and obtain the values of $x(t)$ which are subsequently tabled. At this stage, the first approximation is complete.

In order to obtain the second approximation, we evaluate $f_{3}$ and $f_{1}$ using equation (*). We then repest the calculation of $x_{3}$ and $x_{4}$ using the new and more accurate values for the differences.

The final results are given in table $B$. In practice, the first approximation is written in pencil while the second is inked in. In Table $B$, the quantities obtained by extrapolation and those not corrected by the second approximation are printed in italics


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 -Red. Instead of obiaining the dif. - . ces required for the evaluation si Red by extrapolation it is kettain * extrapolate immediately the $R$ d $\therefore$ antity itself. Such an extra.cl: is shown in table $C$, where the exiorglated values are printed ; . .ics.


The values of $x_{5}$ and $x_{6}$, given in table $B$, have been found by the above mentioned extrapolation. The first value is exact and the second value is incorrect by only one unit., However, the values $f_{5}$ and $f_{6}$ evaluated using these vaiues for $x_{5}$ and $x_{6}$ are fal and do not require re-evaluation.

In conclusion, we shall demonstrate how the same problen could be solved using Adam's method (of differences). In this method, the calculation is carried out using equation (9), which may be rewritten as follows

$$
\begin{aligned}
& A_{1,}: \rightarrow J_{n}: \mathrm{Red} \\
& \begin{array}{ll:l}
x_{n i 1} & x_{n} & د_{n} ; \\
\cdots
\end{array}
\end{aligned}
$$

The results of the calculation are given in table D．The exact values， from which we start the numerical integration，are printed in bilis type．The function $f$ and its differences as we11 as the Red correction are calculated in six Aigits．

Table D

| $t$ | $\boldsymbol{x}$ ， | $f$ | $f^{1} f^{2}$ | $f^{i} \quad!$ | $f^{\prime}$ | $f^{3}$ | Red |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $-02$ | 1．0100三10 | －0．010 101） |  | ！ |  | ！ |  |
|  |  |  | $+5087$ | 1 |  |  |  |
| －01 | 1.00250 | －0005013 | $-74$ |  |  | ＇ |  |
|  | $1 \quad 1$ |  | $\therefore 50131$ | ＋741 |  | 1 |  |
| 01 | 1.00000 | O．1ヶм）（\％）0 | 0 |  | 0 |  |  |
|  | 1 |  | $\div 5013$ | ＋ 711 |  | ： 6 |  |
| 0.1 | 1．00250） | 0.005013 | 1 ： 74 |  | ＋ 6 |  |  |
|  |  |  | － 5067 | 180 |  | $-6$ |  |
| $1 . ?$ | 1.01005 | $0.0101(0)$ | ＋17 |  | 0 | 1 | $+2002$ |
|  | 11270 |  | $+5241$ | 180 |  | .71 |  |
| U． 3 | 111.275 | 0.0153 .11 | $1+231$ | － | ＋$i$ |  | ＋ 2719 |
|  | 1 180 |  | ＋5．175 | ＋ 87 |  | －21 |  |
| 0.1 | 1110 sl | （1320816 | 1 • 321 |  | ＋ 4 |  | ＋28i3 |
|  | $+23 i 8$ | $1$ | $i .5796$ | －以 |  | ＋1 |  |
| 115 | $1.05 i+9$ | 00.66612 | ＋+17 |  | $+10$ |  | 13069 |
|  | ＋ $2 \times 68$ |  | ． 620.3 | 1 linj |  |  |  |
| 116 | 114.1417 | 00.12 d 25 | $1: \therefore ?$ | 1 |  |  | ＋3320 |
|  | －． 1615 |  |  |  |  |  |  |
| 117！ | 113432 | 00.34001 |  | 1 i |  | ， | $+36.9$ |
|  | － 4.114 |  |  |  |  | ＋ |  |
| 0.8 | 117351 |  |  | ， |  | 1 |  |
| ， |  |  |  |  |  |  |  |

All the calculated values of $x$ are quite accurate and the difference in accuracy between Adams＇method and the method of quadratures will be considerable．only if $t^{\prime}$ ．e numerical integration is continued significantly
further ior if larger intervals, $w$, are used) especially when function $f$ is $2 v a l u a t=d$ with a single spare index, as it has been done in the present example.
52. Integration of Second-Order Equations. Evaluation of an Integral

## Assigned by Two Values

We shall now consider methods for the integration of equations of type

$$
\frac{t^{2} x}{d d^{d}}=f(x, t)
$$

Each integral of a second-order equation can be assigned either by its values in two points, e.g. by the values

$$
x_{-1}-x\left(f_{0}-:(1), \quad x_{0} \cdots x\left(f_{0}\right),\right.
$$

or by the values

$$
x_{0}=x\left(t_{0}\right), \quad x_{0} \cdot\binom{d x}{d t}_{t},
$$

that this integral and its derivative take at the initial point $t_{0}$.
In this section we shall only c consider the first case. We assume that the initial values $x_{-1}$ and $x_{0}$ are given. We then have to evaluate the subsequent values $x_{1}, x_{2}, \ldots$ of the unknown function.

Alongisde the differences

$$
\therefore_{k-1}^{1} \cdots x_{k}-x_{h-1}, د_{h:}: \cdots x_{h+1}-x_{k}
$$

we introduce into consideration the second difference

$$
J_{k}=د_{k+1}-د_{k-} 1=x_{2 ; 1}-2 x_{1}+x_{k-1}
$$

Expanding $x_{k+1}=x\left(t_{k}+w\right)$ and $x_{k-1}=x\left(t_{k}-w\right)$ in powers of $w, w \in$ obtain

$$
J_{4}^{\prime}=\frac{2}{2!} u^{\prime 2}\binom{d^{2} x}{d!}_{2} \quad \frac{2}{1!} n^{\prime \prime}\binom{d^{4} x}{d l^{t}}_{1}!\ldots .
$$

Assuming that

$$
!n^{\prime} F(x,!)-f(!), \quad::^{\prime \prime} F\left(x_{n},!_{n}\right) \quad i_{n},
$$

we obtain

$$
د_{k}^{\prime}=f_{k} \cdot \frac{1}{12} w^{\prime}\binom{d / f}{d t^{2}}_{k}+\frac{1}{. k+1} u^{\prime \prime}\binom{d / f}{d / 1}_{k}+\cdots .
$$

Using equations (4), we express the derivatives in terms of differences. We obtain
which represents Cowell's method. Applying equations (6) yields the principal formula of Stormer'\} method (simjlar to Adam's method, which reads

The comparison between the last two formulae immediately indicates the superiority of Cowell's method. If the values of $f_{k}$ are evaluated with an error equal to $\pm \epsilon$, then the first, second, third, ... order differences will have errors within the limits $\pm 2 \in, \pm 4 \in, \pm 8 \in, \ldots$ In equation (20), all the coefficients are of the $1 / 12$ order and hence the errors in the higher-order differences will significantly modify the value of $\Delta_{K}^{2}$. Equation (19) is free from this defect because of the rapid decrease of its coefficients.

Calculation using Cowell's method is done in the following way. The values of $x_{-1}$ and $x_{0}$ are giver, and hence we can evaluate

$$
د_{-1}:=x_{11} \quad \Sigma_{1} ;
$$

Neglecting for the moment all the unknown terms in equation (19), we find the approximate value

$$
د_{i}^{*}=\mu_{11} .
$$

This yields

$$
\begin{aligned}
& د_{\frac{1}{z}}=\Delta_{-} \frac{1}{y}+د_{6}^{2} \\
& x_{1}=x_{0}+د_{\frac{1}{z}} .
\end{aligned}
$$

Having cbtained the values $f_{-1}, f_{0}$ and $f_{1}$, we determine $f_{0}^{2}$ and subsequently use the more accurate value

$$
s_{0}^{\#} \cdots f_{0}!\frac{1}{1!} f_{n}^{z}
$$

to evaluate $x_{1}$ once again. We similarly evaluate $x_{-2}$. We then find $f_{0}^{4}$ and use the more ac.rrate expression

$$
y_{i} \because_{0} \cdot \frac{1}{12} f_{0}^{2}-\frac{1}{240} f_{v}^{4}
$$

and so on until we obtain the final value of this quantity and consequently the value of $x_{1}$. Thus obtaining a few values $x_{1}, x_{2}, \ldots$ we evaluate ${ }^{\wedge} \underset{k}{2}$ by extrapolating the values of the unknown differences. If the interval w is not large, the extrapolation is done so well that it is never necessary to improve the accuracy cf the resulting values of $\Delta \underset{k}{2}$.

Instead of finding $x_{k+1}$ in terms of $\Delta_{k}^{2}$ by using the double summation

$$
\begin{equation*}
د_{k+}, \rightarrow J_{k-1}: J_{k}, \quad x_{k+1} \quad x, x_{i+1} \tag{21}
\end{equation*}
$$

it is possibletapplyl,the following formula
which is more easily done using a calculating machine. Hcwever,
avoiding the writing of differences does not save much time and it only prevents the possibility of checking and controling the calculations.

The double summation that explicitly appears in formilae (21) and which is implicitly invo'ved in formula (22) leads to a greater accumulation of errors than the single summation obtained by applying Cowell's method to first order equations. It is thus clear that the replacement of Cowell's method by the corresponding method of quadratures is very essential in the integration of second-order equations, especially when the calculations are prolonged.

Summing equation (21) from $k=0$ to $k=n-1$ yields

$$
\begin{equation*}
د_{n}:-د_{-}: \ddot{:}_{0}^{n} د_{i}, \quad x_{n} \quad x, \quad " \vdots د_{4} \frac{1}{\vdots} \tag{!3}
\end{equation*}
$$

Formula (19) gives

Since, according to our system of notation,

$$
\begin{aligned}
& f_{1,}^{m}=\cdot f_{!}^{m \cdot 1}-f_{-}^{m}! \\
& f_{1}^{m}=f_{n}^{m} \quad-f_{!}^{m} \\
& f_{n}^{m},=f_{n}^{m-1}-\frac{1}{\vdots}-f_{n-1}^{m-1} .
\end{aligned}
$$

The addition of these equations give

$$
{ }^{n} \stackrel{1}{\prime} J_{k}^{n} \cdot f_{n}^{m} \frac{1}{\vdots}-f^{m}!
$$

Taking this into consideration, we write the first of equa+ions (23) in the following inanner

$$
\begin{aligned}
& 1 د-!f_{-1}^{-1}-1!f^{\prime}-\frac{1}{!} \because 11^{3}-\frac{1}{!}
\end{aligned}
$$

We then choose the arbicrary initial serm in the column of first sums In such a manner that the second ine of this equation vanishes.

By assuming that

$$
f_{-1}^{-1} \quad د_{-1}^{2}-\frac{1}{12} \int_{-}^{1}!\frac{1}{10} t^{3}:-\frac{31}{r .1+0} f^{5} \frac{1}{2}-i
$$

We finally obtain

We replace $n$ by $k+1$ and sum from $k=0$ to $k=n-1$. The second of equations (23) may then be written as follows

However,

$$
\because y_{1}^{m}: \frac{1}{2} \quad J_{a}^{n+-1} \cdots t^{m-}
$$

Consequently,

We assume that the initial term in the column of the second sums is given by

$$
\begin{equation*}
!\cdot x-\frac{1}{12}: \sum_{210}^{1} f_{n}^{z}-: 01 \tag{25}
\end{equation*}
$$

Then, the latter formula will be given by

In applying these formulae, it is worthwhil.e to take into consideration that

Formula (26), supplemented with equations (24) and (25), represents the method of quadratures. This method is applied in the same way as Cowell's method. First, starting from the given valuex $x_{0}$ and $\Delta_{-\frac{1}{2}}=$ $x_{0}-x_{-1}$, the adjacent values $x_{1}, x_{-2}, x_{2}, \ldots$ are found using the method of successive approximations. Then, the differences $f_{n}^{2}, f_{n}^{4}, \ldots$ required for the application ofequation (26) are found by extrapolations and corrected if necessary in the second approximation.

In Cowell's method, the calculation is carried out using formulae (21) and (22), which miy be writicen as

$$
\begin{aligned}
& s_{i}=f!H \\
& 1_{-+}+\frac{1}{!} \quad x_{1}-x_{-1}+\frac{1}{1} 1_{i} \\
& \text {.r. r. } \quad=\underbrace{\prime}_{i} J_{A+\frac{1}{2}},
\end{aligned}
$$

The Red correction isthus subject to a double summation. However, in the method of quadratures, the calculation is done using a formula of the type

$$
B_{n}=f_{n} \quad \therefore \quad \therefore \cdot n
$$

Thus, the errors made when evaluating Red are not increased by the summation.

The method of quadratures that corresponds to Stormer's method consists in the application of the foilowing formula

In order to obtain the initial terms of the columns of sums, we put QUALITV' in this formula $n=0$ and $n=-1$. We then obtain

$$
\begin{aligned}
& f_{0}^{2}=x_{0}-\frac{1}{12} f_{11}-\frac{1}{12} f^{\prime},-19,211^{19}: \cdots . \\
& f_{1}^{\prime}=x_{1}-1_{12}^{1},=-\frac{1}{12} f_{-}^{\prime}:{ }_{2}^{11} f^{\prime}, \ldots \ldots
\end{aligned}
$$

We obtain for the column of the first sums

$$
f_{-1}^{-1}:=f_{11}^{-2}-f_{1}^{-1}
$$

Stör"er's formula is applied in the following form

$$
x_{n}=f_{n}^{2}+1_{12}^{1}\left(f_{n}, \vdots f_{n-1}^{\prime}+f_{n-2}\right)-{ }_{210}^{1} f_{n-n}
$$

The tabular interval $w$ is chosen so small that the third difference may be neglected.

## Annotation

All the methods of numerical integration of the equation

$$
d: x=f(x, 1)
$$

are applied without alteration to systems of the type

$$
\begin{array}{llll}
d: x \\
d f
\end{array}=F(x, y, z, t), \quad d=y \quad(\|, x, y,=, t) \quad d: \quad d y=H(x, y, z, n) .
$$

Naturally, the integration of equations coupled in a system is carried in paralle.
53. A Second Case in the Evaluation of the Integral of a Second-Order

## Equation Assigned by Two Values

The method of quadratures forthe second-order as well as for the first-order equations can be represented in two ways depending on wi ther we want to calculate the valur $x_{n}=x\left(t_{0}+n w\right)$ or the values of the functions in the middle of the intervals, i.e. $x\left(t_{o}+\left(n+\frac{1}{2}\right) w\right)$. The first case was considered in the previous section. Here, we shall derive the necessary formule: for the second case. We assume that

$$
\begin{aligned}
& i_{i!}: \cdots i_{i n}-i_{k}
\end{aligned}
$$

Taylor's formula gives

Consequently
or, using equation (5),
or, finally, using the following formula

$$
f_{k}^{r-1}!f_{k}+!f_{k+1}^{1}
$$

which has already been applied in Sec. 48,

This formula gives a method of integration similar to Cowell's method Summing this formula from $k=0$ to $k=n-;$, we obtain

$$
\begin{aligned}
& i_{n}-i_{0} \cdot J_{n}^{\prime}-\frac{1}{24} f_{n}^{\prime}+\frac{17}{19} f^{\prime} \cdots . .
\end{aligned}
$$

Assuming that

$$
\begin{equation*}
j_{v}^{-1} i_{v} f \cdot \frac{1}{2 f} f_{0}^{1}-\frac{17}{1920} f_{1}^{1} \quad: \ldots \tag{27}
\end{equation*}
$$

Then

$$
z_{n} \because f_{n}^{1}-\frac{1}{21} f_{n}^{1}-1 \cdot{ }_{1!100}^{i} f_{n}^{1} \cdots \cdots
$$

Replacing in this equation $n$ by $k$, and summing from $k=0$ to $k=n-1$, we obtain

$$
\begin{aligned}
& \Gamma_{-}^{2}:+_{24}^{1} f_{-}-{ }_{1920}^{17}, 1+\ldots
\end{aligned}
$$

Assuming that

$$
\rho-\frac{1}{2}\left(t_{0}-\frac{i}{1}\right) \cdot \frac{\vdots}{1} f_{-\frac{1}{2}}^{1, \cdots 1}-1+
$$

we finally obtain

$$
1 \quad\left|\therefore-\frac{1}{2}\right|=1_{m-2}^{-2}-\frac{1}{\therefore 1} f \cdot \frac{1 i}{1!1+1} r^{2}:
$$

When the initial values $x\left(t_{0}-\frac{w}{2}\right)$ and $x\left(t_{0}+\frac{w}{2}\right)$ that define the integral under consideration of equation (18) are given, then formila (29) together with equations (27; and (28) enable us to successively obtain the values of $x(t)$ for $t=t_{n}-\frac{w}{2}$, where $n=2$, 3, 4, ...

It is advisable to have formula (28) written in a slightly different form. Since

$$
\begin{aligned}
& f_{1}, \cdots f_{4} \quad!_{2}^{\prime} f^{1} \ldots \ldots
\end{aligned}
$$

then this formula may be replaced by

Similarly, we replace formula (27) by

$$
\begin{equation*}
f_{1}^{:}=z_{0}+\frac{1}{2} f_{11} \cdot 1 \cdot \frac{1}{2 \cdot 1} f_{0}^{1}-17 f_{10}^{1}+\ldots \tag{:27nin}
\end{equation*}
$$

In evaluating the right-hand side of formula (29), one has to know the velues $x\left(t_{n}\right)$. The application of this formula is thus complicated by the requirement of finding the values of $x\left(t_{n}\right)$ by means of ir.terpolation into the middle of the intervals. We can obtain a formula that immediately gives the values $x\left(t_{n}\right)$. We apply the formula of the interpolation in the average, given in Sec. 50, to the function $x\left(t_{n}-\frac{w}{2}\right)$, defined by formula (29). Assuming again that

$$
\zeta(:, n)=x\left(!_{n}-u^{\prime \prime}\right)!
$$

we obtain

$$
\begin{aligned}
& \bar{F}_{n}!\quad ; \quad-1 \jmath_{n},-\frac{17}{1920} j_{2}-\ldots
\end{aligned}
$$

Therefore
where the initial values of the columns of sums are defined by formula (27) and (28) or (27 bis) and (28 bis).
54. Evaluation of the Integral Assigned by the Initial point and

## initial velocity

We have been considering the evaluation of the integral of the euqation

$$
\begin{aligned}
& d x \\
& d t^{2}
\end{aligned} \quad(\because x, x)
$$

that is assigned by the two values of trs independent variable, e.g. giving values of $x(t)$ for $t=t_{0}$ and $t=t_{0}-w$ and for $t=t_{0}-\frac{w}{2}$

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and $t=t_{0}+-\frac{w}{2}$. We now consider another problem of particular interest to celestiai mechanics which is the evaluation of the integral defined by the values $x$ and $\frac{d x}{d t}$ in some point. We shall also consider here two cases:

The first case
Let the unknown integral $x(t)$ be assigned by the initial values

$$
x\left(t_{1,}\right)=x_{11}, \quad\binom{d x}{d!}, \quad \& \quad \cdots \quad x_{11}^{\prime} . \quad \text { is }
$$

We shall now see how the adjacent values $x_{1}, x_{2}, \ldots$ can be obtained Considering equations (26), (24) and (25) which born the basis of the method of quadratures. We manipulate theconstants $f_{-\frac{1}{2}}^{-1}$ and $f_{0}^{-2}$ in such a way that formula (26) gives a function that satisfies the initial conditions (30.. It is clear that the quantity $\mathrm{f}_{\mathrm{o}}^{-2}$ should be left in the same form as given by equation (25); the first condition of which is given by equations (30) would be satisfied. Now, denoting by $\Delta^{4}, \Delta^{2}, \Delta^{3}, \ldots$ the successive differences of the function $x(t)$, we obtain on the basis of equation (4) the following relation

Formula (26) yields

$$
د^{\prime \prime \prime}=t_{n}^{n \cdots}+\frac{1}{1: 1_{n}^{m}} \quad 1 \quad 210 f_{n}^{m}+\cdots .
$$

Consequeratly

Noting that

and puting $n=0$, we finally obtain

It is therefore sufficient to calculate the initial values of the sum columns using formulae (31) and (25) to enable formula (25) give the values of the integral that satisfy the initial conditions (30). The second case

Consider now the case, when the integral is assfigned by the fcllowing initial conditions

$$
x\binom{t_{0}-w}{2}=X_{0}, \quad x^{\prime}\left(t_{0}-\begin{array}{c}
w^{\prime}  \tag{i32}\\
2
\end{array}\right) \quad X
$$

It is required to calculate $x_{n}=x\left(t_{0}+n w\right)$. Let us start by applying che previous approach that has been applied to formula (29). For a change, we will use another method which will as rapidly lead us to our aim. Replacing equation (18) by the system

$$
\begin{gather*}
d x^{\prime}  \tag{3,3}\\
d l
\end{gathered}=F(x, t), \quad \begin{gathered}
d x \\
d l
\end{gather*}
$$

We integrate the first of this equation by using formula (17). Since, in our present notations,

$$
w f(x, t)=-\frac{1}{w} f(t) .
$$

we obtain
where the first sums $f^{-1}$ are defined by condition (16), namely

We now consider the second of equations (33). Adopting that

$$
w^{\prime} x^{\prime}(t)=h(t)
$$

then, using equation (17), we obtain
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$$
x\left(t_{n}+\begin{array}{l}
w  \tag{36}\\
2
\end{array}\right)=n_{n, i}^{-!}:+{ }_{24}^{1} n_{n+1}^{\prime}-\frac{17}{i} 760_{n, 1}^{:}+\ldots
$$

In order to evaluate the right-hand-side expression, we have to know the quantity

$$
h_{n}=h\left(f_{0}+n u^{\prime}\right)=w^{\prime} x^{\prime}\left(f_{0}+n w^{\prime}\right),
$$

because formula (34) yields the values $\cap \mathrm{f} \mathrm{x}^{\prime}(\mathrm{t})$ for points $\mathrm{t}_{\mathfrak{n}}+\frac{\mathrm{w}}{2}$, which lie in the middle of our intervais.

Hence, we recall the formula of integration in the average, (Sec. 50),

Substituting in this formula

$$
\left.\begin{array}{c}
氵_{n}=-t_{n}-\frac{u}{2} \\
\because\left(\because_{n}\right)==\xi_{n}=h\left(t_{n}-\frac{u \prime}{z}\right.
\end{array}\right)
$$

and taking into account that on the basis of equation (34)

$$
\begin{aligned}
& \left.\varphi_{n+1}=\frac{1}{2}\left(\hat{F}_{n}-1 \hat{o}_{n+1}\right)=\begin{array}{c}
w \\
2
\end{array} \left\lvert\, x^{\prime}\left(t_{n}-\begin{array}{c}
w \\
2
\end{array}\right)+x^{\prime}\left(t_{n}+\begin{array}{c}
w \\
2
\end{array}\right)\right.\right]= \\
& ==\delta_{n}^{1}+{ }_{24} f_{n}^{1}-17{ }_{5} 760{ }_{n}^{1}+\ldots .
\end{aligned}
$$

and, consequently,

$$
\varphi_{n+}^{2}:=f_{n}^{1}+\frac{1}{91^{\prime}} f_{n}^{2}-{\underset{5}{5} 7 i 0}_{17}^{f_{n}^{5}+\ldots}
$$

we finally obtain

$$
A_{n}=\div\left(x_{n}+\frac{!u}{2}\right)=f_{n}^{-1} \cdots \frac{1}{12^{\prime}}{ }_{n}^{\prime}+\frac{11}{720} f_{n}^{1}-\frac{191}{00180} f_{n}^{3}-1 \cdots
$$

Substituting these values of $h_{n}$ into formula (36) yields

Pusting $n=-1$, we obtajn the following formula
wich deiermines the initial term of the coinn for the second sums.
Formulae (37), (38) and (3j) give the solution to the problem under consideration. Formula (38) may be replaced by another formula, which gives an expression for $\mathrm{f}_{0}^{-2}$. This will be more convenient. Since,

$$
\begin{aligned}
& f_{-1}^{2}=\frac{1}{2}\left(f_{0}^{2}+f_{1}^{2}\right) \cdot \Gamma_{0}^{2}-{ }^{2}{ }^{\prime} \\
& f_{-1}^{1}=f_{0}-\frac{1}{2} f_{\ldots}^{1} \ldots . .
\end{aligned}
$$

then equation (38) may be transformed into

$$
f_{0}^{-3}=X_{0}+\frac{1}{2} f_{-1}^{-1}+t_{24}^{1} f_{0}-\frac{1}{48} f^{\prime}-1-\frac{17}{1420} f_{0}^{:} \int_{38+11}^{17} f^{3}:-
$$

Adding this equation to equation (35) multiplying each term by $\frac{1 / 2}{}$, we obtain
or, finally,

There on'y remains to replace formula (37) by another formula which gives the values of $x\left(t_{n}\right)$. However, this has been done in the previous section.

Thus, in both cases, the values of the integral are svaluated using the following formula

However, in the first case, the initial values of the column of sums are calculated by using formula: (31) and (25), while in the second case, they are obtained from formulae (35) and (39)

## 55. An example of integrating second-order equations

Let us consider se equation

$$
\begin{equation*}
d_{d}^{2} x=\frac{1}{2}\left(1+\frac{1}{2} t^{\prime}\right) x \tag{'}
\end{equation*}
$$

We want to calculate a table for the values of the integral, determined by the condition

$$
x(\quad 0.0 .5)=x(-90.05) \quad 1.000(025 .
$$

This can be solved using either the formulae given in See. 52, or those of Sec. 53. Ir the first case we obtain a table for the arguments $t=-0.05,0.15, \ldots$. In the second case, the arguments of the table will have the values $t=0.0,0.1,0.2 \ldots$.

We choose the second way and accordingly we put

$$
\begin{aligned}
& f=(i .005)(1+!!i x \text {. }
\end{aligned}
$$

We shall apply the formulae of Sec. 53 in our calculations, nanely

$$
\begin{aligned}
& r_{n}=f_{n}^{-2}+\operatorname{Red},
\end{aligned}
$$

We obtain the initial values $x_{-2}, x_{-1}, x_{0}, x_{1}, x_{2}$ by successive approrimations and 78 a first approximatior we take
and evaluate the corresponding values of $f$. These are given in table A. Using the above formulae: we obtain

## Table A

| f | 1 | $\rho$ | J |  | $t^{\prime}$ | $J$ | 1.10 | $x$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $1: 1$ | 10016 |  | U00.30: | , |  |  | . 1. | 100:31 |
| 00 | $09495!$ |  | 110.710 |  | $-3$ | : is | $+12$ | 10.0011 |
| 01 | $1 \mathrm{CO2} 1 \mathrm{r}$ ! | + $10(x) \cdot$ ( 1 | (101506 | $!$ | $+3$ |  | - 13 | 1(x)251 |

$$
f_{z}^{1}=0.100250, \quad J^{-2}=(1.999 .94
$$

and by successive additions we fill the column of sums. In order to obtain new more accurate values for $x$, we have to calculate the red correction. With these values, we obtain the second approximation given in table B. In this table, the values $x_{-2}, x_{2}$ and $x_{3}$.

Table B

| 1 ; | $f^{-7}$ | $\mathrm{fl}^{1}$, | $f$ | $s^{\prime}$ | $1{ }^{\circ}$ | Prd | $\boldsymbol{x}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0.2 ; | $1.1 \times 962$ | -11.0075m | 0.1nis 35 | - 114 | . 70 | i 127 | 101008 |
| -0.1. | $100.00: 1$ | 1-0.00:500 | 000.51 .38 | .. 3 i | : 70 | $+118$ | 109250 |
| 0.01 | 0.909581 | +0.012 50, | $1100 . j 30)$ | : ix | : 'ti | $\because 110$ | $1(x) y$ |
| 01 | 1.602081 | +110017!23 | $11(0) .0038$ | $+111$ | 10 | $1 .+18$ | 1.0以~N |
| (1) | 1.0018.2 | - 11012 0 m | (1818: | $+161$ | 1. 30 | 1.1.7 | 1110115 |
| 0.1 | 1.102: 112 | 1 | (1)NS if: |  | . 76 | 1115 | 11127 |
| 114 : | 1.17511348 | - $10.11 / 511 i^{\prime}$ |  |  |  | $\ldots f(n)$ | 119081 |

are obtained by extrapolating the values of $f_{-2}, f_{2}$ and $f_{3}$ (indicated as all other extrapolated values, in italics). Finaliy, $x_{4}$ is caiculated by extrapolating the values of Red (cf. Sec. 51).

Table C


In the third approximatior: the extrapolation is carried further by three intervals. The values of $x_{5}$ and $x_{6}$ are exactly obtained, but there is an error of two units in the fifth decimal place in the value obtained for $\mathrm{f}_{\mathrm{x} 7}$.

If we wish to avoid successi": approximations in evaluating $x_{5}, x_{6}$, $x_{7}, \ldots$ we can use Störmer's formula

$$
\begin{align*}
& x_{n}=f_{n}=-1 \cdot K \cdot d \tag{}
\end{align*}
$$

The application of this formula in the case when the third difference is negligible, is as simple as calculating ky means of the conventional formula

$$
\begin{equation*}
\operatorname{Red}=\frac{1}{1: 2} f_{n}-\frac{1}{210} f_{n}+i \cdot \tag{}
\end{equation*}
$$

The advantages of the two methods are made use of from time to time in order to correct the values obtained ty formula (**) using the more exact formula (***), If for example we tabulate the values of the integral of equation ( $\dot{( })$ that satisfies conditions

$$
x(-0.05)=1.00002 \pi=\quad x^{\prime}(-0.0 .3)=-0.0 \therefore 01 . \pi i
$$

for $t=0.0,0.1,0.2, \ldots$, we evaluate the initial terms of the columns of sums using formulae (35) and (39), which may be written as

Finally, we note that the function

$$
\cdot(1) \quad r^{\prime}
$$

calculated in this section, is identical to the function calculated in Sec. 51. The comparison between these two examples suggests that the numerical integration of the second-order equation is not more difficult, even easier, than the numerical integration of first-order equations.

## 56. The formulae of quadratures

Applying the formula derived in Secs. 49 and 50 , to the integration of the following equation

$$
\begin{equation*}
d x=f(t) \tag{1}
\end{equation*}
$$

The solution of this equation that satisfies the initial condition $t=t_{0}, x=0$ is given by the integral

$$
x=\left.\right|_{t_{0}} ^{1} F(t) d t
$$

On the other hand, the particular value of this function at $t=t_{n}$ is given by formula (14). Hence,

$$
\begin{equation*}
\int^{9} f(f) d t \quad f_{n}^{\prime} \quad 1, f_{n}^{\prime}: \frac{11}{7!1 f_{n}^{:}} \ldots . . \tag{41}
\end{equation*}
$$

where

$$
f(t)===(u f(t)
$$

The first term of the column of sums is defined by

$$
\begin{equation*}
f_{0}^{-1} \therefore \frac{1}{12} f_{0}^{1}-\int_{120}^{11} f_{0}^{1} 1 \ldots \tag{42}
\end{equation*}
$$

Similarly, applying formulae (17) and (17 bis) to the calculation of the integral of equation (40) that satisfies the initial condition $t=t_{0}-\frac{w}{2}, x=0$, we obtain

11
where, in this case, the initial term of the column of sums is given by

We obtain another formulae of quadratures if in equations (43) and (44), we.make the initial term of the column of sums subject to condition
(42). We then put $n=0$ in equation (44) and substract it from equation
(43). In this way, we obtain

In order to obtain the formula of quadratures that solves a double integral, we consider the second-order equation

$$
\begin{array}{ll}
d r  \tag{47}\\
d: & f(t) .
\end{array}
$$

## ORIGINAL PAGE IS

The solution of this equation, which satisfies the conditions $x\left(t_{o}\right)=0$ and $x^{\prime}\left(t_{0}\right)=0$, is given by

$$
x-\int_{i,}^{0} d t \int_{i,}^{\infty} r(t) d t
$$

On the other nand, this solution is given by formula (26). Therefore, we find that

$$
\begin{equation*}
\int_{i}^{1} d l \int_{i}^{1} r(t) d f \quad f_{n}^{-2}+\frac{1}{12} f_{n}-\frac{1}{-10} f_{n}^{*}+\cdots \cdot \tag{48}
\end{equation*}
$$

where the initial terms of the column of sums is given dy

We now consider the solution of equation (47) that satisfies the conditions $x\left(t_{0}-\frac{w}{2}\right)=0$ and $x^{\prime}\left(t_{0}-\frac{w}{2} j=0\right.$. Using formulae (37) and (37 bis), we obtain

$$
\begin{aligned}
& \because=: \quad \text { : }
\end{aligned}
$$

where we now obtain

Applying a similar approach, to the one used in obtaining formula (46) from equations (50) and (51), we obtain
where the columns of sums are determined by condition (49). ORIGINAL PAGE F F All the formulae, which have been obtained, may be unified in the following manner

$$
\begin{align*}
& \frac{1}{w^{2}} \int_{i}^{t_{n}} d t \int_{i}^{t} i(t) d t \quad f_{-}^{2} \cdot \frac{1}{12^{\prime}} f_{n}-\frac{1}{240} f_{n}^{2}+\frac{31}{(0) 180} t_{n}^{4} \cdots . \tag{II}
\end{align*}
$$

$$
\begin{aligned}
& +\frac{17}{1+20} f_{n}^{2}: \frac{36,7}{\sqrt{1936,514} f_{n+}}
\end{aligned}
$$

where, if $A=t_{0}$, the initial terms of the column of sums are evaluated by equations (42) or (49), and if $A=t_{0}$ they are evaluated by formulae (45) or (52). For other values of A these initial terms are evaluated by the general formulae, that have been obtained in Secs. 49, 50 and 51.

It is interesting to note that equation (42) is equivalent to the first of equations (49).

## Annotation

Thefollowing effect, which is met with in any interpolation formulae, and in particular, in numerical integrations, is easily showr on the simple equations (4e) and (47): For any interval w there exist some functions $F(t)$, for which formulae (I) or (II) give results that differ by an arbitrarily large amount from the actual results, even if all the terms on the right-hand side of these equations, that affect the
result, will be taken into consideration. Indeed, let us consider the following function

$$
\therefore(1)=\therefore \cdots: \frac{2 \pi\left(t-t_{0}\right)}{i} .
$$

All the values of this function that correspond to $t=t_{0}+n w$ are zero. Formulae (41) and (42) give for the integral

values, which are also equal zero. However, the integral is evidentiy not equal zero. It can be as large as possible when the initial value $A$ is properly chosen. Hence, the integrals of functions $F(t)$ and $F(t)+\varphi(t)$ will differ by an arbitrarily large amount while the numerical integration will give the same value for the two functions. This example illustrates that in numerical integration as well as in any other processes inviving interpolations, it is not sufficient to have a table for the values of the function, but it is also necessary to know its analytical character.
57. Basis of the Successive Approximation Method

Let us take a first-order equation, e.g.

$$
\frac{d n}{d t}=r(x, d
$$

and investigate the calculation procedure of its integral, which satisfies the initial condition $t_{o}, x_{o}$. The unknow integral evidently satisfies the following integral equation

$$
r=i_{0}+\int_{t_{2}}^{t} r^{-}(x, t) d t
$$

Inversely, the function that satisfles equation (54) also satisftes the differential equation given above as well as the required initial condition. Hence, obtaining the required par sular solution of the above-mentioned differential equatior is equivalent to solving the integral equation (54).

Let us take an arbitrary function $\xi_{0}(t)$ that satisfies the condition $\xi_{0}\left(t_{0}\right)=x_{0}$, and then evaluate a new function $\xi_{\rho}(t)$ using the following formula

$$
\xi_{1}(f)=i_{0}+\operatorname{t}_{0} F\left(\xi_{0}(i),\{ ), \ldots\right.
$$

The new function also satisfies ${ }^{t_{0}}$ the condition $\xi_{,}\left(t_{0}\right)=x_{0}$. Continuing this procedure, we obtain a set of functions

$$
\xi_{0}(t), \xi_{1}\left(* 1, \xi_{2} \mid f 1, \ldots\right.
$$

connected together by the following relation

$$
\xi_{n+1}(t)=x_{0}+\int_{t_{0}}^{t} F\left(\xi_{n}(t), t\right) \cdots
$$

We shall prove that

$$
\lim _{n \rightarrow \infty} ;(1):=x(f) .
$$

For this purpose let us substract equation (55) from euqation (53), term by term, and consider the difference

$$
x-\xi_{n+1}=\int_{i}^{!}\left\{F(x, t)-F\left(s_{n}, n\right)\right\}: d t
$$

The theorem of finite increments gives

$$
F\left(x, f ;-f \xi_{n}, n!<n x-\xi_{n}\right.
$$

where $M$ denotes the upper limits of the partial derivatives $F_{x}^{\prime}\left(x_{1} t\right)$ in the required variation region $x$ and $t$.

Consequent1y,

$$
1-1.1 \cdot .111-\varepsilon_{0} \cdot 1 \quad \vdots
$$

We subordinate the interval $t-t_{0}$ to the condition

$$
\begin{equation*}
.11|t--t,|<4 \tag{5,6}
\end{equation*}
$$

where q is a proper fraction such that $0<\mathrm{q}<1$. Then

$$
\left|x-i_{1+1}\right|<4 \quad x \cdots \theta_{4} .
$$

Applyfng this inequality for $n=0,1,2, \ldots, n-1, n$, and multiplying the resulting inequalitiés : term:by term, we obtain

It is thus clear that when condition (56) holds, function $\xi_{n}(t)$ tends to the unknown function $r(t)$ when $n$ tends to infinity. Therefore, if we take an arbitrary function $\zeta_{0}(t)$ and apply formula (55) a sufficiently large number of times, we obtain $x(t)$ with an arbitre $y$ accuracy. The required accuracy will be obtained more rapidly, with the smaller numbers for $q$, i.e. the smaller the interval $t-t_{0}$ is and with the better choice
for the initial approximation $\mathcal{E}_{0}(\underset{)}{ }$.
In the example considered in Sec. 51 , we could take $\mathcal{E}_{G_{3}}(t)$ to be the constant value $x_{0}$. After two successive approximations, we would be able to find the correct values of $x(t)$ for values of $t$ near the initial value $t_{0}$. When a few values for $x(t)$ are obtained, it would be better not to use the previous trail function but to constrctu by extrapolation a talk of values for $x(t)$. It is then possible to construct a new function, using this table, which would be closer to the unknown function $x$ ( $t$ ). We then substitute this function into formula (55), and proceed in the same way as we have done before in applying the method of quadratures.

The application of successive approximations to the method of quadratures for second-order equations can be justified in a similar way.
58. Different methods for the reduction of the number of successive approximations

The methods oî numerical integration of differential equations can be divided into two groups. The methods of Adams and Sotrmer and the corresponding methods of quadratures belong to the first group, while Cowell's method and the conventional method of quadratures belong to the second group.

The methods of the first group make use of the differences located in the ascending diagonal. Each value of the unknown function is thus evaluated using only its preceding values. These methods may be called extrapolational. On thecontrary, the methods of the second group' made use of the differences lying on the horizontal line. Therefore, the preceeding and the following values equivalently, participate in the calculation of each of the values of the unknown function. That is why these methods may be called interpolational.

The interpolational methods lead to more accurate results, however, in the calculations of this method, one has to make use of several successive approximations. If the nimber of the required successive approximations is not greater than two. this is then not considered as a weakness, because successive approximations help in avoiding errors of calculations. However calculations with a large number of successive approximations are not valid. In the fcllowing, we shall consider the method for reducing the number of approximations.

## The first method

The simplest and most convenient method to reduce the number of apprcximation consists of decreasing the interval w. This interval may be taken of such dimensions that the extrapolation of the values of the unknown function becomes sufficiently accurate in order to obtain the final values of the function $f$. In this case, the second approximation will only change thedifferences and sums of the first approximation and will be accepted as a final result.

The authoritive astronomer Comri mentioned only this method in his precept for using numerical integrations in problems of celestial mechanisms (1) and particularly warned against using large intervals during interpolations. He wrote: "The computer should be warned against attempting to use tov large an interval, the result of which is that checking by difference, which is essential in these methods, becomes ineffective. A safe guide is that the sixth difference should not exceed two figures". Comri suggested the use of the following rule: "the interval should be such, that the sixth difference should not be more than two significant Eigures.
(1) Planetary Co-ordinates for the years 1800-1940 referred to the Equinox of 1950.0. Lonidon, 1933.

If we express in the conventional formula for quadratures,
the equantities $f_{n}, f_{n}^{2}, \ldots$ (the presence of which makes the successive approximation indespensible) in terms of quantitiess located in the ascending diagonal, we obtain the quadratic form of Stormer's method namely

$$
x_{n}=f^{-2}: \frac{1}{1!}!_{-1}+\frac{1}{1!} \prime_{1-\frac{3}{2}}^{\prime}!\frac{11}{!^{\prime} 11} f_{n-2}^{2}+\cdots \cdot
$$

Formula (58) does not require successive 9 pproximations. It,yields however somewhat less accurate results than formula (57) (cf.Sec. 52).

In order to unify the conclusions obtained from there two formula, we pause midway in transition from equation (57) to equation (58). We have


Stopping, for example on the second of these expressions, we obtain

$$
x_{n} \quad x_{n}: \sigma_{1,}
$$

where

$$
\begin{aligned}
& \left.r_{n} \cdot f_{n}^{\prime}+\frac{1}{12}\left(J_{n-1}+J_{n}^{\prime}-\frac{1}{2} f_{n}^{2}\right)-\frac{1}{240} f_{n}^{2} \right\rvert\, \cdot \begin{array}{c}
31 \\
60+181 J_{n}^{\prime}-\ldots
\end{array} \\
& \begin{array}{ll}
\sigma_{n} & \frac{1}{12} \prime_{n-1}^{\prime}
\end{array}
\end{aligned}
$$

We immediatel.y obtain the final result for $x_{n}$, since the errocm fuo $G U A L I M$ extrapolating the values $f_{n}^{?}, f_{n}^{1}, \ldots$ will not effect the value of $x_{n}$ as an result of the smaliness of their coefficients. Only the correction $\sigma_{m}$ will be alterted in the subsequent approximations. Hcwever, this correction is very small and its evaluation is not difficult.

## The method of Tiet:jen

Tietjen ${ }^{(1)}$ observed that the need for successirs approxjmations results from the presence of the term $1 / 12 f_{\mathrm{n}}$ in the right-side of equation (58). A11 the other terms can be obtained by extrapolation and the resulting values are pracically quite accu ste. Therefore, taking into account that

Tietjen replaced formula (57) by
 Covern visintiv
where

$$
s_{n} \cdot f_{n}: \frac{1}{2411} f_{4}^{:}: \begin{gathered}
31 \\
w(1) 400
\end{gathered} f_{n}^{\prime} \cdots
$$

The final value of $S_{n}$ can be immediately obtained. The solution of equation (59) yields the unknown value $x_{n}$ add in this way, only the differences $f_{n}^{2}, f_{n}^{1}, \ldots$ will be altered during the successive approximations. This method will be of practical interest only if the solution of equation (59) sor $x_{n}$ is sufficiently simple.
(1) F. Tietjen, Specielle Storungen in Bezug auf Polar coordinaten, Berl. Astr. Jahrbuch fur 1877, M.F. Subbotin, On the Numerisal Integration of Differentia! Equations ( 0 cislennom integrirovanif differencial'nyh uravnenij) Proceedings of the Academy of Science of USSR. (Izvestija Academii Nauk SSSR), 1933, No. 7, pp. 895-902.

## The method of Numezov

The same procedure winich was followed by Tietjen in developing the method of quadratures was later on applied ly Numerov ${ }^{(1)}$ in order to develop Cowell's method. Let us consider the principal formila in Cowell's method namely

which can be used together with the following equation

$$
x_{k+1}=\because x_{k}-x_{k}, د_{4}
$$

to obtain successively the values $x_{1}, x_{2}, \ldots$. Let us introduce the new vari $\mathfrak{b}$ le is by adopting that

$$
x-r-\frac{1}{12} f
$$

When $t=t_{k}$, we shall have

$$
\begin{equation*}
x_{1} \therefore \lambda_{a}-\frac{1}{12} t_{1} \tag{6,2}
\end{equation*}
$$

"Denoting by $\Delta, \Delta^{2}, \ldots$ the differences of $x$, we obtain as a conseqיence of equation (62) the following relation

$$
s_{k}^{2} \cdots+3-\frac{1}{12} f_{i}
$$

(1) B. Numerov, Methode nouvelle de la determination des orbites et le calcul des ephemerides en tenant compte des perturbations, Transactions of the Principal Russian Astrophysieal Obsorvacory (Trudy G1. Rossijskoj Astrofiz. Observatorri) rol. II, 1923.

Thus, we obtain for the new variable

$$
9 \div f_{k} \quad{ }_{2}^{2}-40 f_{k}^{\prime}+\begin{gathered}
31 \\
00) 480 \\
f_{2}^{\prime \prime}-\ldots
\end{gathered}
$$

and, as for any function,

$$
\therefore \therefore_{n+1} \quad \ddot{x}_{k} \cdot x_{n-1}+1_{k} .
$$

The two equations are quite equivalent to equation (60).
In order to calcrlate $f_{k}=w^{2} F\left(x_{k}, t_{k}\right)$, it is necessary to know $x_{k}$. Therefore, it is necessary to express the special coordinates $x_{k}$ obtained in this method in terms of the conventional coordinates $x_{k}$.
It is necessary for this purpose to solve equation (62), which may be written as

$$
\begin{equation*}
x-x_{k}-u_{1: 2}^{2} f\left(x_{1,}, b_{n}\right) \tag{15.3}
\end{equation*}
$$

and which corresponds to equation (59) in the method of Tietjen. Jet us assume that the interval $w$ is chosen in such a manner that the terms
may be neglected. Ne then obtain a very simple formula for t'se subsequent evaluation of the "special coordinate" $x$, namely

$$
x_{1,1}=\ddot{x_{1}}-x_{1}, 1!
$$

This formula is equivalent to equation (60), which may be written as

The so-called method of Numerov or method of extrapolation ${ }^{(1)}$ consists in the application of this formula,

In the method of Numerov, one does not need to construct the differences of the quantities $f_{k}$. If these differences could be constructed, it would be easier to calculate using formula (61) than to make a
(1) This latter name is not convenient, for the word extrapolation has a generally accepted wider meaning,
transition into the "special coordinates", and the accuracy would not then be easily controled. The guarantee against errors may be ubtained only by the use of various special control calculations.

This modification of Cowell's method may be applied when the following =ron conditions hold: 1) It is possible to guarantee beforehand the insensitivity of terms (64) during the whole process of integration. (In Cowell's method and in the method of quadratures it can always be sean which of the terms cau be neglected and which cannot). 2) The solution of equation (63) for $x_{k}$ is sufficiently simple. The drawbacks of Cowell's method become more severe in this case (Sec. 52).

If we proceed by dropping terms in formula (61) and finally
choosing a new variable
we then arrive to the following perfectly accurate equations

$$
\begin{gathered}
\quad \therefore \ddots_{4}=-J_{4} \\
r_{i 11} \quad \exists_{k} \quad \therefore_{1}, f_{k} .
\end{gathered}
$$

In this case, Cowells method is improved to the maximum and we obtain, as we easily see from equation (57),

$$
r_{1}-I_{1} .
$$

i.e., in other words, the method of quadratures.
59. Laplace's Method of Quadratures and Related Methods of the Numerical

Integration of Equations
The methode of numerical integrations, considered in the previous sections, can be modified in many different ways. Let us consider the methods given in Sec. 49 for the integration of the following equation

$$
\therefore \quad 111,11
$$

These methods are based on the calculation of the difference

$$
\left.\therefore 1 \quad x_{i!1}-x_{1} \quad x_{1} 1_{1,1}\right)-x(1,)
$$

by means of the successive differences of the function

$$
1(1) \quad 1 .: 11.11
$$

Let us represent this difference in the following way
or, using again equation (6),

This formula leads to a method of integration, which is similar to Cowell's method in thesense that it avoids the extrarolation in the evaluatio, $f f_{k+1}$, ... . Summing equation (67) from $k=0$ to $k=n-1$ yields

On applying this formula to the case, when the right-hand side of equation (6) does not depend on $x$, we obtain

$$
x_{4}-i_{v}-\int_{0}^{i x} f(t) d t \cdot \int_{d}^{0} f(t) a t
$$

Consequently

This formula is in many cases not convenient because it expresses the integral not only in terms of the values of the functions, $f_{0}, f_{1}, \ldots$, $f_{n-1}, f_{n}$, but also in terms of the values $f_{-1}, f_{-2}, \ldots$, which correspond to values of the argument lying outside the limits of integration. This inconvenience can be avoided if we use the following relations


Finally, we obtain the following formula of quadratures

which has been obtained by Laplace. This formula had been widely used till the forties of the 19 th century in the calculation of perturbations. Láer on, it was replaced by Gauss method discussed in Sec. 56. 60. The coefficients of the Formulae of Numerical Integracion

Tn conclusion, we point out another way of calculating the coefficients of the formulae which are applied in the numerical integration of equations. When the form of a formula of numerical integration is established, we can apply this formuia to a properly chosen specific case and obtain its coefficients. For example, if we consider the following equation
and apply formula (10) to its solution $x=e^{t}$. Choosing $t_{c}=0$, the differences of the function $\psi=e^{t}$ will be given by

$$
\begin{aligned}
& \left.\because_{1}^{\prime}, 1=e^{16114}-e^{A-} \quad e^{2 \omega}\left(e^{4 \pi}-\right)^{-1}\right) \\
& \dot{r}_{k}^{*} \cdot\left(e^{* *}-e^{14}+1\right)\left(e^{*}-1\right)=e^{(t-1)}\left(e^{*}-1\right)^{*}
\end{aligned}
$$

$$
\begin{aligned}
& \text { • . . . . . . . . . . }
\end{aligned}
$$

or, assuming that

$$
u e^{\frac{\pi}{3}}-e^{-\frac{\omega}{\gamma}}
$$

we will obtain the following expressions

$$
\begin{aligned}
& \because, 1\left(1+\frac{1}{2}\right)_{4}
\end{aligned}
$$

Consequently, we will have for the function $f=$ we $^{t}$

$$
\begin{aligned}
& \text { f.. } \quad \text {, }: \cdots=: \ldots 1 \\
& \therefore \quad \vdots \quad 4 e^{(=1} \cdot h_{2}^{a}
\end{aligned}
$$

where

Substituting these expressions into formula (10), we finally obtain the following identity

$$
\frac{\Delta}{u^{\prime \prime}\left(1: \frac{u^{\prime}}{\because}\right.} \quad 1 \quad 1, a^{2} \cdot{ }_{i!1}^{11} u^{4} \quad \frac{191}{60480} u^{\circ}+\ldots .
$$

Since

$$
u \quad \min \left(\begin{array}{lllll}
1 \\
2 & u & \vdots & 1 & \vdots \\
4
\end{array}\right)
$$

then this identity may be written in the following way

Thus, in order to obtain the coefficients of formula (10), it is sufficient to expand the function on the left-hand side in power series. Similarly, the coefficients of formulae (11), (19) and (26) can be found using the following identities

$$
\begin{aligned}
& 4
\end{aligned}
$$

the proof of which is relatively simple.

## 61. Introduction

At the beginning of the previous chapter, we pointed out that the numerical integration of differential equations has been considered as a method for the calculation of perturbations. It has never been applied to the calculation of unperturbed coordinates. The only exception is the approach suggested by Krueger for the evaluation of a true anomaly (Sec. 62). Only when numerical integration was applied to the study of perturbed coordinates (and not perturbations) by the initiative of Cowell, the possibility of appiying this method to the calculations of urperturbed coordinates was frequently considered.

In this chapter, we shall considez the calculations, by numerical integration, of the orbital coordinates $r$ and $v$ (or $\mathcal{\xi}$ and $q$ ), and the equivalent heliocentric ccordinates $x, y$ and $z$ used in the calculations of the ephemeride.

In the calculation of a more or less long ephemeride, the method considered in this chapter is usually preferred than the calculation of the coordinates of a star by the conventional methods, stuaied in the first volume.
62. The calculation of coordinates defining the position of a luminary in
an orbit
It is easier to apply the following equations to the simultaneous calculation of a true anomaly $v$ and radius vector $r$

$$
\begin{align*}
& d i \quad \frac{1 \cdot n}{r^{2}}  \tag{111}\\
& d=\frac{p}{1+\cdots \cdots:}
\end{align*}
$$

In order to obtain a table for the values of $r$ and $v$ for the moments $t_{0}, t_{0}+w, t_{0}+2 w, \ldots, i c$ is sufficient to calculate these quantities for the moment $t_{0}$ by conventional methods (Volume $I$, Chapter III).

It may sometimes happen that the integration of the following equation
which is easily obtained from equations (1) and (2), rather than equation (1), is more useful.

Krueger ${ }^{(1)}$ suggested the following particularly convenient order of calculation. First of $a 11$, the coordinates $v$ and $r$ are found by Kepler's formulae for the first three moments. Then, a table of the approximate values of $r$ is constructed by extrapolation. This gives the possibility of calculating the right-hand side of equation (1), which after integration yields $v$. With these values of $v$, the values of $r$.are reevaluated. Repeating the integration, we obtain more accurate values for v (cf Sec. 57).

If we have to only calculate the values of the radius vector, we can apply Legandre's formula (Sec. 5).

$$
\frac{a!}{a r} \quad \because \cdot\left(\begin{array}{ll}
1 & 1 \\
r & a
\end{array}\right) .
$$

Integrating this equation, we obtain a table for the values of $r$. We then evaluate $v$ from equation (1) by means of a simple quadrature. Of course, we can evaluate $v$ using equation (2) if we have already obtained r. However, the calculation of the true anomaly by integrating equation (1) leads to mpre accurate results and requires less amount of
(1) A. Krueger, Die Wiederkehr des Olbers'sehen Cometen 1887, Astr. Nachr. 117, 1887.
work. Hence, it is generally recommended to follow this procedure.
In order to start integrating equation (3), it is possible to calculate the values of $r$ for two moments, for example $t_{0}-w$ and $t_{0}$, or $t_{0}-\frac{W}{2}$ and $t_{o}+\frac{W}{2}$. It is also possible to calculate $r$ only for one moment, but it is then necessary to calculate for this moment the derivative

$$
\begin{equation*}
d r \frac{\sin i n}{1 \beta} \frac{\alpha \tan 3, i n n}{1 a} \tag{1}
\end{equation*}
$$

The method of calculation of $r$ and $v$ for a series of different moments, which we have suggested above, is particularly useful when the calculation of these values by conventional formulae, is complicited, and in particular, for those orbits whose eccentricities slightly differ from unity.

In order to calculate the rectangular orbital coordinates $\xi$ and $\%$ it is possible to use the following differential equations

It is, however, simpler to express the coor-inates $\xi$ and $\eta$ in terms of the quantity $\sigma$ (Vo1. I, Sec. 23), defined by

$$
\begin{equation*}
\frac{\pi}{6!} \frac{1}{1!} 4^{+}+\frac{1-:!}{1+:!} \tag{i.}
\end{equation*}
$$

where

$$
4 \quad a(1-1) \quad \cdot \quad!11 \quad \text { e) }
$$

and subject to the condition that $\sigma=0$ in the moment, when the luminary passes by the perihelion. The coordinates $\xi$ and $\eta$ are evaluated by the formulae (loc. cit.)'

$$
\vdots v\left(1-z^{\prime}\right) \quad 4=24=\sqrt{ } 1-i v 1-\therefore .
$$

In conclusion, we point out that the numerical integration of equation (1) or equation (3) may be applied for correcting or checking the c alculated values of $v$ or $r$. For example, taking as a first approximation the value of $v_{1}$ given in table XV (Vol. I) with an accuracy of 0.00005 , and integrating equation (1), we obtain a value of $v$ with a larger number of decimal places (Sec. 57).
63. An example of calculating the orbital coordinates using the numerical

## integration

Let the orbit of a comet be defined by the elements

It is required to calculace $r$ and $v$ for the moments $t_{o}+k w$, where
and when time is evaluared starting from the moment the comet passes by the perihelion.

This problem can be solved by several methods

## Krueger's method:

First of all, we use the conventional formulae (Vol. I, Sec. 23) to calculate the followin values of $v$ and $r$

| 1 $t_{0} \ldots \ldots$ | 1 | 101.00:000 | r | 1.3750 |
| :---: | :---: | :---: | :---: | :---: |
| !, |  | 101.05 .5 |  | 1.8198: |
| $6_{10}: \prime \prime$ |  | 111212647 |  | 1.4:9 |

In equation (1), the function
ノ ukipr
can be represented by

$$
f-2.111215 r \quad-|0.3245321| r \div
$$

since the Gaussian constant, expressed in degrees,equals

$$
k=0.0 .1 \times 5 \text { 6.075, } \quad \log k \quad 4.49 .17010,10
$$

In the present case, $w=2$ and

The calculation of the function $f$ for the indicated vaiues of $r$ and for the two extrapolated values (printed in italics) are shown in the following table

TABLE A


Since we integrate equation (1) using the formula

$$
\because \quad!\quad \because_{n} \quad f_{n}^{1} \quad 11 f_{n} \quad \cdots
$$

we imnediately write in table $B$ the aritrmetic averages of the values of $f$, obtained in table $A$.

TABLE B
ORIGINAL PAGE: OE POOR QUAL."


## TABLE C

| $t$ | \| +-rios ${ }^{\text {a }}$ | $r$ | idifferences | $r^{\prime}$ | $r$ | , |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| f., |  | 119 | 11. |  |  |  |
|  |  |  | 1 . 61.5 |  |  |  |
| $t_{1}$ |  | 119 | - 1: |  |  | i 119 SH |
|  |  |  | - $30: 3$ |  |  |  |
| $!$ | 0780.20 | 117111 | $-15$ | 064026 | $110.270{ }^{\text {a }}$ | 0177087 |
|  |  |  | 3事" |  |  |  |
| $t$ |  | I, M M 41 |  | 01,665.3: | (1)1120ts | 013/3191 |

The initial value of the column of sums is defined by the condition
which yields in the present case

The details of the calculation are given in tables $B$ and $C$ in the following way.

1. Using the extrapolated value Red $=+335$, we obtain $v_{2}=103^{\circ} .12439$.
2. Using this value for $v_{2}$, we calculate $r_{2}$ by means QRIGINAI PAGE IS noting that in the present case $p=1.147088$.
3. The obtained valuc $r_{2}$ allows us to find the final value $f_{2}=0.97687$. This shows that the value $f_{3}=0.99766$ used in the integration may not be changed.
4. We find $r_{3}$ by extrapolation and then evaluate $f_{3}$.
5. We write in table $B$ the corresponding value $f_{5 / 2}=0.95740$.

Taking the extrapolated correction Red $=4314$, we calculate $\mathrm{v}_{3}$ 。
In conclusion, we note that the accepted accuracy in calculating $r$
does not provide us with a fifth decimal place number for the true wily. For this purpose, it is necessary to find a value of $r$ consisting of seven significant figures. However, such an accuracy is never expected to be necessary.

## The Second method

First of all, we integrate equation (5) using the following initial values

$$
t_{1} \quad \text { (in) } 41(6), \quad=12-1.210320 .
$$

Since varies very slowly, th2 interval w can be increased and taken as $w=4^{d}$, We can either calculate $\sum$ and $\eta$, or find $r$ and $v$ by using the following formulae

$$
r=q(1+e s), \quad n=\sqrt{\frac{1-8}{1-5 z^{2}}} .
$$

We can also evaluate $r$ by using the first of these formulae and then find $v$ by integrating equation (1). This requires an evaluation of a quadrature (Sec. 56 ), since the right-hand side of this equation will be already known for all the moments under consideration.

## The third method

Using the initial values
we integrate equation (3) taking $w=4^{d}$. The true anomaly may then be found similarly to the previous method.
64. Calculation of the ephemeride by means of the numerical integration of the equations of motion:

The rectangular coordinates of the luminary, $x, y$ and $z_{1}$ required for the calculation of the ephemeride may be obtained by the numerical integration of the following equations

$$
\frac{\pi^{\prime}}{d r^{2}}+\frac{n^{2} x}{r^{3}} \quad 1, \quad d i \quad \frac{t^{2} y}{r}, \quad \frac{r^{2}}{d!} \frac{x^{2}}{r^{2}}-1
$$

We start integrating either by calculating the coordiaates

$$
\left(!_{-1}, v_{-1}, z_{-1}, \quad \text { I, , Yo, }, 1\right.
$$

for moments $\mathrm{f}_{\mathrm{o}}-\mathrm{w}$ and $\mathrm{t}_{\mathrm{o}}$, or by calculating the qurntities

$$
\left(n_{0}, l_{n}^{\prime}, 1\right) \quad\left(\lambda_{0}, y_{0}^{\prime}, z_{0}^{\prime}\right)
$$

defining the position and velocity of the luminary in the moment $t_{0}$. Wo dan apply for this purpose the convencional Kepler equations. For example, if the motion proceeds by an allipse of moderate eccentricity, we may then apply the following formulae (Vol. 1):

$$
\begin{aligned}
& \text { - } 252 \text { - }
\end{aligned}
$$

$$
\begin{aligned}
& \because \quad: /{ }^{\prime} \text { fror } \quad \therefore \quad \therefore 1, \cdots i .
\end{aligned}
$$

$$
\begin{align*}
& \begin{array}{c} 
\\
1 \\
\vdots \\
\vdots \\
1 \\
1
\end{array}
\end{align*}
$$

where the eccentric anomaiy $E$ is defined by the condition

$$
A-C \sin E \quad . \hbar
$$

The following equations may be applied to control the calculation

$$
\begin{aligned}
& r=-\boldsymbol{y}^{\prime} \boldsymbol{x}^{2}: y^{2}:=-a(1 \quad \text { co, } 1 \text {. }
\end{aligned}
$$

The unperturbed ephemeride is usually calculated for a short time interval. The initial moment $t_{o}$ is chosen in the middle of this interval. Under these conditions, the method of quadratures has no advantages ovc Cowell's method since the number of integrals involved in the integration is not large. Both methods can equally be recommended.

For the integration of equations (6). we have to calculate the functions

It is easiertcalculate this using a table which gives the values of $w^{2} k^{2} r^{-3}$ that correspond to diffeient valies of the argument $1^{2}=x^{2}+$ $y^{2}+z^{2}$. Such a table is given in volume $I$. A more detailed table has been given by Comrie (Comrie, Planetary Co-ordinates for the years 1800-1940, London 1933). On using this table, one should tate into ronsideration that


The evaluation of the heliocentric ephemeride of a planet by means of this method has already been given in volume $I$.

We note that it is more useful to calculate by means of formulae (7) several position of the planet, rather than to calculace one or two positions that are necessarily sufficient for solving system (6). This will simplify the integration and will also render possible a good control.
65. Other Methods for Calculating the Ephemeride by Numerical Integration

Let us assume the coordinates, $x_{0}, y_{0}$ and $z_{0}$, and velocity components $x_{o}^{\prime}, y_{o}^{\prime}$ and $z_{o}^{\prime}$, of a luminary are known for some moment $t_{o}$. It was shown in volume $I$ that the coordinates $x, y$ and $z$ of this planet at any particular time could be expressed in terms of the above mentioned initial values by the following equations

$$
\begin{aligned}
& \therefore \quad i x, j x \text {, } \\
& \because \quad i \quad: \quad \text { (in } \\
& \therefore \text { - } \therefore \text { r } 1 \text {. }
\end{aligned}
$$

The functions $F$ and $G$ are given by the following series-expansions
where $\theta=k\left(t-t_{0}\right)$, and $r_{0}$ and $r_{0}^{\prime}$ are given by

$$
r_{0} \therefore r_{0}: y_{n} \quad \because, \quad r_{1,1}-\lambda_{1,} x_{1,} \cdot y_{1,}, z_{1,} z_{0}
$$

The dashes here denote, the derivatives with respect to $\theta=$

- $t_{0}$ ), so that

$$
x^{\prime}=\begin{aligned}
& 1 d x \\
& k d t^{\prime}
\end{aligned} \quad y=\begin{aligned}
& 1 t^{\prime} y \\
& k d t^{\prime}
\end{aligned} . \quad z^{\prime}=\begin{aligned}
& 1 d z \\
& k d i^{\prime}
\end{aligned} \quad r^{\prime}=\begin{aligned}
& 1 d r \\
& k d t^{\circ}
\end{aligned}
$$

The evaluation of the heliocentric coordinates $x, y$ and $z$
required to obtain the ephemeride can be made by means of formulae (8).
The functions $F$ and $G$ can be easily calculated by means of the series (9)
for the near moments $t$. For further points, it is more convenient
to apply the numerical integration than to use the final expressions of these functions, which has been given in volume $I$. By differentiating equations (8) twice, and noting that

$$
\begin{aligned}
& d x \\
& d:
\end{aligned}=-k^{2} x r^{\prime}, \quad \frac{d^{2} y}{d l^{2}}=-k y r^{3}, \quad \frac{d^{2} z}{d t^{2}}=-4^{2} z r^{3},
$$

we obtain
where

$$
\begin{equation*}
r^{2} r_{11}^{2} r+r_{1} \theta_{i}^{2}: Q r_{1} r_{i} \tag{11}
\end{equation*}
$$

The system of equations (10), in which $r$ is defined by equation (11), is easily integrated by means of the methods described in the previous chapter.

Equations (9) show that

$$
\therefore \quad 1,\binom{d!}{d!}_{-1} \quad 11 . \quad(i,=1), \quad\binom{d r i}{d i}_{0} \quad n i
$$

It is thus convenient to define the integral curve $b$ the initial position and the initial velocity.

There is ancther method for calculating the rectangular heliocentric coordinates. Tt consists in taking formulae
which express the condition that the three positions of a luminary are 1.jcated in one plane which passes through the centre of sun, and considering that in these formulae $\left(x_{1}, y_{1}, z_{i}\right)$ and $\left(x_{2}, y_{2}, z_{2}\right)$ are the coordinates of the 1 uminary in the moments $t_{1}$ and $t_{2}$, and ( $x, y, z$ ) the unknown coordinates corresponding to moment $t$. In this case, the quantities $n_{1}$ and $n_{2}$ are considered as functions of time that satisfy the following equations ${ }^{(1)}$ :
where

$$
\begin{equation*}
r=r n_{1}^{2}+r+n-: r_{1} r, n, n \tag{114}
\end{equation*}
$$

The initial values of the functions $n_{1}(t)$ and $n_{2}(t)$ are suitably taken as

$$
\begin{array}{ll}
n_{1}\left(t_{1}\right)=1, & n_{1}\left(t_{2}\right)=0 \\
n_{2}\left(f_{1}\right)=c, & n_{n}\left(f_{2}\right)=1 .
\end{array}
$$

The integration of systems (10) or (13) is simpler than the direct integration of the equations of motions which have been considered in the previous section. Each of these systems involves only two equations and, not three. However, this simplification is considerably mullified

[^0]by the difficulty of equations (11) and (14) and by the indispensible returns to equations (8) or (12) in order to obtain the unknown coordinates.

The moments $t_{1}$ and $t_{2}$ are suitably chosen not far from the middle of the ephemeride.

## CALCULATION OF THE PERTURBATIONS IN THE ELEMENTS

## 66. General Considerations:

The differential equations that determine the osculating elements in the general case of an arbitrary perturbing acceleration have already been given in Sec. 12. For the timebeing, we shall abandon the orbits with eccentricities near unity. We introduce, instead of the eccentricity, angle $4=$ are sin e. Equations (37) and (38) of Ser. 12 will then give

$$
\begin{aligned}
& \frac{d i}{d l}=r \cos u w^{\prime} \\
& \frac{d \mu}{d t}=r \sin u \operatorname{cosec} i W^{\prime} \\
& \frac{d}{d!}=a \cos \varphi \sin u S^{\prime}+a \cos \%(\cos v+\cos t) T^{\prime} \\
& \left.d \pi=-p \operatorname{cosec} \because \operatorname{con} v S^{\prime} \cdot+\operatorname{cosec} ;(r+p) \operatorname{sir} v T^{\prime}+1 \boldsymbol{q}_{2}^{i} r \sin u W^{\prime}\right\}(1) \\
& \frac{d n}{d l}=-\frac{3 r}{V a} \cdot \sin ; \sin u S^{\prime}-\begin{array}{cc}
3 k & p \\
V_{a} & r
\end{array} \\
& \frac{d \varepsilon}{d t}=-\left(2 \cos p t-p \operatorname{tg} \xi_{2}^{\cos }{ }^{\prime}\right)_{j}^{\prime} S^{\prime} \operatorname{tg}_{2}^{\prime}(r+p) \sin u T^{\prime}+ \\
& \text { F- } \underline{g}_{\underset{2}{i}}^{i} r \sin u W^{\prime} .
\end{aligned}
$$

These equations form a system of coupled first-order equations. They can be numerically integrated by any of the methods considered in Chapter VI. In astronomical calculations, the method of quadratures, considered in detail in Secs. 49-51, is usually applied. In the present case, the application of the method of quadratures is partiaularly easy. Tn the right-hand sides of equations (1), the coefficients of the perturbing masses are very small. Therefore, they can be sufficiently accurately
evaluated by means of the approximate values of the unknown elements $i, \Omega, \varphi, \pi, n$ and $\in$.

We denote by $t_{0}$ the moment for which the values of the osculating elements $i_{0}, \Omega_{0}, \varphi_{0}, \Pi_{0}, n_{0}, E_{0}$ and $\epsilon_{0}$ are known, or in other words, the epoch of osculation. Let us assume that

$$
1-i_{0}+1 i, \quad u=U_{0} \vdots 1 U_{1} . . \quad:
$$

In this case, quantities $\Delta i, \Delta \Omega, \ldots$ will be defined by the differential equations

as well as the initial conditions $\Delta i=0, \Delta \Omega=0, \ldots$ for $t=\mathcal{C}$ We choose a definite interval of integration, say $w$, and denote the epoch of osculation $t_{0}$ by $t_{0}-\frac{w}{2}$. We try to find the values $\Delta i, \Delta \Omega, \ldots$, and, consequently, the osculating elements for the moments

$$
t \cdot t
$$

where $k$ is an integer.
Since; on one hand, we hove to evaluate the right-hand side of equation (1') muitiplied by $w$ in the course of integration of these equations, and on the other hand, it is preferbale to obtain the quantities $\Delta_{i}, \Delta \Omega, \Delta \varphi, \ldots$ we shall put

$$
\begin{align*}
& i j=r \cos \| W \\
& \text { } 2,1=r \sin u \text { cosec } i W \\
& \lambda_{i}=a \cos \% \sin u S+a \cos \%(\cos u-1-\cos E 17 \\
& i \pi=-p \operatorname{cosic} ; \cos u S \cdot j \cdot \operatorname{cosec} \dot{\gamma}(r+p) \sin \cup \gamma+- \\
& +\operatorname{tg}{\underset{2}{i} r \sin u W}^{r} \\
& u^{\prime} \dot{\jmath} n=-\frac{3 k w}{V_{a}} \sin ? \sin v S-\begin{array}{l}
3 k u^{\prime \prime} p \\
l_{a}
\end{array} T
\end{align*}
$$

$$
\begin{aligned}
& f-\operatorname{tg}_{2}^{i} r \sin u w .
\end{aligned}
$$

Taking the values of $S^{\prime}, T^{\prime}$ and $W^{\prime}$, obtained in Sec. 11 , into consideration, we obtain

$$
\begin{align*}
& T=\begin{array}{c}
w \\
k V^{\prime} p_{\text {arc } 1 "} T, \quad W={ }_{k} V_{p \text { arc } I^{\prime \prime}}^{W} \quad, ~
\end{array} \tag{3}
\end{align*}
$$

where $k$ is the Gaussian constant expressed in seconds of arc.
We shall now apply the method of quadratures, which has been considered in detail in Sec. 50. We shaol use, in particular, equations (17 bis) and (18) of that section. Taking for $f$ each of the functions $6<, S \Omega, \ldots$, we obtain the following formulae
where. $x_{u}$ dendetes the corresponding values of $\Delta i, \Delta \Omega, \ldots$ for the moment $t=t_{0}+n w$.

The average longitude is calculated by the following equation (sec. 12).

ORIGINAL PAGE IS
where its unperturbed value equals $\epsilon_{0}+\mathrm{a}_{0}\left(t-\tau_{0}\right)$. The perturbation of the average longitude wili then de given by

$$
D_{1}=; \quad\left|z_{0}: \pi_{1}\left(t-i_{0}\right)\right|-1=\therefore \int_{0}^{!}!n-n_{0} i u
$$

Consequently, the calculation of $\Delta \lambda$ is reduced to che evaluation of the quantity

$$
J^{\prime} \lambda=: \int_{1}^{1}(n-n,) d t
$$

which satisfies the differential equation

$$
\begin{equation*}
\frac{\left.d s^{\prime}\right\rangle}{d t^{2}}=\frac{d n}{a t} \tag{5}
\end{equation*}
$$

as we11 as the initial conditions

$$
s^{\prime} x=0, \quad \frac{d s^{\prime} \lambda}{d f}=0
$$

for $t=\tau_{0}$
The right-hand side of equation (5) is defined by the second to last of equations (1). In order to integrate this equation, we apply formulae (37 bis), (36) and (39) of Chapter VI. This yields the the following formulae
where $x_{n}$ denotes the quantity $\Delta^{\prime} \lambda$ for the moment $t=t_{0}+n w$. After obtaining the value of $\Delta^{\prime} \lambda$, we calculate the perturbation of the average longitude $\lambda$ for the moment $t$ using the following formula

Once we obtain the perturbed values

$$
i=i_{0}+\Delta i, \quad U=u_{0}: \Delta \underline{1}, .
$$

of all the elements for this moent, we can calculate the position of the lumfnary by the conventional formulae of the elliptic motion:

$$
\begin{align*}
& i \because-e \sin E-\lambda-\pi  \tag{x}\\
& r \sin v=a \cos \xi \sin t \\
& r \cos u=a(\cos E-\sin \psi),
\end{align*}
$$

where the semimajor axis is obtained by means of the following relation

$$
\left.\begin{array}{lll} 
& \vdots & \vdots
\end{array}\right)^{\prime}
$$

in which

$$
\text { het } 3 \text { Buculomit. }
$$

if $n$ is expressed in seconds of arc, and

$$
1 \because 0^{3} \quad 9.210: 7141
$$

if $n$ is expressed in fractions of a second.
The calculation of the rectangular equatorial coordinates, required when the perturbed position of the luminary is to be compared with the experimental value, is done by the following formulae:

$$
\left.\begin{align*}
& x-r \sin a \sin \left(A^{\prime}: v\right) \\
& y=r \sin b \operatorname{cin}\left(B^{\prime}\right.  \tag{113}\\
& z-r \sin r \sin \left(C^{\prime}\right. \\
& \hline- \\
& v)
\end{align*} \right\rvert\,
$$

where the Gaussian constants $a, b$ and $c$ as well as the quantities

$$
A^{\prime}=A-+\cdots, \quad B^{\prime}=A: \cdots, \quad \because^{\prime}-\cdot C \cdot \cdots
$$

are defined by

$$
\begin{align*}
& \sin a \sin A=-\cos : \\
& \text { iin } a \cos A=-\cos i \sin 9 \\
& \sin h \sin B+\ldots \sin 0 \cos \varepsilon  \tag{il}\\
& \sin b \cos 3 \quad \text { ros } i \cos 2 \cos \varepsilon \quad \sin i \sin = \\
& \sin c \sin C=\sin 19 \sin \varepsilon
\end{align*}
$$

where $\epsilon$ denotes the slope of the ecliptic with respect to the equator.
If we thus disregard the simpie interrations, we see that the calculation of the perturbed values of the elements leads to the calculation of the functions (2), which depend on the componsnts of the perturbing acceleration, $\underline{S} ; \underline{T}$ and $\underline{W}$ as well as the coordinates $r$ and $v$ of the perturbing body. We have already considered the calculation of the latter quantities, which can be done by means of formulae (7) and (8). We still have to find the most convenient method far the calculation of the quantities $\subseteq$, $T$ and $W$, which depend on the components of the perturbing acceleration.
67. Calculation of the components of the perturbing acceleration

Let us consider the following zuniilary coordinate system. We take the axis $\mathcal{Z}$ along the radius vectus $\approx$. rhe perturbed body in the direction of increase of $r$, the axis $\eta$ along the perpendicular to the radius vector in the orbital plane in the direction of increase of the true anomaly, and the axis $\zeta$ along the normal to the orbital. plane so that the coordinate system becomes right-handed. We denote by $\underline{S}, \underline{T}$, and $\underline{W}$ the components of the perturbed acceleration in the directions $\xi, \eta$ and $\mathcal{Z} \mathcal{Z}$. We can calculate these quartities using different approaches. The most widely used approach is the following.

Let us denote by $m_{1}$ and $\xi_{y,}, 7$, and $\mathcal{Z}_{;}$the mess and coordinates of one of the perturbing planets. Since the coordinates of the perturbed
planet are evidently equal to ( $\mathrm{r}, 0,0$ ), then the general expression for the component of perturbing acceleration along any axis, namely (sec. 3),
yields in the present ase
where

$$
s_{i}^{\prime}-:\left(\xi_{i}-r_{i}^{\prime}: r_{1}^{\prime \prime} \cdots \cdots,\right.
$$

and $r_{1}$ is the radius vector of the perturbing planet. Taking equation (3) into corsideration, and putting

$$
\begin{equation*}
\kappa_{i} \cdot{ }_{p}^{\prime, m_{1}}\left(\frac{1}{5}-1,\right. \tag{12}
\end{equation*}
$$

we obtain

$$
\dot{S}_{1}=\kappa_{1} i_{1}-\begin{gather*}
u^{\prime \prime \prime} m_{1} \\
1 \rho
\end{gather*} \frac{1}{j_{1}}, \gamma_{1} \quad \kappa_{1} \gamma_{11}, \quad \| \cdots \kappa_{1} i_{1}
$$

If we consider perturbations caused by other planets having masses $\mathrm{m}_{2}, \ldots$ and coordinates $\left(\xi_{i}, \mathcal{Z}_{2}, \tau_{2}\right), \ldots$. we can then calculate by formulae, similar to equations (12) and (13), the corresponding components $\left(S_{2}, T_{2}, W_{2}\right), \ldots$ and then adopt in equations (2) tnat

$$
\begin{aligned}
& \therefore-S_{1}-\therefore S_{n} \\
& r \quad r_{1} \cdots r_{n}
\end{aligned}
$$

$$
11 \cdot 11 ; 11
$$

$$
\begin{aligned}
& 11:-m_{1}\left(\begin{array}{cc}
1 & -1 \\
1 & 1 \\
1 & 1
\end{array}\right) \text {. }
\end{aligned}
$$

We still have to show how
the coordinates of the perturbing planets, ( $\left.\xi_{1}, \eta_{1}, \tau_{1}\right)$
can be found. We can find in the astronomical annual the ecliptic $h \in 1$ locentric coordinates, namely the radius vector $r_{1}$, longitude
 $\ell_{1}$ and latitude $b_{1}$, for each of the large planets. These quantities

Figure 10 ase given with high accuracy in Comrie's table for most of the interesting cases.

Let us denote by $L_{1}$ and $B_{1}$ the longitude and latitude of the perturbing planet relative to the orbital plane of the perturbed planet. We shall consider that the longitude $L_{1}$ is calculated from the ascending node $\Omega$ of this orbit (Fig. 10).

In the spherical triangle formed by the position of planet $P_{1}$, the pole of the ecli, tic $E$, and the pole of the orbit 0 , the angles at the apexes 0 and $E$ are respectively equal to $90-L$, and $90-\epsilon_{1}-S$. Hence, in the evaluaticn of $L_{1}$ and $B_{1}$, we may apply the following relations

Denoting by $u$ the argument of latitude of the perturbed planet, the quantities $L_{1}-v$ and $B_{1}$ wili be the spherical coordinates of the perturbing planet that correspond to the coordinate system $\xi, \eta, \zeta$ Therefore

$$
\begin{aligned}
& \because \quad .1 \text {, 吅 } \therefore
\end{aligned}
$$

Formulue (14) and (15) constitute the solution of the problem of calculating the coordinates $\zeta_{1}, ₹_{1}$ and $\zeta_{1}$ in terms of the given quantities $r_{1}, \ell_{1}$ and $b_{1}$.

At an earlier date when calculating machines were not widely used, the separation distance between the planets, $\Delta_{i}$, was calculated not by means of formula
but by means of the following equations

$$
\begin{aligned}
A_{1} \cos \varphi_{1} \cos O_{1} & =-a_{1}-r \\
\Delta_{1} \cos \theta_{1} \operatorname{sini} O_{1} & =r_{1} \\
I_{1} \sin \theta_{1} & =a_{1} .
\end{aligned}
$$

wher $\quad$ i. are auxiliary quantities, unnecessary to calculate further.
68. Another Method for Calculating the Components of the Perturbing

## Acceleration

We shall here consider the case, when we need so calculate the perturbations that occur after a few rotations of the luminary, provided that these pertu\&bations are small. In this case, we can calculate in another way the coordinates of the perturbing planet, which we shall call Jupiter.

We shall assume that the motion of Jupiter and the motion of the perturbed luninary proceed in the invariable pianes defined by $t_{\text {u }} e$ elements $i_{1}, \Omega_{1}$ and $i_{2}, \Omega_{2}$. We first consider the spherical
triangle $\Omega \Omega_{1} N$, (Fig. 11), formed by the ascending nodes $\Omega$ and $\Omega<$, of the orbit under consideration relative to the ecliptic, and

ingure 11 the ascending node $N$ of the orbit of Jupiter relative to the orbit of the perturbed body. In order to evaluate the angle $J$ betwien the orbit and the arcs $\Omega \mathrm{N}$ and $\Omega, N$, whish will be denoted by $N$ and $N_{1}$, we shall apply the the following formulae, which can easily be obtained from the
study of the triangle under consiaeration:

Let us now consider a new coordinate system $\xi^{\prime}, \eta^{\prime}, \tau^{\prime \prime}$, which
 directed towaras the orbital point $B$, removed from the node by an angle $\Omega \mathrm{R}=\beta$, and the axis $\eta^{\prime}$ towara a point, removed from the nocie an angle $\beta+90^{\circ}$. The coordinates of Jupiter in the new coordinate system can be calculated by means of the conventional formulae (formulae (8) in Sec. 10), assuming in these Formulat that the
longitude of the node equals to

$$
B . V^{\prime}-\lambda \quad \ddots
$$

and the argument of the latitude equals to

$$
N P_{1}-H_{1}-N_{1}-\underline{U}_{1}
$$

where $\lambda_{1}$ is the longitude of Jupiter in the orbit. This yields

In analogy with the Gausian constants, we introduce the following quantifles

We then finally obtain

In order to obtain the required unknown coordinates $\xi$ nnd $\underset{\vdots}{ }$ we rotate the coordinate system around the axis $\zeta$. This yields

We can now samplify the calculations performed by means of equations (18), (19) by an appiopriate choice of the arbitrary angle $\beta$. The most
simple formulae ${ }^{(1)}$ are obtained in the following cases
where

Since, in the second case, argle $\beta=u$ depends on the position of the perturbed luminary, constants $A, A^{\prime}, \ldots$ will then also depend on the coordinates of the perturbed 1 -minary. The corresponding formulae are suitably applied only when it is required to calculate the functions (2), subject to integration, of only a few orbital points (cf. Sec. 69).

The appifation of formulae (18) and (19) rather than the conventional formulae (14) and (15) is useful only if the elements $i, S$,. $i_{1}, \Omega_{1}$ are approximately constant during a considerable interval of time, and, moreover, when the longitudes $\lambda_{1}, \lambda_{2}, \ldots$ of the perturbing planets in the orbit are known.

The influence of the small variations di, $d \Omega, \ldots$ on the constants $A, A^{\prime}, \ldots$ can evidently be evaluated by differentiation. We shall not consider here the derivation of the corresponding formulae. 69. The tabulation of thecoefficients

The calculation of the expressions given by equations (2) is reduced to the evaluation of the quantities $S, T$ and $W$, which have been considered in detail in the previous two sections, as well as the coefficients
(1) These formulae were given by Merton for the case $=u$ : G. Merton, The periodic comet Grigg (1902 II) = Skjellerup (1922 I) (1902 to 1927), Memoirs of the R. Astr. Society; 64, Part III, 1927. The formulae that correspond to the case $=\overrightarrow{a r e}$ given in: N.I. Idel'son, La comete d'Enke en 1924-1934, Proceedings of the Principle astronomical observatory in Polkov (Izvestija Giavnoj astronomiceskoj observatnpii v Pulkove) vol. XV, I, 1935.
obtained by muliiplying these quantities. The calculation of these coefficients is reduced to the evaluation of the following quantities

| $A=\begin{array}{r} r \\ g \end{array} \because$ | $\therefore \quad \therefore . \\|$ |
| :---: | :---: |
| \% rosesin ${ }^{\prime \prime}$. | П - いV: |
|  |  |
| () - s'11: = ! ! |  |

which are functions of only $\varphi$ and $M$. Here, we have singled out the multipliers which depenci on $a, i$ and $w$, since their evaluation is simple enough.

The calculation will be significantly simplified by constructing tables for these coefficients. In order to reduce the volume of these tible, wa shali transform equations (1) by introducing a new independent variable $M$, instead of the variable $\mathfrak{c}$, using the relation

$$
\therefore \quad \therefore 1 \quad \ldots
$$

In this case, it is possible to calculate the quantities (20) for a few round values of $M$. Tnstead of tables of two arguments, $M$ and we shal.1 chen have a table of a single argument, $\varphi$. Such tables have been constructed by Crommelinin (1) for values of the argu'nent $e=\sin \varphi$ varying fiom 0.37 to 0.84 by intervals of 001 , i.e. corresponding to the orbits of short-periodic comets. The first table gives the values of the cefficients, which slightly differ in form $A, B, \ldots$, for $M=0^{\circ}, 7.5^{\circ}, 15^{\circ}, 22.5^{\circ}$, ... with five figures. The second table
(1) A.C.D. Crommelin, Tables for facilitating the computation of the perturbations of periodic comets by the planets, Memoirs of the R. Astr. Society, 64, Part V, 1929.
gives the logarithms of these coefficients with four figures for $M=0$, $1^{\circ}, 2^{\circ}, \ldots, 25^{\circ}, 26^{\circ}$. The first table is devoted to the calculation of perturbations caused by Jupiter and Saturn, and the second to the calculation of perturbations caused by the four internal planets, which have an appreciable influence only when the planet passes near the perihelion.

In some cases, it is more useful to choose the excentricity, rather than the mean anomaly, as an independent variable. In these cases, equations (1) will be transformed by means of the relation

$$
\begin{array}{ccc}
\because & \because a \\
\vdots & d
\end{array}
$$

The points that correspond to equidistant values of $E$, are located on the orbit more uniformly than the poincs that correspond to equidistant values of M. This is particularly perceptible for large values of the eccentricity. Indeed, the series-expansion, considered in Ch. XII, indicates that
if we only keep the first power of the eccentricity. Hence, equidistant values of $E$ give "more equidistant" values of $v$ than do equidistant values of $M$. This can be shown in a more convincing way by the following table, which gives the values of the three anomalies for the case $e=0.85$, i.e. for a round value of the eccentricity of Enke's comet:


We obtain the most homzeneous distribution of the positions of the perturbed luminary in the orbit by choosing the true anomaly as the independent variabie by which we carry out the integration. This is easily done by using the relation

$$
\begin{array}{ccc}
d \\
d u & \cdots & \cdots=\left(\begin{array}{ll}
+ & d \\
d & 1 \\
d
\end{array}\right)
\end{array}
$$

We note that the choice of the eccentricity or the true anomaly as an independent variable considerably complicates the calculation of the coordinates of the perturbing planet.
70. Comparison between the formulae

As a rule, the perturbations of the elements are calculated to within $0^{\prime \prime} .0001$ In the average daily motion and to within $0^{\prime \prime} .001$ for all the other elements. In order to obtain such an accuracy, we take the case of a small planet with an interval $w=40^{\mathrm{d}}$ and perform the calculation to five decimal places. In the case of a comet, it is advisable to change the interval w , depending on whether the comet is near or far from the perturbing planet and also on its dis ance to the perihelion.

One should pay attention that the elenents of the perturbed planets as well as the coordinates evaluated by these elenents should be related to the same equator and equinox, as the coordinates of the perturbing planet.

In order to avoid unnecessary interpolations, we have to 'hoose the moments $t_{0}+k w$ such, that the coordinates of the perturbing planet are known for thess moments.

Let us consider the osculating system of elements a, e, i,... fur the epoch $\tau_{0}$. Choosing the interval $w$, we define the initial moment $t_{0}$ by the relation

$$
t_{0}-\frac{w}{u} \begin{aligned}
& 2 \\
& 2
\end{aligned}=-0
$$

We then calculate $r, v, u, p$ for the moments $t_{0}-2 w, t_{0}-w, t_{0}$ and $t_{0}+w$ using the given (i.e. unperturbed) elements by means of the formulae

$$
\begin{gather*}
E-e \sin t \cdot d t \\
r \sin b=a \cos \because \sin t  \tag{11}\\
r \cos u=a(\cos E-\sin \%) \\
u=v+\infty, \quad f=a \cos \varphi .
\end{gather*}
$$

We then take from the astronomical annual (or from Comrie's tables) the values of the coordinates $\left(r_{1}, \ell_{1}, b_{1}\right),\left(r_{2}, \ell_{2}, b_{2}\right), \ldots$ of the perturbing planets and calculate the corresponding orbital elements using the following relations:

$$
\begin{align*}
& \text { cu: } B_{1} \cos L_{1}=\left(\cos I_{1}-1\right) \cos t_{1} \\
& \cos B_{1} \sin L_{1} \cdots \sin i \sin b_{1}+\operatorname{cosicov} b_{1} \text {, in }\left(l_{1}-\underline{y}\right)  \tag{II}\\
& \sin B_{1}=\cos i \sin b_{1}-\sin i \cos b_{1} \sin \left(I_{1}-2\right) \\
& \begin{array}{l}
\therefore_{1} \quad r_{1} \cos B_{1} \cos \left(l_{1}-u\right) \\
r_{1}=-r_{1} \cos \beta_{1} \sin \left(l_{1}-u\right) \\
\therefore=r_{1} \sin \beta_{1} .
\end{array} \tag{!!}
\end{align*}
$$

We evaluate thedistance between the unpertu. bed body and the perturbing planet under consideration by

$$
\begin{align*}
\Delta_{i}^{n} & =\left(\xi_{1}-r\right)^{2}-1-i_{i} \because_{i}^{\prime} \\
& =r_{i}^{2}+r^{2}-2 r_{1}, \tag{IV}
\end{align*}
$$

We calculate the components of the perturbing accelfration by using the following relations:

$$
S_{1}=A_{1} \xi_{1}-\frac{u k^{\prime \prime} m_{1}}{k_{1}^{\prime}=\frac{u k^{\prime} \prime}{r} m_{1}\left(\begin{array}{c}
1  \tag{V}\\
j_{i}
\end{array}-\begin{array}{c}
1 \\
r_{1}^{\prime}
\end{array}\right)}
$$

where the values of $w k^{\prime \prime} m$, are given in table IV at the end of this book. We then sum the components of acceleration caused by the action of different planets, and obtain

$$
\begin{array}{ccc}
i & 1 & l \\
!i & \therefore i & \vdots!
\end{array}
$$

We calculate the functions which are subject to integration by means of the following formulae:

$$
\begin{aligned}
& \therefore \text { : © . . .! } \\
& \because r \text { r . } \therefore(1,1!
\end{aligned}
$$

$$
\begin{aligned}
& i_{i}=-(1, \operatorname{con}-r-r i=\text { contij) }
\end{aligned}
$$

We carry the integration of the first five elements using the following formulae:

We construct for w 8 n a column of second sums, the initial value of which is chosen as

We carry out a first integration, which yields wn as well as a secoad integration which yields $\Delta^{\prime} \lambda$ by using the following formula

$$
\begin{equation*}
\therefore, \rightarrow /_{n}^{2}: \frac{1}{12} f_{n} \cdots{ }_{211}^{1} f_{n} \cdot \frac{1}{140!} f_{n} \tag{1..}
\end{equation*}
$$

In all these 'ntegrations, the unavailable differences are obtained by extrapolation.

After integration, we obtain the perturbations of the elements $\Delta{ }_{i}, \Delta \Omega, \Delta \varphi, \Delta \pi, \Delta \epsilon, W \Delta n$ and $\Delta \lambda^{\prime}$, for the four moments mentioned above. Adding these perturbations to the initial osculating elements, we obtain the perturbed elements for these moments. The per= turbation of the semimajor axis can be found by the following differential relations


In order to obtain more accurate values for the quantities $\mathcal{S}_{i}, \delta \Omega, \ldots$ we repeat all the abovc mentioned alculations starting with the perturbed elements and re-integrate For this purpose we use the same forralae, with only one small modification. We calculate the average anomaly using equations

$$
; \quad ; \quad \because \quad ; \quad 1 ; 1
$$

in which $\epsilon_{0}$ and $n_{0}$ are the initial values of the elements $E$ and a which correspond to the epoch $\tau_{0}$. This calculation repetion is continued chosen, the second approximation may be considered as final. One then continues calculating the perturbations for the next moments $t_{0}+2 w$, $t_{0}+3 w, \ldots$ (or $t_{0}-3 w, t_{0}-4 w, \ldots$ depending on the direction that should be fol:owed starting from the initial epoch). If the perturbations are large, and the final values of tle differences of the functions $\delta i, \ldots$ are significantly different from the primary values, we then should once more repeat the perturbation calcula ions for 'he moments $t_{0}-2 w$, $t_{0}-w, t_{0}$, and $t_{0}+w$.

The continuation of the calculation is not difficult and since r.he osculating elements vary slowly, extrapolation gives such accurate values for the perturbations, that it is never necessary to repeat the calculation of $\delta i, \hat{\delta} \Omega, \ldots$, provided that the interval $w$ is appropriately chosen.

Calculations have to be done on several separate sheets. The values of $r, v, \ldots, \delta i, \ldots, w \delta n$ are obtained on the first sheet. The second sheet is subsediary to the first one. There, we calculate the coordinates of the perturbing planets, $\xi_{1}, \bar{\eta}_{1}, \bar{\gamma}_{1}, \ldots ., \ldots$, the conresponding components of acreleration $\mathrm{S}_{1}, \mathrm{~T}_{1}, \ldots$ and the quantities $\mathrm{S}, \mathrm{T}$ and W . The computations that correspond to a given moment should be written in the same vercical column in both sheet.s. The integration of each of the functions $\delta i, \delta \Omega, \ldots$ has to be done in a separate sheet according to the scheme indica ed in chapter VIII.
71. Particular cases for calculation of perturbations of the elements of small plariets

Two kinds of difficulties are encountered in the application of the methods considered in the frevious sections to small planets.

1- If the slope of the orbit is very smal1, then the computation of the quantity

$$
\therefore \quad \begin{array}{llll}
\therefore \because & . & 16 \\
\vdots & 1 & 1
\end{array}
$$

wil1 be accompanied by a large decrease in accuracy, and leads to the inaccurate definition of the longitude $f$ the node.
2. If the eccentricity of the orbit is small, then a similar difficulty occurs on calculating the longitude of the perihelion, which requires the computation of

In the first case, the easiest way is to change the basic plane so that the elements $i, \Omega, T$ and $\in$, related to the plane of the aciiptic, are transformed into elements $i^{\prime}, \Omega^{\prime}, \pi^{\prime}$ and $E^{\prime}$ (instead of the last two, one may take $w^{\prime}$ and Mo), related to the plane of the equator.

In the cases when only ar. approximate calculations is required with an genuraidy not exceeding the first powers of the mass, it is better to introduce the arxiliary variables $p$ and $q$ instead of $i$ and $\Omega$ as defined by

$$
\begin{aligned}
& \text { ap } 1: 1 . \cdots i=i
\end{aligned}
$$

The perturbations of these elements are calculated by
which can easily be deduced from equation (2). The perturbaticns of $i$ and $\Omega$ can be found up to within quantities of the first order by means $f$ the relations $\quad \therefore \cdot(0, \ldots \cdot$ ip $1: 1 \% 1 \%$ $\because 1: \because \quad H \cdot I_{1} \cdot\left(1, \ldots \cdot S_{4}\right.$.

We sow consider the second case, in which the eccentricity of the orbit is small. instead of $e$ and $\pi$, we introduce the new variables

$$
\text { h ri!!: } \quad \text { rine. }
$$

Since,
we then easily see according to equations (2) that
where

The integration of these equations gives the perturbed values of $h$ and Equation (21) can then be used to calculate the corresponding va-ues of e and .
72. Some aspects on thecalculatinn of perturbations of the orbital elements
of comets
The calculation of the pert bations of short-periodic elements, for which the eccentricity is not so large that the use of the eccentric anomaly is impossible, is performed by means of the formulae given in Sec. 70. However, for a comet whose eccentricity is near unity, these formulae should be partiaily transformed in such a way, that instead of elements $\varphi, n$ and $\mathcal{E}$ the perturbation of the elements $e, q$ and $\mathcal{Z}$ are obtained (the time of passing by the perihelion). We shall not consider here these transfolmations. After cowell's and Cromnelin's work on

Halley's comet, the perturbations of comets of this type are nct calculated for the $e^{1}$ ements but for the rectangular coordinates by the methods which will be considered in the following chapter.

The comet may pass su close to one of the large planets, that the gravitation of this planet becomes stronger than the gravization of the sun. In tihis case, it is advisable, as Laplace (Mochanique Celeste, t.4, Livre IX, Chap. II) pointed out, to consider the planet as the central body and the sun as the perturbing one. Let us denote by $x, y$ and 2 the beliocentric coordinates of the planet, and by $m_{1}$ its mass. We write the differential equations for the comet's motion under the influence of the sum's gravitation and planet:
and the equations of the unperturbed motion of the planet:

If we take the centre of the planet as the origin of a cuordinate system, the axes of which are parallel $s o$ the axes of the heliocentric system, and denote the corresponding coorcinutes of the cluet by $\bar{\zeta}, \eta$ and $\tau$, we obtain

Using equations (22) and (23), we obtain

$$
(: 11
$$

where

$$
1 \div, \quad \vdots \because \because
$$

When the sun is $t$ ken as the central body, the quantity $R$ is the acceleration it imports to the comet, and $F$ is the perturbing acceleration caused by the attraction of the planet. Equation (22) shows that

On the other hand, when the planet is taken as the central body, we dencte the acceleration it imparts to rhe comet by $R_{:}$, and the perturbing acceleration caused by the sum by $F_{1}$. Then, it follows from equations (24) that

The region in epace, in which it is more useful to take the sun as the central body, is separatel from that, in which it is more useful to consider the planet as the central body, by points satislying the following equation

$$
\begin{array}{ll}
r & \because \\
i & \therefore
\end{array}
$$

## Putting

$$
x_{1} 亏: \begin{aligned}
& y_{1} y_{1} \cdot 1 z_{1}:=\cos \sigma_{1} \cdot \begin{array}{l}
\Delta \\
r_{1}
\end{array}-u .
\end{aligned}
$$

we obtain

$$
\begin{aligned}
& x x_{1}-\dot{y} y_{1}+z x_{1}-x_{1}\left(x_{1}+j\right) \mid y_{1}\left(y_{1}+x_{1}\right): u_{1}\left(x_{1}-i\right)= \\
& =r_{1}^{\prime}(i \mid u \cos 1)
\end{aligned}
$$

Substituting these equations into the previors ones, we obtain

$$
\begin{aligned}
& \left.-2(1+2 \cos 4)(1+2 u \cos 9) \cdot u^{\prime}\right)\left.\right|^{\prime} \mid \vdots .
\end{aligned}
$$

Expanding the right-hand side of this equation in powers of the small quantity $u$, we obtain

Keeping only the first term of this expansion, we obtain

$$
\perp=r_{1}\binom{m_{i}^{i}}{1 / 1-j-3 \cos -11}^{\prime}
$$

This is an approximate equation for the plane under consideration in the polar coordinate system. The surface (25) is evidently a surface of rotation around the polar axis i.e. around the radius vector of the planet. This surface slightly differ from a sphere; the ratio of the maximal to minimal value of $\Delta$ is equal to

$$
\because \square^{\prime}=1.15
$$

The sphere, drawn from the centre of the planet with a radius equal to

$$
s_{0}=r_{1} r_{i_{1}}{ }^{\prime}
$$

is cailed the sphere of activity of the planet. The radi of the spheres of activity of some planets aregiven in the following table

| Mercury . . . . . . . . . . . . . . . | 0.001 | Jupiter . . . . . . . . . . . . | 0.322 |
| :--- | :--- | :--- | :--- | :--- |
| Venus . . . . . . . . . . . . . . . . . | 0.004 | Saturn . . . . . . . . . . . . . | 0.362 |
| Earth . . . . . . . . . . . . . . . . . . | 0.006 | Uranus . . . . . . . . . . . . . . | 0.339 |
| Mars . . . . . . . . . . . . . . . . . . . | 0.004 | Neptune . . . . . . . . . . . . | 0.576 |

Inside the sphere of activity $F: R \geqslant F_{1}: R_{1}$. Outside this sphere $F: R<F_{1}: R_{1}$ so that it is more useful to consider the planet as the central body.

It is interesting to evaluate the ratio $F: R$ of the perturbing force to the gravitational force of the sun for points located on the sphere of activity. It is easy to sec that for such points

For Jupiter, the limits within which this ratio varies are from
$m_{1}=\sqrt[5]{0.25}$ to $\sqrt[5]{16 m_{1}}=0.43$.
We have to deal quite of ten with comets passing through the sphere of Jupiter's activity. This is explained by the fact that the aphelions of most of the short-periodic comets are grouped near Jupiter's orbit. In order to make a transformation from the heliocentric coordinate system to the Jove-centric coordinate system, in which Jupiter is taken as a central body, it is necessary to find the Jove-centric coordinates at the time $t_{0}$ for which the distance $\Delta$ is reduced to 0.3 . These will be denoted by $\quad:, x=x-x_{1} . \quad r_{11} \quad v-l_{1,} \quad i_{0} \quad \&-z_{1}$
while their derivatives will be given by

$$
\therefore \quad 1 \quad 1.1 \quad r_{1},-y \quad y_{1}, \quad \therefore \quad 4-x_{1} .
$$

Using the formulae obtained in Chapter IV of Volume $I$, we can obtain the Jove-centric elements of the comet. The perturbations of these elements are computed by means of conventional formulae. The sun in almost all cases can be considered as the only perturbing body. It is interesting to note that the Jove-centric orbit of the comet is usually a hyperbola of large eccentricity. The eccentric anomaly appearing in the formulae of Sec. 70 will then be imaginary (vol. I, Sec. 15). Consequently, it is worthwile transforming these formulae, adopting that

$$
E-i l!\quad 1
$$

We shall not consider here these complicated transformations. We find that, in the case under consideration when the comet is inside or even near the planet's sphere of activity, it will be more convenient to calculate the perturbations in the coordinates using Cowell's method. This method will be considered in the next chapter. In order to use this method, it is sufficient to find the coordinates of the comet $x_{0}, y_{0}$ and $z_{0}$ and their derivatives $x_{0}, y_{0}$ and $z_{0}$ at the time $t_{0}$ by means of the osculating elements (Ser. 74),

The computation of the perturbatjons of the Jove-centric coordinates has been carried out by Kamienski ${ }^{(1)}$ in his study on the motion of Wolf's comet during its approach to Jupiter in 1922. In rhis case, the eccentricity of the Jove-centric orbit varied from $e=6.457$ to $e=6.480$, and the
(1) M. Kamienski, Recherches sur Ie mouvement de la Comite periodique de Wolf (IX Partie), Publications of the Astr. Obs. of Warsaw University, 2, 1926.
semimajor axis from $a=-0.022800$ to $a=-0.022912$.
When the cowet under consideration closely approaches Jupiter, one has to take into account the perturbation caused by the compression of Jupiter. Such an approach took place,for example with Brook's comet (1889V), when the least distance between Jupiter's surface and the comet became apprcximately 1.14 times Jupiter's radius (approximately 80.000 km ). In considering thiscase, the interval w used in integration has to be reduced to 0.25 hours in the vicinity of the time of approach. For details on the effects of the compression of Jupiter on a comet's motion, we refer the reader to the literature quoted herein ${ }^{(1)}$.

## 73. Approximate Calculation of Perturbations of Small Planets

The exact calculation of the perturbed coordinates of a small planet by a numerical integration requires a large amount of work. This does not depend on whether we are calculating the perturbations in the elements or immediately calculating the purturbed coordinates (Secs. 74, 75). This type of calculation is only carried out for planets, which are interestIng in some aspects. The analytical theory of motion of the type developed by Leverrie and Newcome for large planets has been applied to an even smaller number of small planets. It is not, however possible, to calculate the perturbations of small planets for, after a few years their actual motions will differ so much from their unperturbed motions, that the planets will hardly be distinguished and would thus be
(1) G. Deutschland, Der Eintluss der Abplattung auf die Attraction der Himmelskorper nach der Theorie der speziellen Storungen, mit Anwendung auf den Kometen 1889 V (Brooks) bet seiner Kupiternahe un Jahare 1886 (Diss) Berlin 1909.

The results are partially given in. Astr. Nachr., 181, 1909, 1-8.
$-284-\quad$ ORIGNAL PAGE in
OF POOR QUALSTY a simple method which would allow us to calculate the perturbations of small planets with an accuracy that would identify these planets. Several methods have been suggested for this purpose and amongst them, the following two approximate forms of an accurate method suggested by Starke ${ }^{(1)}$, have been widely practised.

We first of all point out that it is suffieient to only take into account the perturbations caused by Jupiter for the approximate calculations of the perturbations necessary to identify these planet.

We shall limit ourselves to an accuracy within 0.0001 for the perturbations of $n$ and up to within 0.0001 in the perturbations of all the other elements. We shall then be able perform all computations to three decimal places. The interval w can always be taken equal to $80^{\mathrm{d}}$, but when the planet is far from Jupiter (e.g., when the heliocentric angular distance between them is greater than $60^{\circ}$ ) the interval w may be set equal to $160^{\mathrm{d}}$. It is useful to take all the constant factors for the interval $80^{d}$ and complete the missing ones by interpolation at $w=160^{\mathrm{d}}$. Under these conditions, we shall be able to make considerable simplifications in the calculations of the formulae obtained in Sec. 70. Instead of computing $v$ and $r$ using equations ( $I$ ), we can find their values by consulting the special tables given in Volume 1 , Chapter III. We can always substitute in equations (II) $\cos \mathrm{b}_{1}=1$, and, If the slope $11 s \operatorname{small}\left(i<8^{\circ}\right)$, put
(1) G. Starke, Genaherte Storungsrechnung und Behverbesserung, Veroff des Astr. Recherinstituts, Nr. 44, 1921. Berlin; Tafein zur genaherten Speziellen Storungsrechnung, Veroff. des Astr. Recheninstituts, Nr. 48, 1930, Berlin.

Furthermore, we have to replace the coefficient wk ${ }^{n} m_{1}$ in formulae (V) by $80 \mathrm{k}^{\circ} \mathrm{m}$, where $\mathrm{k}^{0}$ is the Gaussian constant expressed in degrees. Fo: Jupiter,

$$
\lg \left(50 \Leftrightarrow m_{1}\right) \cdot 1.577 .
$$

In calculating using formulae VII, we use

$$
\lg \left(3 k^{k}\left(w^{\prime}\right)=1.616\right.
$$

This yields $\mathcal{S}_{\mathrm{n}} \mathrm{n}$ in units of $0^{\circ} .0001$. The increments of the other elements will be expressed in units of one-thousandth of a degree. We can generally neglect the differences in the calculations of formulae (VIII) and (IV). The integration will then be reduced to a sinple summation. Finally, we note that the perturbed value of $n$ is not required, since we calculate the average longitude using formula

$$
\left.\lambda . t+n,\left(t-t_{1}\right): a^{\prime}\right\rangle,
$$

where $n_{0}$ denotes the unperturbed value of $n$. However, the quantity $\omega \Delta n=80 \Delta \mathrm{n}$ is necessary for the computation of the perturbation of the semimajor axis. This latter quantity may be calculated by formula
where $n_{0}$ is expressed in seconds of the arc.
In 1930, Starke developed another version of this method. It requires the uss of subsidiary tables, but once these tables are available ${ }^{(1)}$ the amount of work necessary will be considerably reduced.
(1) These tables are quoted in the previous foot note.

Here, the average anomaly $M$ is taken as the incependent variable. The coefficients in formulae (VII), Sec. 70, are functions of . They are tabulated by the argument $\varphi$ for the values $m=0^{\circ}, 12^{\circ}, 24^{\circ}, \ldots 348^{\circ}$. The computation of the components of acceleration is also simplified by special tables. We shall not consider here the construction of these tables.

Finally, we remind the reader of the method of approximate computation of perturbations in the elements, suggested by Stromgren ${ }^{(1)}$. This method is similar to Storke's method mentioned above in that both methods are designed for their use with calculating machines. Strömgren transformed the formulae in such a way, that they can be quickly and easily used to find the perturbed values of the directing cosines $P_{x}, P_{y}, P_{z}, Q_{x}, Q_{y}$ and $Q_{z}$.
(1) 8. Stromgren, Formelu sur genaherten Srorungsrechnung in Bahnelementen, Publikatione og mindre Meddelelser fra Kobenhavens Observatorium, Nr. 65, 1929.

## CHAPTER XI

## GALCULATION OT THE PERTURBATIONS IN THE COORDINATES

74. Direct Calculation of the Perturbations in the Coordinates (Cowell's

Method).
We consider the motion of a luminary, the mass of which is denoted by $m$ and the heliocentric coordinates by $x, y$ and $z$. We denote by $m_{i}$ and $x_{i}, y_{i}, z_{i}$ the masses and coordinates of the pertarbing planets. The euqations of relative motion, derived in Sec. 3, yield

$$
\left.\begin{align*}
& d-x  \tag{1}\\
& d t^{2}=-b^{2}(1+m) \frac{x}{r^{3}}: f_{x} \\
& d^{3} y=-k^{2}(1+m) \frac{y}{r^{3}}: f_{y} \\
& d d^{2} \\
& d^{2} z=-k^{2}(1+m) \frac{z}{r^{s}}+f_{2} \\
& d t^{2}=-1
\end{align*} \right\rvert\,
$$

where

$$
r_{1} \sum_{1} m_{1}\left(\begin{array}{c}
x_{1}-x \\
\Delta_{1}^{3}
\end{array}-\begin{array}{l}
x_{1} \\
r_{1}
\end{array}\right)
$$

Here, the summation is over all the perturbing bodies and the quantities $r^{2}, r_{i}$ and $\Delta_{i}$ are defined by

$$
\begin{gathered}
r^{2} \quad x^{2}-y^{2}+z^{\prime \prime}, \quad r_{1}^{2} \quad x_{1}^{2}+y_{i}^{:} \quad z_{1}^{2}, \\
\lambda_{i}^{*} \cdot\left(x_{i}-x\right)^{2}+(y-y)^{2}
\end{gathered}
$$

If it is required to compute the values of the coordinates $x, y$ and $z$ for a relatively short interval of time for e.g. some decsjes, the easiest way then is to integrate equations (1) numerically. Any of the numerical integration of differential equations methods that have been considered in chapter VIII, enables us to calculate the values of $x, y$ and $z$ with an arbitraril.y high accuracy. The calculation by means of any of oŕ these methods is straightforward and elementary. This is extremely
important, especially when is doing substantial work, since it enables the use of calculating machines. Another advantage of this particular method of computation of the perturbed coordinates, is its universal character. The analytical methogs of finding the perturbations are only valid if perturbations are small. The methods of numerical calculation of the perturbations in the elements, given in the previous chapter, are oaly preferentially applied if the perturbations are not very large, although they are generally valid for arbitrary perturbations. On the other hand, the question of the magnitude of perturbations is not raised in the numerical integration of equations (1). Consequently, this method is conveniently used for small planets, subject to small perturbations, and for planets which approach Jupiter so closely that their perturbations become particularly large. Similarly in this method, there is not difference between comets which are far removed from planets and comets entering the sphere of activity of a planet. Finally, the numerical intrgration method enables us to compute in a very simple way the positions of such bodies like Jupiter's eighth satelite, whose magnitude of perturbation is comparable to the value of the central force. We have seen in Sec. 41 that the most complicated cases of motion can be studied by means of mettods of the numerical integration of equations. These methods had been used by Darwin, Tile, Burrau and Stormgren long before Cowell applied numerical integrations to solve astronomical problems. Indeed the direct numerical integration of the equations of motion had been applied before Cowell. Moreover, the method which be first developed was inferior to the method of quadratures suggested by Gauss and applied by Tile and the other previously mentioned authori. This is the reason for our raturning later on to this method.

We however, owe Cowell the introduction of the method of direct numerical iniegration of equations (1) into the field astronomy as a relatively rapid and practical method of obtaining the perturbed ephemeride. For these reasons and for convenience sake, we call the method of calculating the perturbations of the coordinates of a luminary by means of the numerical Integration of equations (1) as Cowell's method.

We have already mentioned the advantages of Cowe?l's method, which in many cases makes this method the most practical way of obtaining the ephemeride of a luminary, taking into account its perturbations. We can also point out, that in this method the trigonometric calculations are completely singled out. The only auxiliary table necessary for these computations is the table that gives the values $r^{-3}$ by values of the argument $r^{2}$,

One of the most serious difficulties of Cowell's method consists in that all the intermediate calculations must be carried out to at least the same number of significant figures; the final result is required to have (In practice, one should carry out the intermediate calculations with a large number of significant figures to guarantee against the accumulation of errers). For example, if we have to connect two appearance times of a comet separaced by a large interval of time, if we use Cowell's method we then must calculate the perturbed coordinates to seven sfgnificant figures for this interval. If we apply the methods developed in: the previous chapter, it will then be sufficient to calculate the perturbations of the elements during this interval of time., rom three to five significant figures.

If the 1 uminary closely approaches the sun, which is the case for most of the comets near the perinelion, the value of the interval of integration must then be significantly deß'eased. The amount of computation
required by Cowell's method will then be more substantial than that required, for example, for Euke's method (Sec. 76).

Ke finally note that Cowell's method is at least convenient for accurate calculation only in the case when the computations will be carried out by means of calculating machines. ${ }^{(1)}$

## 75. A Compilation of theformulae used in Cowell'. method

We denote by w the interval which will be chosen for the integrations of equ.
(1). We have already pointed ot in Sec. 58 that this interval should never be chosen very large, otherwise there will be less chance to control the computation accuracy by means of the differences. In addition the calculations would become less practical.

It is advisable to start the computation by using a small value fow w. If the fourth differences are then found not to affect the calculated values of the coordinates, the interval magnitude may then be doubled. For small planets, the interval may be set io equal 20-40-80 days, depending on the required accuracy. We have to take smaller interval values, say $5-10$ days, for comets close to the sun. If it
(1) The integrations of equations (1) can, of course, be carried out using any method for the numerical integration of differential equations. In particular, instead of applying the method of quadratures, we can use Cowell's method (Sec. 52), supplemented by Numerov's method for the reduction of successive approximations (Sec. 58). We then obtain Numerov's method or the method of extrapolation". A detailed account of this method, as well as an example of its application are given in: Belletin of the Institute of Astronomy (Bjulleten Astronomiceskogo Instituta) No. 12, 1926.

Probably, the method of quadratures (Secs. 52-54, 58) is the best method for integrating equations (1) from the point of view of accuracy of the results and the simplicity of computation.
is required to furtiner reduce the interval, it is then advisable to replace Cowell's method by Enke's method (Sec. 76).

We shall consider that the values of the osculating elements $a$, $e, i, \Omega$, $w$ and $M_{o}$ of a luminary at moment $t_{0}$ are known. In order to find the solution of equations (1) which describes the metion of this luminary, we have to calculate the values of coordinates $x_{o}, y_{o}$ and $z_{0}$ and their derivatives $x_{0}^{\prime}, y_{0}^{\prime}$ and $z_{0}^{\prime}$ at moment $t_{0}$. For this purpose we use th? following formulae (Vol. I).

$$
\left.\begin{array}{cc}
x & a P_{1}(\cos E \\
y & a J_{v}(\cos E-e)+b Q_{x} \sin E \\
z & a P_{y}(\cos E-c) \cdots b Q_{i} \sin E
\end{array} \right\rvert\,
$$

in which

$$
\left.b=a \cos p, \quad r \cdots x^{2}+\left|\cdot y^{2}+\because^{2} \quad a\right|-e^{\prime} \cos t\right),
$$

where $E$ is defined by

$$
\text { E--esint } I t
$$

For checking the calculations, we use the following relation

$$
k e y \text { asin } t \cdot x x^{\prime}+y y^{\prime}+2 x^{\prime}
$$

The directing cosines of the orbital axes, $P_{x}, F_{y}, \ldots, ?_{z}$, may be calculated by using the following formulae (Vol. I, Sec. 25):

$$
\begin{aligned}
& A_{1} \cdot \cos \operatorname{Li} ; \quad d \cdot \cdots \cos i \operatorname{lin}!
\end{aligned}
$$

$$
\begin{aligned}
& F_{x}-A_{1} \cos \omega-\frac{1}{-} A_{4} \sin \omega ; \quad O_{8}-A_{1} \cos =-\lambda_{1}+\ln \omega \\
& P_{y}=B_{1} c(1) \omega 1-i_{2} \operatorname{cill} \omega ; \quad Q_{y}=B_{2} \cos \omega-i_{1} \sin \alpha \\
& f_{z}-C_{1} \cos \omega+C_{2} \sin \alpha ; \quad O_{2} \quad C_{2} \cos \omega C_{1} \text {, } 11 \cdots,
\end{aligned}
$$

where $\epsilon$ denot 3 s the slcpe of the ecliptic with respect to the equator. For checking, we app ${ }^{1 / y}$ the following relations

When the initial values $x_{0}, y_{0}, \ldots, z_{0}^{\prime}$ are slreav; calculated, we start integrating equations (1). For this purpose, we first of all calculate for each considered moment $t_{n}=t_{o}+n w$ the values of the functions

$$
\begin{aligned}
& x=\sum_{1}^{1}\left(w=m_{1} x_{1}-x-x_{1}^{\prime}\right) \\
& \because \quad \sum_{1}^{1}\left(u_{1}^{2} \alpha_{1}-m_{1}^{x} y^{x}-r^{4}\right) \\
& Z \quad \sum_{1}^{1}\left(w^{2} k^{2} m, y_{1}^{y} \quad-z\right) \text {. }
\end{aligned}
$$

where the particularly singled-out quantities

$$
x^{2}=-w^{2} k^{2} m_{1}{ }_{r_{i}^{3}}^{r_{1}}, \quad Y^{i}=-w^{2} k^{2} m_{i} \dot{r}_{i}^{3}, \quad Z=\omega^{2} k^{2} m_{1} r_{1}^{r_{i}^{3}}
$$

depend entirely on the coordinates and masses of the perturbing planets.
The irreplaceable hand-book used for these calculations; are Comrie's tables: "Planetary co-ordinates for the years 1800-1940" which we have mentioned several times before. The extension of these tables to the years $1940-1060$ is expected to appear in the near future. Besides the coordinates ofall the large planets (except Mercury and Plutonus), these tables also give the corresponding values of $X^{i}, Y^{i}$ and $z^{i}$. Moreover, several other auxiliary tables are given there, and in particular the table by which the quantity $\mathrm{r}^{-3}$ can be found from the value of the argument $r^{2}$.

If Comrie's tables are not available, we can compute the rectangular coordiates of the perturbing planets by the values of the corresponding ecliptic coordinates obtained from the year-book. For this purpose, we use the following relations
where $r_{i}, \ell_{i}$ and $b_{i}$ are the radius vector, the longitude and latitude of the planet respectively.

The constants involved in the expressions $f, g$ and $h$ are given in table IV at the end of this volume.

We apply the following formula during integration:

as well as two similar formulae defining the values of $y$ and $z$ which correspond to the moment $t_{c}+n w$. The initiai terms of che columns of sums are defined by the following formulae (Sec. 54):

In order to avoid some of the successive approximations at the very beginning of the computation (cf. Sec. 55), it is easy to determine, using formulae (I), the values of the unperturbed coordinates for a few moments $t_{-2}, t_{-1}, t_{1}, t_{2}, t_{3}, \ldots$ near the initial moment $\varepsilon_{0}$. One approximation wili then be sufficient to obtain the final (perturbed) values of the coordinates and the moment. For this purpose it is also possible to use the following Taylot expansion.

$$
x\left(t_{0}-\vdots n u\right)=x_{n} ;-x_{0}^{\prime}\left(t_{n}-t_{0}\right) \cdots \frac{1}{j} f_{0} u^{\prime \prime}\left(t_{n}--t_{0}\right): \ldots
$$

When some of the initial values for the coordinates are corrected and the final values of the quantities (VI) are obtained, we further integrate quite easily by means of formulae (V)

## Annotation I

When the Gaussian constants $a, A, b, B, \ldots$ corresponding to the initiai osculating elements are known, then formulae (I) and (II) are recomended to be replaced by

$$
\begin{align*}
& x \quad r \sin a \sin (: 1 \text {; } u) \\
& y=r \sin b \sin (i) ; a)  \tag{111}\\
& =-r \sin c \sin (c \cdot!-m)
\end{align*}
$$

$$
\begin{align*}
& \left.\left.y^{\prime}-r^{\prime}\left|j^{\prime} r^{\prime} k^{\prime} l^{\prime} p \operatorname{sinbcw(l|}\right| u\right)\right]  \tag{II'}\\
& \left.z^{\prime}=-r^{\prime} \mid \because r^{\prime} \text { 仅 } r \text { sinc cosic: } u\right) \mid \text {. }
\end{align*}
$$

where $p=a \cos ^{2} \varphi$, quantities $r, r^{\prime}$ and $u$ are calculated by means of the following formulae

$$
\begin{gathered}
r-a(1-c \cos t), \quad r=\operatorname{tec} a \sin t, \\
\lg \frac{t}{2}=\sqrt{\frac{1-1}{1-c}} \operatorname{tg} \frac{t}{2}, \quad u \cdot v-w,
\end{gathered}
$$

and for checking, we use the following auxiliary equation

$$
r r^{\prime}=x x+y y^{\prime}:=z^{\prime} .
$$

## Annotation II

If the luminary, whose motion is under investigation, exists at a sufficiently largedietance from the sun, then the perturbations that Mercury, Venus, ... produce in its motion, are either insensible or quite small. In such cases, it is sufficient to only consider the secular parts of these perturbations. This can simply be done by correspondingly increasing the mass of the sun.

The factor $1+m$, involved in equations (1), will be replaced by unity, since the mass m of a small planet or comet may always be set equal to zero. If we consider that the mass of the sun is unity, and take into consideration the masses of the perturbing planets, ther formulae (III) will be replaced by
where factor $M$ has one of the following valces

$$
\begin{aligned}
M & =1.0000014 \quad \text { (sum of masses of sun and Mercuty) } \\
& =1.00000260 \quad \text { (sum of masses of sun, Mercury and Venus) } \\
& =1.00000564 \quad \text { (sum of masses of sun, Mercury, Venus and Earth). } \\
& =1.00000596 \quad \text { (sum of masses of sun, Mercury, Venus, Earth and Mars). }
\end{aligned}
$$

76. Emke's Method

Instead of finding the perturbed coordinates of a luminary by means of the numerical integration of equation (1), it is possible to calculate the differences

$$
\xi=x-x, \quad \eta=y-\ddot{y}, \quad:=z-z
$$

between the perturbed coordinates $(x, y, z)$ and the unperturbed ( $\bar{x}, \bar{y}, \bar{z}$ ). Since the unperturbed coordinates satisfy the following equations:

$$
\begin{aligned}
& \frac{d^{2} x}{d f^{2}}=-k^{2}(1-m) \dot{x} r^{2} \\
& d t^{2} y=-k^{2}(1 \mid-m) y r^{-}= \\
& d z=-k \cdot\left(1 \cdot(\cdot m): r^{-3} .\right.
\end{aligned}
$$

then by subtracting these equations term-wise from equations (1), we obtain'

$$
\begin{align*}
& { }_{d!}^{\prime \prime \prime}=r_{x}+R^{2}(1-1 m)\left(\begin{array}{l}
x \\
r
\end{array}-\frac{x}{r}\right)
\end{align*}
$$

Thus, the calculation of the differences $\mathcal{\xi}, \eta$ and $\mathcal{\tau}$, which are nothing else but the perturbations of the rectangular coordinates, is reduced to the integration of equations (1).

Particular attention should be devoted to the evaluation of the second terms in the right-hand ride of equ. (4), The direct computation of the differences in the brackets is accompanied by a large accuracy 10ss. It is more corvenient from the computational technique point of view, to transfoim these equations into the following form

$$
\left.\begin{array}{lllll|lll}
\vdots & \vdots & r_{r}^{\prime} & \therefore & \frac{!}{r} & (1 & r & 1
\end{array} \right\rvert\,
$$

It is thus clear that all three coordinates could be found if the difference $1-\frac{\mathbf{r}^{3}}{r^{3}}$ is obtained. Since

$$
\begin{aligned}
& r^{2}=(x+3)^{2}+(y+1)^{2}+(z+:)^{2}=
\end{aligned}
$$

we can then write that

$$
\frac{r}{r}=1+20 .
$$

where

$$
\begin{equation*}
u=\frac{1}{r^{2}}\left|\left(x+\frac{1}{2} \dot{\xi}\right) \vdots+\left(y-\frac{1}{2} r\right) r_{1}+\left(z+\frac{1}{2} y^{\prime}\right) \dot{\mid}\right| . \tag{5}
\end{equation*}
$$

Therefore,

Following Enke, we adopt that

$$
\begin{equation*}
f=3\left(1-\frac{5}{2} a: \frac{35}{6} q^{2}-\frac{315}{24} q^{3} ; \ldots\right) \tag{b}
\end{equation*}
$$

This yfelds

$$
1-r_{r^{3}}^{r^{3}} \quad q \rho
$$

Hence,

$$
\underset{r}{x}-r_{r}^{x}=r_{r}(4 f x-5)
$$

When the differences are presented in this form, they can be computed without any accuracy loss. The reevaluation of the following functions
required for the numerical integration of equations (4), will be carried out by means of the following equations

$$
\begin{align*}
& F=x+\frac{w^{2} k^{2}}{r^{\prime}} \cdot(q / x-i) \tag{VII}
\end{align*}
$$

where quantities $X, Y$ and $Z$ are those defined by formulae (IV) of the previous secticn.

Quantity $f$, involved here and defined by equation (6), may be determined by using table III at the end of this volume. The value of $f$ is given in sixteen decimal places in Comrie's tables previously mentioned.

In the following, we enumerate the operations that have to be carried out for the application of Comrie's method.

1- Starting with given values for the osculating elements, we calculate the unperturbed equatorial coordinates using equations (I) and (I') for a series of moments $t_{h}=t_{0}+h w$, where $h=-2,-1,0$, $1,2, \ldots$. It is useful to choese the initial moment so that the epoch of osculation takes place in the moment $t=t_{o}-\frac{w}{2}$. In the following, this will be assumed to be the case.

2- We use Comrie's tables to find the values of the rectangular coordinates $x_{i}, y_{i}$ and $z_{i}$ of the perturbing planets and the corresponding quantities $X^{i}, Y^{i}$ and $Z^{i}$ for all the mements whose perturbations intend to calculate.

3- In order to start integrating we compute the values of quantities $F_{-2}, F_{-1}, F_{0}, F_{1}$ and $F_{2}$ using formulae (IV) and (VII). In the first approximation, we assume that

$$
\vdots=r_{1}=:=-0, x x, y=y_{1} z \therefore q-0 .
$$

We determine the initial terms of the column of sums by the following formulae (cf. Sec. 54 ; in the present case we have $\xi_{\overline{7}}=\boldsymbol{\eta}=\tau=0$ and $\xi^{\prime}=\eta^{\prime}=\mathcal{\varepsilon}^{\prime}=0$ for the epoch of osculation $t_{0}-\frac{w}{2}$ ):

$$
\begin{align*}
& F^{-1}=-\frac{1}{2} F_{-\frac{1}{2}}^{1}+\frac{1}{339} r^{3} \frac{1}{2}-. .  \tag{VIII}\\
& \left.F_{0}^{2}=1 \cdot \frac{1}{24} F_{-1}-\frac{1}{333}\left(2 F_{-1}^{:}+F_{4}^{2}\right)+\cdots\right\}
\end{align*}
$$

we calculate the perturbations using the following formulae

$$
\begin{equation*}
\bar{B}_{n}=F_{n}^{-1}+\text { Red, } \quad \text { Red }=\frac{1}{!2} F_{n}-\frac{1}{\cdots 40} \Gamma_{n}+\frac{1}{19 j \mid} F_{n}^{4}-\ldots . \tag{IX}
\end{equation*}
$$

We then repeat the calculation of quantities $F_{-2}, F_{-1}, F_{0}, F_{1}, F_{2}, \ldots$ using the values obtained for $\xi, \eta$ and $\mathcal{C}$. We repeat this prosedure until we find that these quantities do not improve further.

4- When the above-mentioned calculations give the final values' for the quantities (VIII), we start integrating conventionally using formulae (IX) and substituting therein, the extrapolated values of the differences or the Red correction (Sec. 55). We add the perturbations of the coordinates obtained thisway to the unperturbed coordinates previously obtained (item 1). This leads us to the required values of the perturbed coordinates:

$$
x=-x-\vdots y=y: r_{1}=\sim z:=
$$

In conclusion, we note that the calculations made by Enke's method are more useful than the calculations made by Cowell's method only in the case when perturbations $\xi, \eta$ and $\mathcal{T}$ are small. The values of these perturbations can always be reduced by changing the epoch of osculation of the elements for which we are calculating the unperturbed coordinates. However, the calculation $n f$ the new osculating elements by means of the obtained perturbed coordinates (Vo1. I, Chap. IV) constitutes method. Thereture, this method is only applied if the perturbations are small and when it is necessary to evaluate them for small intervals of time. Under these conditions (e.g., for comets only observed unce) Enke's method is the best.

Sometimes, in the study of the motion of comets, the two methods, Enke's and Cowell's, can be combined. When a comet is far from the sun and is subject to considerable perturbations from the planets, it is better to apply Cowell's method since this method is the most general and the most independent from the magnitude of perturbations. On the other hand, when the comet is near the perihelion, its perturbations are generally small (because of the high speed of motion and the large distance from planets of large masses), and the unperturbed coordinates vary very iapidly. In this case Cowell $s$ method is not useful because the interval w should be strongly decreased. It is more useful then to replace this method by Enke's method.

## Annotation I

We can use Titjen's method for the reduction of the number of successive approximations (Sec. 58) during the integration of equations (4). Indeed, noting that on the basis of equations (VII).

$$
f_{n} \quad X_{n}+h q / x_{n}-h_{n} .
$$

where

$$
\begin{equation*}
h=\frac{w^{\prime \prime} k^{\prime \prime}}{r_{n}^{i}} \tag{7}
\end{equation*}
$$

we may write equation (IX), by which we compute $\mathcal{Z}_{n}$, in the following way

$$
i_{n}\left(1+\frac{1}{12} h\right) \cdots S_{n}^{4}-\frac{1}{12} h q / x_{n}
$$

where

$$
S^{\prime}=\frac{1}{12} X_{n}: F_{n}=-\frac{1}{210} r_{n}^{\prime} \cdot \frac{1}{1951} F_{n}^{\prime} \cdots \ldots
$$

Similarly,

$$
\begin{align*}
& n_{n}\left(1+\frac{1}{12} n\right)=S_{n}+\frac{1}{12} n q j y_{n} \\
& \because_{n}\left(1+\frac{1}{12} n\right)=S_{n}+\frac{1}{12} n q f z_{n} . \tag{9}
\end{align*}
$$

We immediately obtain the final values of the quantities $S_{n}^{X}, S_{n}^{y}$ and $S_{n}^{z}$ owing to the smallness of the coefficients $\frac{1}{240}, \ldots$ On the other hand, when we calculate quantity $q$ by means of formula (4)
we can replace the quantities $\frac{1 / 2}{2} \varepsilon_{n}, \frac{1 / 2}{2} \eta_{n}$ and $\frac{1 / 2}{2} \tau_{n}$ inside the brackets by their extrapolated values. This will introduce a slight error because these quantities have small multipilers $\left.\xi_{n},\right\}_{n}$ and $\tau_{n}$. The quantities $\xi_{n}, \ldots$ standing outside the brackets can be replaced by their values which may be obtained by formulae (8) and (9). This yields
where

$$
\begin{align*}
& r_{n}\left(11_{12}^{1} h\right) \quad \because \quad \because \quad r_{n}^{2}\left(1+\frac{1}{12} n\right) \quad \because \quad \overline{r_{n}}\left(1-1 \cdot \frac{1}{12} n\right) \tag{10}
\end{align*}
$$

Therefore, we finally obtain for the calculation of $q$ the following formula

$$
q-\frac{a S_{n}^{\gamma}+3 S_{n}^{v}+\gamma S_{n}^{s}}{1-\frac{1}{12} h f\left(x x_{n}-+\beta y_{n}+\gamma-\gamma z_{n}\right)}
$$

Thus calculating the coefficients $\alpha, \beta$ and $\gamma$ using formulae (7) and (8) by means of theextrapolated values, we can obtain $q$ from equations (II). Then, formulae (8) and (9) yield new and more exact values for $\xi_{n}, \eta_{n}$ and $\mathcal{F}_{n}$.

In the calculation of $q$ by means of formula (II), we usually prefer to extrapolate the denominator rather than find the denominator


The above-mentioned application of Titjen's method to the integration of equations (4) was first suggested by Oppol'cer and was called Oppol'cer's method.

## Annotation II

In the absence of Comrie's tables ${ }^{(1)}$ the rectangular equatorial coordinates of the ferturbing planets may be calculated by formulae (3). In this case however, it is more useful to calculate the perturbations in the ecliptic coordinates. Forthe perturbing planets, these are calculated by the following simple formulae
(1) We quote herein the sonvenient tables, published in: H. Q. Rasmusen, Hilfstaflen fur die numerische Integration der rechtwinkligen Koordinaten eines Himmelskorpers, Astr. Nachr., 260, 1936, 325-376.
The following article includes a table which simplifies the application of Titjen's method in the computation of the coordinates by Cowell's method.
M. Th. Subbotin, Sur le calcul des coordinnees heliocentriques des planets et des comets au moyen des quadratures, Povlkovo Observatory Circular, No. 9, 1933, 15-25.

$$
\begin{aligned}
& x_{1}^{\prime}=r_{1} \cos b_{1} \cos l \\
& y_{2}^{\prime}=r_{1} \cos b_{1} \sin l \\
& z_{1}^{\prime}=r_{1} \sin b_{1} .
\end{aligned}
$$

When the perturbations $\xi^{\prime}, \geqslant$ 'and $\tau^{\prime}$ of the ecliptic coordinates are obtained, the perturbations of the equatorial coordinates are calculated by the following evident relations

PARI THREE
ANALYIICAL METHODS FOR STUDYING PERTURBED MOTIONS

CHAPTER XII

## THE SERIES EXPANSI?N OF THE COORDINATES

OF THE ELLIPTIC MOTION

## 77. Introduction

The equations governing perturbed motions are generally complicated. The analytical integration of these equations is only possible when the perturbing accelerations involved in these equations are explicit functions of the independent variables. Usually, the independent variable in the theory of perturbed motion is taken to be time (or, equivalently, the mean anomally of the perturbed luminary), or the eccentric anomaly of the perturbed luminary, or finally, its true anomaly. The true anomaly is often replaced by the true latitude. The perturbing accelerations can be expressed in terms of the perturbation function in a straightforward manner. The task of integrating analytically equations of motion can thus be reduced to the task of expressing the perturbation function, by a function of one of the above-mentioned variables. The first step which has to be performed is to express the coordinates of the elliptic motion by an expiicit function of time or, equivalently, the arerage anomally. This wili constitute the topic of the present chapter.

The coordinates of the elliptic motion $r$ and $v$, and the function of coordinates, $F(r, v)$, are periodic functions $f(M)$ of the average anomaly $M$ with a period of $2 \pi$. Therefore, any such function $F(r, v)=$ $f(M)$ can be expanded in a Fourier series

$$
\begin{align*}
& \left.f(1)=\frac{1}{2} a_{n} \cdot 1 \cdot a_{1} \cos A i_{-1} \quad \therefore \cos k M \right\rvert\, \ldots . \\
& +h_{1} \sin M+\quad . \quad: \quad \sin k .1 H_{1}+. .,
\end{align*}
$$

Tils series converges for $\varepsilon \cdot$, wes of $M$ since we are only consid-
 is well knmen that, 20 suc; $: \quad \therefore$ the expansion coefficients $a_{k}$ and $b_{k}$ will decay so rapidiy $\because$ : tie products $a_{k} k^{2}$ and $b_{k} k^{2}$ will tend to zero for any arbitrayy cist factor $\alpha$. These coefficients are given by

Series (1) is often replaced by the corresponding Maclaurin series. Indeed, let us put

$$
z=e^{\prime \prime} \text { expi.it. }
$$

where $i=\sqrt{-1}$, then

$$
\begin{aligned}
& 2 \operatorname{cosk} k \quad \text { expish|exp(-ikin) } \\
& \text { 2isinRM- expik.1-..exp(-ik.h) }
\end{aligned}
$$

Hence,

We introduce coefficients $a_{k}$ and $b_{k}$ for negative indices by adopting that

$$
a,=a_{k}, \quad b_{-k}=b_{r} ;
$$

Series (1) will then assume the following final form

$$
\begin{equation*}
f(M) \quad \sum_{i}^{i} P_{k} z^{k} \tag{3}
\end{equation*}
$$

where


$$
\rho_{n}=\frac{1}{2}\left(a_{n}-i b_{k}\right) .
$$

The coefficients of the Maclaurin series (3) are given by the following well-known formula

$$
{ }_{n}-\cdots \underset{n-i}{1} \int \frac{f(n) d z}{z+1}
$$

where the integration is carried over the contour $C$ that encloses the point $z=0$ in the plane of the complex variable $z$. Taking this contour as a sphere of unit radius witn centre at the point $z=0$, we obtain

$$
\begin{equation*}
I_{4} \quad \frac{1}{2 \pi} \int_{i}^{2} \int_{1}^{2}(A) \exp \left(-i R_{k} H\right) d A . \tag{4}
\end{equation*}
$$

This formula is squiv.lerit to equation (2)
For these functions which we will consider later on, the integrals (2) or (4) cannot generally be expressed in terms of elementary functions. They are convenfently expressed in terms of Bessel functions. Hence, we shall start by studying some properties of these functions.
78. The Bessel Functions

Let us consider the following expression

$$
\begin{array}{llll}
4.1 & \ddots & \ddots & \ddots \\
\hline
\end{array}
$$

Since
then by multiplying these absolutely convergent scries, we obtain

We expand this seri $s$ in powers of $z$ and put $x-\beta-n$.

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We then obtain

$$
\begin{equation*}
i i(z)=j_{n}(x) z^{n} \tag{6}
\end{equation*}
$$

where $n$ varies from $-\infty$ to $+\infty$, and $\beta$ from 0 to $+\infty$ if $n \geqslant 0$ and from $-n$ to $+\infty$ if $n<0$. Consequently, the expansion coefficients are given by

$$
\begin{equation*}
J_{n}(x)=\sum_{h=n}^{i}(1 \cdot \beta+n)^{\prime}\binom{x}{2}^{n} \tag{7}
\end{equation*}
$$

if $n \geqslant 0$, and
if $n<0$. These expansion coefficients, $J_{n}(x)$ are known as the Bessel iunctions of indices $n$. The series (7) converges for all values of $x$, and can be considered as a definition of the Bessel function $J_{n}(x)$. Assuming in the latter equation $n=-m$, where $m>0$, we obtain the following relation

$$
\begin{equation*}
J_{-m}(x)=(-1) J_{m}(x) . \tag{8}
\end{equation*}
$$

which indicates that it is possible to consider only the Bessel functions t'rat have positive indices. It alidio follows from equation (7) that

$$
\begin{equation*}
I_{n}(-x) \quad(-1)^{n} J_{n}(x) . \tag{?}
\end{equation*}
$$

We can obtain otiner properties of the Besse 1 functions by means of equation (6). Differentiatirg this equation with respect to $z$, we obtain

Substituting here the expression (6) of the function $\Phi(z)$, and equating the coefficients of $z^{n-1}$ in both sides, we obtain

$$
a+1) \quad \text { ! ! , }(x) \cdots, 1,1
$$

(10)

On the other hand, differentiating equation (6) with respect to $x$, we obtain

$$
\frac{1}{a}: 2-=1, X^{\prime} J_{2}(x)="=\bigcup_{4}^{\prime} f_{7}^{\prime}(x)=1
$$

Equating the coefficients of $z^{n}$ in both sides of this equation yields

$$
\begin{equation*}
J_{n}^{\prime}(x)=\frac{1}{2}\left|J_{n},(x)-J_{n+1}(x)\right| \tag{11}
\end{equation*}
$$

This equation enables us to express any derivative of the Bessel function as a linear combination of these functions. For example,

$$
J_{n}^{\prime}(x) \int_{2}^{1}\left|J_{n}(x)-J_{n i 1}^{\prime}(x)\right|
$$

which may be reduced to

$$
\begin{equation*}
J_{n}^{\prime \prime}(x): \left.=\frac{1}{f}\left|J_{n}:(x) \cdots-2 J_{n}(x) \cdot\right| J_{n i n}(x) \right\rvert\, \tag{12}
\end{equation*}
$$

We shall now show that the Bessel function $J_{n}(x)$ satisfies a linear second-order differential equation. We consider equation (10), which yields

$$
\begin{aligned}
(n-1) J_{n}(x) & =\frac{x}{2}\left|J_{n}:(x): J_{n}(x)\right| \\
(n!1) J_{n: 1}(x) \cdots & : x\left|J_{n}(x): J_{n}(x)\right|,
\end{aligned}
$$

Adding these equàt fons and for simplifying dropping the argument $x$, we obtain

$$
n\left|J_{n},: J_{n+1}\right|-\left|J_{n-1}-J_{n, 1}\right|: \underset{2}{x}\left|J_{n}-2 J_{n} \cdots J_{n \mid:}\right|+2 x J_{n}
$$

We replace the square brackets by the expressions given in equaiions (10), (11) and (12). We then obtain

This differential equation enables us to study the Bessel functions $J_{n}(x)$ for both real and complex values of the indices $n$. It is usually considered as the basic equation in the general theory of Bessel functions.

If we substitute into equation (6)

$$
\begin{equation*}
z-\text { expi". } \tag{11}
\end{equation*}
$$

we then obtain

$$
\exp (x \sin \because)=\sum J_{n}(x) \operatorname{cop}(i n:)
$$

Assuming that both $\varphi$ and $x$ are real, and equating tise real and imaginary parts in the equation, we obtain
where, we have taken equations (8) into consideration. The replacement of $\varphi$ by $\varphi+\frac{\pi}{2}$ yield

In conclusion, we derive some simple integral representations of the Bessel function. Applying formulae (9) to the coefficients of the series (6), and taking into account equation (14), we obtain
or,

$$
\begin{equation*}
J_{n}(x)==\int_{2 \pi}^{1} \int_{i}^{\ddot{n}} \exp (i x \sin ;-i n ;) d x \tag{17}
\end{equation*}
$$

Equating the real parts of this equation yields

$$
\begin{equation*}
J_{n}(x)=\int_{2 \pi}^{1} \int_{0}^{2-} \operatorname{cov}(n \div-x \sin \%) d ; \tag{18}
\end{equation*}
$$

This formula is not useful for calcuiating the Bessel function $J_{n}(x)$ fo: large values of the index $n$, because the integrand will have I.arge maxima and minima values. Moreover, this formula does not stress the property that the Bessel function $J_{n}(x)$ heinaves like $X^{n}$ for small values of $n$ and $x$, while this property is very important in astronomy applications. It is thus useful to rewrite equation (7) in the following manner

For arbitrary integral values of $n$ and $\beta$, the following relation holds

Therefore,
or, finally

This formula isfree from the shoricomings of equation (18) as previously mentioned.
79. The Computation of Bessel Functions

In the problems wo are going to consider, we have to compute for a

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ORIGINAL $P^{\prime \prime}$
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given value of $x$, all the function $; J_{0}(x), J_{1}(x), \ldots$ that differ from zero up to within the accepted number of decimals. We will now show the simplest and most convenient method of doing this. We first not $\geqslant$ that formula (19) yields

$$
I_{n}(x) \quad 1.3 .5{ }^{x}{ }^{n} \cdot\left(e_{n}^{n} 1\right)
$$

This inequailty evables us to find the maximum value for $n_{1}$ for which the fanction $J_{n}(x)$ differs from zero within the accepted accuracy. The first method

We can write equation (7) in an unfolded form as follows

These series are convenient for the purpose of the rapid calculation of the Bessel functions when the values of $x$ and $n$ are not larye.

It is sufficient to compute the values of only two functions, e.g. $J_{0}(x)$ and $J_{1}(x)$, and then find the values of other functions by means of the successive application of formulae (10). For example,

We shoulc however point out, that owing to the presence of fartors $\frac{2}{x}$, $\frac{4}{x}, \frac{6}{x}, \ldots$, we will have a progressive loss of accuracy, which will be all the more significant for smaller values of $x$.

## The second method

Let us introduce the ratio $p_{k}$ of two velghbouring Bessel functions, defined by the relation

$$
J_{1}(x) \quad p_{1} f_{0}(x), \quad J_{1}(x) \quad n_{1} f_{1}(x) \ldots \ldots I_{11}(x) \quad p_{n} J_{n}(x) .
$$

$$
\begin{aligned}
& I_{1}=J_{0} p_{1} \\
& J_{2} J_{11} p_{1} F \\
& \cdots \cdot \dot{c}_{0} \\
& f_{4}-J_{11} p_{1} p_{2} \ldots p_{n} .
\end{aligned}
$$

$$
(22)
$$

Thus, our task is reduced to the computation of $J_{0}(x)$ on one hand, and to the computation of $p_{1}, p_{2}, \ldots p_{n}$ on the other hand. The function $J_{0}(x)$ can be computed by means of the sertes (20), or, if the value of $x$ is large, by formula (19). Let us now turn to the computation of $p_{i}, p_{2}, \ldots, p_{n}$. Formula (10) leads to

$$
\begin{array}{ccc}
\therefore & I_{1} & \prime . \\
x & J_{k} & \prime
\end{array}
$$

or

$$
\begin{array}{ll}
\because! & 1  \tag{1:}\\
\therefore & p_{k}
\end{array} p_{1}
$$

Substituting here for $k=n-1, n-2, \ldots, 1$, we obtain

$$
\left.\begin{array}{ccccc}
1 & 2 n & 2 & n_{n} \\
p_{n} 1 & & x & & n_{n} \\
1 & 2 n & 4 & & \\
p_{n} & : & x & \cdots p_{n} & \\
\cdot & \cdots & \cdots & \\
1 & \cdots & & & \\
n_{1} & x & p_{n} & &
\end{array} \right\rvert\,
$$

- 

These formulae allow us to compute $p_{n-1}, p_{n-2}, \ldots, p_{1}$ in a sfaple manner without any loss of accuracy, provided that $p_{n}$ is known. However, from the same equation (23), we obtain
so that $p_{n}$ nay be represented by the following contfinued fraction

$$
\begin{align*}
& x \tag{25}
\end{align*}
$$

which converges more rapidly for larger values of $n$.

## The third method

Adding equations (16) term by term, we obtain
where
$f(\dot{y})$ - $\operatorname{cosix} \cos ; 1 \cdot \sin (x \cos y)$

$$
\begin{aligned}
& c_{0}=: J_{11}(x), \quad c_{1} \quad 2 . l_{1}(x), \quad i=\quad 2 . i(x), \quad \text {, } \quad i f_{1}(x) \text {. } \\
& c_{1} \cdots: J_{1}(x) . \quad c_{5} \quad 2 J_{6}(x), \quad \therefore,-2 J(6), .
\end{aligned}
$$

Computing the function $F(\varphi)$ for a series of equa11.y spaced values of $\varphi$, and applying the usuai formulae of the harmonic analysis, we obtain $\mathrm{J}_{0}(\mathrm{x}), \mathrm{J}_{1}(\mathrm{x}), \ldots$.

Let us for example assume that $J_{7}(x), J_{8}(x), \ldots$ is equal to zero. We then introduce the following notation


Then,

$$
\begin{aligned}
& \text { i. i. i ! ll. lc, lii } \therefore \quad 1 \\
& \therefore \quad \therefore \quad \therefore \quad \therefore \quad 2,-1 ; \quad i-1,
\end{aligned}
$$

from which we obtain $C_{0}, C_{2}, C_{4}$ and $C_{6}$ stmilarly, putting

$$
A^{\prime} \quad y_{1} \quad y \quad y^{\prime} \quad y_{l} \cdots y, \quad \because \quad y \cdot y_{1}
$$

$$
\begin{aligned}
& \therefore c_{1} \therefore 3 c \cdots A^{\prime}+C^{\prime}: \quad \because c_{1} \quad \because, 2 \cdot=A^{\prime}
\end{aligned}
$$

from which we obtain $C_{1}, C_{3}$ and $C_{5}$.
For checking, we can apply any particular form of formulae (15)
and (16), e.g. any of the following equations

$$
\begin{aligned}
& : \quad f_{1}(x): 2 f_{2}(x): 1 f_{1}(x) ; \cdot . \\
& 1-I_{1}(x)-2 J_{-}(x): 2 J_{1}(x) \cdots \cdot . \\
& \text { vill } x \quad \because J_{1}(\lambda)-\ddot{O}(x) \quad \because(x) \quad . \quad .
\end{aligned}
$$

Ten-figure tables of the functions $J_{0}(x)$ and $J_{1}(x)$ were given by Bessel ${ }^{\text {(1) }}$ for values of $x$ varying from 0.00 to 3.20 by increments of 0.01 . Hansen (2) gave six-figure tables for these functions for values of $x$ varying from 0.0 to 20.0 by increments of 0.1 .
80. The Expansion of the Excentric anomaly and its Functions by Multiples of the Average Anomaly

It follows from the Kepler equation

$$
\therefore-\infty+111 \quad \therefore 1
$$

that for all values of the eccentricity satisfying the following condition

$$
11, c \times 1
$$

(1) F.W. Besse1, Untersuchung des Teils der planetarischen Storungen, welcher aus der Bewegung der sonne entsteht, Abhaadlungen des Berliner Akademie 1824.
(2) P.A. Hansen, Ermittelung der absoluten Storungen in Blupsen von Belicbiger Exzentrizitat und Neigung, Schriften der Sternwarte Seeberg (Gotha), 1843.
the eccentric anomal.y F is a finfte and continuous (as well as all its derivatives) function of the average anomaly $M$. When $M$ is increased by $2 \pi$, the eccentric anomaly also increases by 277 . Consequently, any periodic function of $E$ having a period of 27 will also be a periodic function of $M$ having the same period.

Let us consider thefunction $\cos \mathrm{mE}$, where m is an integer. This function is evidently a perindic and even function of $M$. We can thus assume that
where, on the basis of equations (2),

In partiaular, by excluding $M$ by means of equation (26), we obtain for $k=0$

If $m>1$, each of these integrals is equal to zero. Then,

$$
a_{n}^{\prime \prime}: \quad 1
$$

If $m=1$, it is easy to see that

If $k \geqslant 0$, then partial integration ylelds

Substituting here for the value of $M$ given by equation (26), we obtain

$$
\begin{aligned}
& f \cos \mid(k \quad m) t: \text { Re vill tidit: . }
\end{aligned}
$$

Using equation (18), we finally obtain for $k>$ ?

Similarly, we can prove that the coefficients of the series
are given by

Ne note that coefficients of the two series, given by equations (27) and (28) can simult aneously be obtained by considering the expansion of the function $\exp (i m E)$.

When $m$, the series (27) and (28) can evidently be representcd in the following form

When $m=1$, then the series (27) can be transformed by means of equation (11) Into che following form

Similarly, using equation (10), we outain

$$
\begin{equation*}
\text { :in: } \quad 1 \tag{1}
\end{equation*}
$$

Substituting these expansions into equation (26) and into the follewing formula

$$
r a(1-r(\sin (i) .
$$

we obtain'

We now derive an expansion for the square of the radius vector. Since
and

We then obtain

$$
\because \searrow
$$

This formula can be derived in a aimpler way, if we note that
and make use of the expansion given by equation (30). Similarly, noting that

$$
\begin{array}{ll}
\because: \quad \text { i }- \text { ert: } & \ddots \\
\therefore i i
\end{array}
$$

we then obtain, using formula (31),

$$
\begin{equation*}
r \quad!\cdots, \Delta f_{1}(!r) c_{1}, b: i \tag{1;4}
\end{equation*}
$$

This equation enables us to find the expansion coefficients

We shall not derive the complicated expressions of the coefficients $g_{1}, g_{2}$, ... in terms of the Bessel functions. For practical purposes, It is sufficient to expand each of these coefficients in powers of e.

These expansions will be given in Section 82. In the following we shall confine ourselves to the evaluation of $g_{0}$ only.

It follows from equation (35) that

$$
\left.\therefore \quad \begin{gathered}
1 \\
\pi
\end{gathered} \right\rvert\,\left(\left.\begin{array}{l}
a \\
r
\end{array}\right|^{2} d u\right.
$$

The integral of area

$$
r^{2}=\begin{aligned}
& d n \\
& u l
\end{aligned} \cdots V^{\prime} \cdot m V^{\prime} u\left(1-c^{-2}\right)
$$

can be represented in the new form:

$$
\operatorname{div}\left(\frac{11}{1}\right)^{2} \operatorname{li}^{\prime} 1 \quad i^{\prime}
$$

$$
1.361
$$

because

$$
M=k V l+\dot{m} a^{-}:\left(\begin{array}{ll}
l & l_{0}
\end{array}\right)+M H_{1}
$$

Consequently

$$
x_{n} \cdot(1 \cdot c)^{-1}: \frac{1}{n} \int_{i}^{3} d n \cdot:(1-c)^{1}
$$

and therefore,

Evaluating this coefficient by the simple squaring of equation (34) and comparing the results, we obtain

$$
11, \quad \because \quad 1 \quad \because \because 1 / 1 \cdot 1
$$

## Annotation

In the expansions derived in this section, and in most of the applications in astronomy, Bessel functions are often encountered in one of the two following forms

$$
\begin{aligned}
& \text { 2J." }(k=)=
\end{aligned}
$$

We point out the following particular cases, which are useful to have in a readily available form

$$
\begin{aligned}
& \frac{2}{6} J_{1}(c)=1-\frac{e^{\prime}}{8} \cdot \frac{e^{\prime}}{192}-9.9
\end{aligned}
$$

81. The Transformation of a series in Yultiplies of the Ecentric anomaly

## into a series in multiples of the average anomaly

Let $S$ be a geriodic lunction of $E$ having a pericd of $2 \pi$. We asume ti: this function is continuous and has continuous derivatives sn that it : can be expanded in a Fourier series

As we have pcinted out, $S$ will also be a periodic function of $i t$ and will have the same period of $2 \pi$. Hence it can also be expanded into the series

$$
\therefore-\frac{1}{2} A_{0} ; A_{1} \text { со. } 11: A_{2} \text { con } 2.11 ; \ldots, A_{1}, \ldots 1 i \ldots .
$$

Our problem is to find the expansion (38) in the case when the coefficients of the expansion (37) are known.

Substituting for cos $n E$ and sin $m E$ their expressions givan by formulae (27) and (28), we ootain

$$
\begin{aligned}
& A_{n} \cdot a_{n} \quad i a_{1} a_{v}^{!} \\
& A_{k} a_{1} a_{k}^{\prime}: a_{2}^{\prime 2}+\ldots . \\
& b_{k} b_{1} b_{k}^{1}+b_{2} \dot{\sigma}_{n}^{\prime}+. .
\end{aligned}
$$

Replacing $a_{k}^{m}$ and $b_{k}^{m}$ by the values found in the previous section, wn obtain

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OF POUR «——

$$
\left.\begin{gathered}
-321- \\
k A=\sum_{m=1}^{A_{11} a_{11}-a_{1} e} \\
\left.\left.R B_{k}=\sum_{m=1} m a_{m}\left|J_{k m}(k, j)-j_{k+m}\right| k e\right)\right]
\end{gathered} \right\rvert\,
$$

Therefore, the transformation of the series (37) into the eries (38) is reduced to the computation of the following quantities

$$
\begin{array}{lll}
J_{0}(e), & J_{1}(e), & J_{2}(e), \ldots \\
J_{v}(2 e), & f_{1}(2 e), & J_{2}\left(2 e^{\prime}\right), \ldots
\end{array}
$$

This can be done algebr.ically up to within a given pover of e, or numerically by means of the formulae given in Secs 78 and 79.

Cauchy had suggested another wethod for treating the problem under consideration. The method is as follows. Let us introduce the following notation

$$
y=\exp i \ell, \quad \quad=\text { exp } i m
$$

We replace the expansions (37) and (38) by the corresponding Maclaurin series

$$
\therefore \quad \because \quad \because p_{k} y^{*}
$$

and

In order to calculate the coefficients $P_{k}$, we consider formula (4) which yj.elds

$$
\begin{equation*}
\because j_{i} \quad \dot{i} \tag{:i}
\end{equation*}
$$

On the basis of equation (26),

$$
\begin{aligned}
& \angle{ }^{*}=\exp (i k: h)=\exp (i k E \cdot \text { the } \sin E) \\
& =y^{2} \exp \left|\begin{array}{l}
\operatorname{se}\left(y-y^{-1}\right) \\
2
\end{array}\right| \\
& \frac{d N}{d t}-1-\frac{r}{2}\left(y+y^{-1}\right)
\end{aligned}
$$

## Consequently

$$
\begin{aligned}
2 \pi P_{k} & =\int_{1}^{0} S y^{k} \exp \left|\frac{k e}{2}\left(y-y^{-1}\right)\right|\left|1-e_{2}^{c}\left(y+y^{1}\right)\right| d E \cdot \\
& =\int_{0}^{3} T y^{-k} d E
\end{aligned}
$$

The latter equation is nothing else but the result of application 0 the generai formula (4) to the finding of the coefficients of expansion of the function

$$
r=S\left|1-\frac{\dot{c}}{2}\left(y: \vdots y^{-1}\right)\right| \operatorname{cxp}\left|\begin{array}{ccc}
k e \\
2
\end{array}\left(y \cdots y^{\prime}\right)\right|
$$

in powers of $y$. We thus obtain Cauchy's first rule. In order to obtain the coefficients $P_{k}$ of the series (40), it is necessary to expand the function $T_{1}$ in which $S$ is replaced by the series (37), in powers of $y$; the coefficients of $y^{k}$ will be equal to $P_{k}$.

On the other hand, sirne

$$
\frac{d z}{d M}=i z
$$

then equation (41) yields

However,

$$
d S=\begin{aligned}
& d S d y \\
& d E^{\prime}
\end{aligned}=\frac{d y}{d t}=\cdot i y \frac{d S}{d y}
$$

Therefore, expressing again $z^{-k}$ in terms of $y^{-k}$, we obtain

$$
2 \pi P_{k}=\int_{0}^{-r} U y^{k+1} d E
$$

where

$\left.1$| $!$ | 1 |  |  |
| :--- | :--- | :--- | :--- |
| $!$ | 1 |  |  | \right\rvert\,

This expression of $P_{k}$ proves Cauchy's second rule. In order to obtain the coeificients $P_{k}$ of th series (40), it is necessary to expand the function $U$ in powers of $y$ and take the coefficients of $y^{k-1}$.

The functions which we usually have to expand are almost in all cases very simply expressed in terms of the combinations $y+y^{-1}$ or $y-y^{-1}$. The application of theabove-mentioned rules leads to the use of the so-called Cauchy's numbers. These are the coefficients $\mathrm{N}_{-\mathrm{p}, \mathrm{j}, \mathrm{q}}$ in the expansion
where $j$ and $q$ are non-negative integers.
At the present time, all the expansions applied in celestial mechanics are available in readily available forms. Therefore, we shall not consider here the properties of Cauchy's numbers ${ }^{(1)}$.
82. The expansion of some functions of the conrdinates of the elliptic motion

We have already found the expansions of the radius of vector in a serifs, (32), in multiples of the average anomaly. In the following, we
(1) See F Tisserand, Traite de Mrcánique Celeste, I, Paris, 1889, 234-237, and references cited therein.
oltain a similar expan-ion for the true anomaly.
We first of all express the true anomaly in terms of the eccentricity by using the following formula

## Putting

we obtain
要化 - ! !
or

Assuming that $\beta=\frac{\mu-1}{\mu+1}$, we obtain

Taking the legarithm of both sides yields
or

Applying this formula to equation (42) ylelds the following expansion
where

Rewriting equation (42) in the form

and applying again formula (43), we obtain

Substituting for $E, \sin E, \sin 2 E, \ldots$ in equation (44) the corresponding expressione given by equations (31) and (28), we obtain the important expansions which defines the equation of the centre, namely
where

In the following, we give the values of $\mathrm{H}_{1}, \mathrm{~F}_{2}, \ldots$ in seconds of arc where we use the logarithms instead of the numerical coefficients.


For convenience, we write equation (22) which determines the radius vector in the following unfolded form

Then

$$
\begin{aligned}
& \text { (j) } \left.\quad 2\binom{c}{\therefore}^{\prime}-\begin{array}{ll}
\prime 1 \\
\therefore \\
2
\end{array}\right)^{\prime} \cdot a\binom{e}{2}^{\prime}
\end{aligned}
$$

Replacing the numerical coefficients by their logarithms, we obtain

$$
\begin{aligned}
& \left.U_{5 n} \cdot|9.7 .3|\right|^{10}-|110,|1, i||^{12}
\end{aligned}
$$

where we have to subtract 10 fron each logarithm given here.
It is useful to note that the expansion (46) may be obtained in another way. Indeed, substituting series (3) -nto formula (36) and integrating, we obtain
(19i)

In the series-expansion of the perturbation function, we make use of the expansion of the following functions
where $p, n$ and $m$ are integers such that $p$ and $m$ either take positive values or are equal to zero. The calculation of the expansion of such functions up to a given power of $e$ is simple enough. We have for example

Replacing $\frac{r}{a}$ and $v-M$ by the series (47) and (46), we obtain the required result.

The coefficients of the expansions of the functions (48) up to $e^{7}$ were given by Leverrier ${ }^{(1)}$. Cuyley ${ }^{(2)}$ gave the coefficients of expansion of the functions within the same accuracy.

$$
\left(\begin{array}{ll}
r \\
u & -i
\end{array}\right)^{\prime \prime} \cdot n \cdot n
$$

where $p=0,1, \ldots, 7$ and $m=0,1, \ldots, 7$, and also for the functions

$$
\binom{r}{a}^{n} \cdots m m
$$

where $n=-5,-4, \ldots,-1,1, \ldots, 4$, and $m=0,1, \ldots 5$.
Some of the most commonly used expansions are given in tables $I$ and II at the end of this volme. These tables give the coefficients of different powers of $e$ in the power series for the coefficients $C_{k}^{n, m}$ and $S_{k}^{n, m}$ involved fin the following expansions

For exampoe, table II shows that
(1) U.I.J. Leverrier, Recher ites astronomiques, Annales de l'nbservatoire de Paris, J, 1885, 343-365.
(2) A. Cuyley, Tables of the Developments of Functions in the Theory of El.liptic Motion, Memoirs of the R. Astron. Society. 29, 1869, 191-306.

In conclusion, we give the following expansions, which can easily be obtained from equations (47).


## Annotation

The coefficients of the series (46) and (46') have a very complex structure. It is much easier to express the coefficients of the expansion of the equation of the centre in multiples of thetrue anonaly. In order to derive this expansion, we consider the following formulae

These formulae lead to


Since
where

$$
1 \quad \vdots \quad . \quad \text { mindx }
$$

then


Therefore,

Replacing E by expression (45), we finally ottain

## 83. Hansen's Coefficients

Instead of separately considering the two expansions given by equations (49), it is possible to stuedy only the following Maclaurin series


The coefficients $X_{k}^{n}, m$ of this series are sometimes called the Hansen's coefficients since Hansen was thr first to give general expressions of these quantities in the form of series-expansions in powers of $\beta$. An alternative and simpler derivation of Hansen's formulae was suggested by Tisserand ${ }^{(1)}$.

The simplest way to obtain these coefficients is to apply Cauchy's first rule. Let us express the function

$$
\therefore=\left(\begin{array}{l}
r \\
a \\
\vdots
\end{array} i^{n}\right.
$$

in terms of $y$. Since
(1) F. Tisserand, Traite de Mechanique Celeste, 1, 1889, Ch. XV.
then, according to Cauchy's first rule, the coefficient $X_{k}^{n, \text {, }}$. will be equal to the coefficient of $y^{k}$ in the expansion of the following expression

It is easy to sec that

$$
{ }_{a}^{r}=11-y^{1}(1-\cdots)\left(1-y{ }^{\prime}\right),
$$

where, we denote as previous ty

$$
\hat{\beta-v} \begin{array}{cc}
e & 1-v-e^{2} \\
1, v & e
\end{array}
$$

On the other hand, the relation
may be rewritten in the following way

$$
\begin{array}{llllll}
x & 1 & 1 & \ddots & y & -1 \\
x & 1 & 1 & -1 & y & 1
\end{array}
$$

Therefore,

Consequently,

Using the binomial formula, we can easily calculate the expansion
and obtain them in the following form
where $F(a, b, c, x)$ is the hypergeometric function. Since,
then the unknown coefficient of $y^{k}$ in the expansion of the function $T$ is equal to

This formula enables us to obtain the coefficients of the expansion $\tau_{k}^{n, m}$ and $S_{K}^{n, m}$ in powers of eccentricity.
84. On the Convergence of the Series-Expansions of the Coordinates of the E11iptir Mot lon

In the previous sections we obtained the expansions of different functions of the eccentric anomaly $E$ in Fourier series, developed by muitiples of the average anomaly M. On the basis of Dirichlet's theorem, these series converge for all values of $M$ and $:$ only if $e<1$ as in this case where the expanded functions and their derivatives are continuous. However, due to the complexity of the expansion coefficients, these coufficients are usually expanded in powers of e in which terms higher than a given power are dropped. Accordingly, we are practically dealing with power series, developed in positive powers of $e$, the expansion coefficients of which are periodic functions of $M$ with a period of $2 \pi$. The radius of convergence of such a series is some
function $\varphi(M)$ of theaverage anomaly, Our task is to find the minimum value of this function, $\varphi(M)$, when the variable $M$ varies from 0 to 2 Let us consider an arbitrary function $F(E)$ of the eccentric anomaly and investigate its dependence on $e$ and $M$, implied by the Kepler equation.

$$
\therefore-r, i: i \quad, i \quad 1,11
$$

For a given $M$, this function is a holomorphic furn+ion for all vailues of e for which the derivative
is finite. Hence, the general singular points of all the functions $F(E)$ are given by the following equation

1 rint 1 . , '
which is to be solved simultaneously with equation (51). The only exception are those functions for which the product $F(E)$ sin $E$ is either zero or infinity for values of the variables satisfying condition (52). We shall not consider these functions now. '. .e radius of convergence
(M) of all the functions under considerations will be equal to the least of the roots e of equations (51) and (52).

Let us now study the function $\varphi(M)$. We primarily note that

$$
\because(-1 \|) \quad \because 1 \|
$$

Indeed, replacing $E, M$ and $e$ in equations (51) and (52) by $\pi \pm E, T \pm M$ and -e does not violate these equations. Hence, the above-mentioned change in the variade M will transform each singular point eo intol the singular point -e that has the same modulus. Consequently, the radius of the circle of convergence will not be changed.

It is somewhat more difficult to prove another, propurty of the

> OPTGNAL
> OF PGR
function $\varphi(M)$, stating that the minimum value of this function is squal to $\left(\frac{\pi}{2}\right)$. Ac.ording to Poincare, we consider the function

$$
f(E)-c x p(2 i m E)
$$

where m is an integer. The derivative of this function satisfies the above mentioned conditions. The c.pansion of this function in a power serier can be done easily usjr. $\mathrm{g}^{-}$-mulae (27) and (7) wh:in yield

$$
r\left((G)=\sum_{k}^{1} \sum_{k} J_{2 m}(k e) z^{*}=\sum_{k}^{1} \sum_{m} z^{k} \sum_{p l(\beta+k-2 m)!}^{(-1)^{k}}\binom{k e}{2}^{A \cdot m+k \cdot t} .\right.
$$

By adopting that

$$
F(t)=\|\left(M_{1}, c\right)
$$

ard considering the sum

$$
\begin{equation*}
\Gamma(A, C)=\|(M, C) f+(=+M, C) . \tag{53}
\end{equation*}
$$

Evideutly,
since all terms with even powers of $z$ are carcelled. Let $M=M_{1}$ be an arbitrary given value of the average anomaly, not equal to $\frac{\pi}{2}$. We $d$ note by $e_{1}$ a real number which satisfies the following condition

$$
\begin{equation*}
f\left(m_{1}\right) \cdot l_{1} \quad 1 \tag{0.0}
\end{equation*}
$$

In this case, the series $\Phi\left(M_{1}, e_{1}\right)$ is evidently divergent. he shall now prove that the sum (53) will diso diverge for the values of these variables. indee?, the particular points $e_{0}$ of the function ( $M_{1}$, e) corl espond to the particu'ar points $-e_{0}$ of the function $\Phi\left(\Pi+M_{1}, e\right)$, aithough $-e_{o}$ cannot be a particular point of $\left(M_{1}, e\right)$
since the substitution of $e$ by -e in equations (51) and (52) replaces $k$ by $\prod_{ \pm} \mathrm{M}$. Hence, the particular point of one of the terms of the expression (53) will definitely be a singular point of the whol? sum since $M_{1} \neq \pi \pm M_{1}$ orce $M_{1} \neq \frac{\pi}{2}$. Consequently, the series $\Psi\left(M_{1}, e_{1}\right)$ diverges, if condition (55) applies. Comparing the terms of this series with thecorresponding terms in theseries

$$
\begin{equation*}
y^{\prime}\left(-\frac{\pi}{2}, i e_{1}\right)=4\left(-\frac{r}{2}, i e_{1}\right)+4\left(+\frac{x}{2}, i e_{5}\right) \tag{F}
\end{equation*}
$$

Evidently, the absolute values of the compared terms will be equal. It the same time, the arguments of the terms of $\mathcal{\Psi}\left(M_{i}, e_{1}\right)$ will be different although che arguments of all terns of the series $\mathcal{F}$ (- $\frac{\pi}{2}, j e_{1}$ ) are equal. Indeed the arguments of each tern of the latter series ara equal to

$$
-\frac{\pi}{2} 2 h+1+3+\cdots(2 h-2 m+2 .)=-=-m \pi
$$

as one ca.l easily see from equation (54). Therefore once the series $\mathscr{I}\left(M_{1}, e_{1}\right)$ diverges, the expansion of the function (56) also diverges. It then follows that for at least one of the functions $\overline{\mathcal{I}}\left(\frac{\pi}{2} \text {, ie }\right)_{1}$ the series expansion in powers of the eccentricity diverges, if it diverges for $\Phi\left(M_{1}, e_{1}\right)$. In other words,

$$
p\binom{\pi}{2} \div p\left(M_{1}\right)
$$

which wis required to prove.
Thus, in order to find the minimun value of the runction, it is necessary to find the root $e_{o}$ of the equations

$$
\begin{aligned}
& 1-c_{0} \cos E_{0}=0 \\
& E_{n}-c_{0} \sin t_{0}=
\end{aligned}
$$

that has the least absolute value. We then obcain

$$
\min \xi(1,1) \sim c_{0} .
$$

These equations yield

$$
t_{0}-\lg t_{1}=\frac{\pi}{2}
$$

or, putting $E_{o}=\frac{\pi}{2}-\epsilon$,

$$
\begin{equation*}
\text { efotge } 0 . \tag{ix}
\end{equation*}
$$

We shall only consider thecomplex roots of equation (56). The real roots of this equation yields

$$
\therefore=\cos _{1,1}^{1} \mid>1
$$

and are thus not interesting to us. To each root $\epsilon$ of equation (56) there will be a corresponding conjugate root $\epsilon^{\prime}$. This equation must thus have at least two roots. Considering any pair of roots $\varepsilon$ and $\epsilon^{\prime}$ of equation (56) and construct the auxiliary functions
;-

These functions satis.iy the following equations
dr

Consequently,
or

We integrate this equation from 0 to 1 . Since at $u=0$

$$
d==0, \quad d q^{\prime}-0,
$$

and at $u=1$

$$
\begin{aligned}
& d_{f} \\
& d u
\end{aligned}{ }^{2}-\operatorname{sill}:-\cos : \quad \frac{d_{f}^{\prime}}{d u}=\cos 2^{\prime}
$$

then on using equation (56), we obtain

$$
\left(e^{\prime}-\cdots z^{2}\right) \int_{i}^{!} \varphi_{i}^{\prime} d u=0
$$

It follows that for each pair of complex conjugate roots $\mathcal{E}$ and $\mathcal{C}^{\prime}$

$$
e^{\prime}=s^{2} \quad .0,
$$

because in this case $\varphi$ and $\varphi^{\prime}$ will also be complex conjagate numbers and thus

$$
\because \because=: 0 .
$$

The latter equation holds true only for either real or imaginary values of $\epsilon$ and $\mathcal{E}^{\prime}$. Thus, equation (56) has only pure and real, and pure and imaginary roots. Since thefirst case is not of interest to us, we shall search for those imaginary ror,ts having least absolute vaiues. Substituting Into equation (56)

$$
=i!
$$

we obtain

$$
\text { s! } \because: 0 l l: 11
$$

winch yields

$$
\therefore=1.1916 \pi 561103: 77.84 \ldots
$$

Since

$$
\min ;(11)-i_{s}: \begin{gathered}
1 \\
\text { slli a } \\
\text { shin }
\end{gathered}
$$

then the radius of convergence of the expansion in powers of the eccentricity will be equal to

$$
0.60: 31.314 .314: \text {. . }
$$

This is the theoretical limit of convergence of the obtained series. Naturally, these series loose their practical value at 2 much earlier stage.

## Annotation

We have p.:oved that the expansions of the functions of the coordinates of the elliptic motion in tigonometric series developed by multiples of the average anomaly are convergent for all values of $M$ and for all values of e satisfying the condition

$$
0 \cdot \quad \ll 1
$$

On the other hend we have just seen that wlen these expansions are developed in powers of 0 , then they converge only inside the interval

This change in the radius of convergence is related to the fact that when the Bessel functions are expanded in powers of the eccentricity as
the accuracy of approximating them by the leading terms of the expansions decreases with increasing values of $k$. For example the ratio of the second terix of this expansion to the first termi equals in absolute value to

$$
\begin{gathered}
k, r^{2} \\
2(\because i=1
\end{gathered}
$$

and thus fncreases to infinity with incroasing values of $k$.
85. The Calculation of the Longitude and Latitude of a Planet

Let $u s$ denote by $v$ the longitude of a planet in an orbit. This longitude is expressed in terms of the previously used guantities by the following relution

$$
\begin{equation*}
u=\pi: u \cdot \because: \omega \because \iota \quad!\quad, u \tag{in}
\end{equation*}
$$



$$
\begin{aligned}
& 4(!-i) \quad \text { coritgu } \\
& \sin b \quad \sin i \sin u
\end{aligned}
$$

We mace use of equation (43) in order to deternine $l$ from equation (58). Since in the present case,

$$
; \begin{array}{lll}
\cos : 1 \\
\cos 1 & 1
\end{array} \quad \text { In } \frac{1}{2},
$$

then

Taking equation (57) inio account, we may write

$$
l w ; R .
$$

where

We shall call the d:fference $R$ between the longitude and latitude in the orbit the recuccion to the ecliptic. For a constant orbital slope $i$, this quantity may be tabulated by the argument $u$. A table which gives the heliocentric latitude $b$ by the argument $u$ may be similarly constructec.

## CHAPTER XIII

## THE SERIES - EXPANSION OF THE PERTURBATION FUNCCION

86. Introduction. Expansions in Powers of the Mutual Slope

In order that the differential equations which define the perturbations (Sec. 15) can be integrated in a general and not in a particular form, it is necessary to have an analytical expression for the perturbstion function

$$
i^{\prime}-k^{2} m^{\prime} R_{61}, \quad R^{\prime}=k^{2} m R_{1,1} .
$$

where

$$
\begin{align*}
& u_{0,1}=\frac{1}{1}-\begin{aligned}
& x x^{\prime}+f+y y^{\prime} \\
& r^{\prime}: \approx z^{\prime}
\end{aligned} \\
& R_{.0}=\frac{1}{\lambda}-x x^{\prime}+y^{\prime}+z^{\prime}+z^{\prime} \tag{1}
\end{align*}
$$

in terns of the orbital elements $a, e, \ldots, a^{\prime}, e^{\prime}, \ldots$. In this Chapter, we only consider the most important methods for obtaining these expressior. which wiil indispensibly have the form of infinite series.

First of all, we shall consider the series expansion of the quantity

$$
\begin{equation*}
د^{-1}=\left(r:-r^{\prime}: \quad 2 r r^{\prime} \operatorname{rov} / f\right)^{\prime} \tag{2}
\end{equation*}
$$

where $H$ is the angle between the radius vectors $r$ and $r^{\prime}$. Tris guantity is known as the principal part of the fertuibation function. lts expansion is the most difficulv part of our problem. The expansions of the other parts of equations (1) are relatively simple.

The expression of the radius vector in terms of time and orbital elements has been studied in the previous chapter. We now consider how to find the angle $H$. Referring to figure ll, we find from the triangle $\Omega, \Omega_{1}$ and $N$ the sides $N$ and $N_{1}$ and the angle $J$ (we are keeping the notations of Sec. 68) either by means of formulae (16) of Chapter XI, or
by means of the following relations

$$
\begin{align*}
& \sin J \sin N=\sin i^{\prime} \sin \left(y^{\prime}-4\right) \\
& \text { UIII } J \cos N=\cos i^{\prime} \sin /-\sin i^{\prime} \cos i \cos \left(\mathbf{Q}^{\prime}-\mathrm{a}(1)\right. \tag{3}
\end{align*}
$$

$$
\begin{aligned}
& \text {, in } \left.J \cos N_{1}=-\cos i^{\prime} \sin i f \sin i^{\prime} \cos i \cos \left(\Omega^{\prime}-2\right), \quad\right)
\end{aligned}
$$

which are consequences of the main theorems of spherical trigometry. We can then represent the longitudes $w$ and $w^{\prime}$ of planets $P$ and $P^{\prime}$ in their orbits in the following way

$$
4^{\prime} \quad: \mathbb{K}^{\prime} . \quad \therefore \quad \cdot \vdots \|
$$

assuming that

$$
=\therefore \quad \therefore \quad \because \quad \cdots
$$

and denoting by $W$ and $W^{\prime \prime}$ the longitudes which are measured from the intersection point of the orkits. Consequently, referring to triangle NPP', we obtain
or

$$
\begin{equation*}
\text { a, } H=\mathrm{cos}\left(W^{n}-W^{\prime}\right)-2: \sin W^{\prime} \text {, in } U^{\prime \prime} \tag{f}
\end{equation*}
$$

where

$$
s=\sin \frac{J}{2}
$$

We substitute this expression of cos $H$ into equation (2) and write $\Delta^{-1}$ in the following way

$$
\mid r^{2} \ldots r^{\prime}: 2 r r^{\prime} \cos (W-W)^{-}:\left\{\begin{array}{l}
W^{\prime}+r^{\prime} \sin W^{\prime} \sin W^{\prime \prime} \\
1 \\
r^{2}+r^{\prime}:-2 r r \cos \left(W^{\prime \prime}-W^{\prime}\right) \mid
\end{array}\right.
$$

We only consider the case in which the second tarm invide fles curved brackets is always less than a proper fraction. Since this * is in absolute
values less than

$$
\begin{aligned}
& 4 s^{2} r r^{\prime} \\
& \left(r--r^{\prime}\right):
\end{aligned}
$$

than this condition will. be satisfied, as it is easy seen, for all the large planets of the solar system. Indeed, the maximum value of angle $J$ betreen the planets in orbit (occurring for Mars and Mercury) equals only $12^{\circ} 30^{\prime}$, which gives $\sigma^{2}=0.0118$. On the other hand, the difference $r$ - $r^{\prime}$ for each pair of planets will always be greater than a given quantity which will be the greater, the greater the product rr'. We shall not consider the cases when the above mentioned condition is not $m$ met.

We thus expand the second factor in a series. Applying the binomial formula and assuming
(1) A more general methed of expansion of the perturbation function, which holds for arbitrary slopes but is in turn much more difficult, has been given by Tisserand: F. Tisserand, Trafte de Mechanique Celeste, 1, Ch. XXVIJI; H.C. P1ummer, An Introductory Treatise on Dynamical Astronomy, 1918, Cambri.se;
0. Backlund, Zur Entwickelung der Storungstunction, Memoirs of the Academy of Science (Memuary Akademii Nauk) VII serie, t. 32, 1884.

A detailed bibliography is given in: H.v. Zełpel, Entwicklung der Storungsfunktion, Encyk.lopedie der Mathem. Wessenschaften Bd. VI, 2, 1912.

$$
د_{1} \quad\left|r: r:-n r r^{\prime} \cos (1)-\left|r^{\prime}\right|^{\prime}\right.
$$

We finally obtain

$$
\begin{aligned}
& د^{-1}=1_{0}^{-1}-r^{\prime} د_{0}{ }^{2} \therefore \text { an } \mathbb{U}^{\prime} \sin \|^{\prime \prime}+
\end{aligned}
$$

$$
\begin{align*}
& \text {. . . . . . . . . . . . . . . . } \tag{i}
\end{align*}
$$

These terms are sufficient for all of the large planets.
Let us denote by $H$ and $F^{\prime}$ the perihelion distances of the planets P and $\mathrm{P}^{\prime}$ from Point N at which their orbits intersect. In this case
where $v$ and $v^{\prime}$ are true anomalies.
Formulae (5) defines the expansion of the principal part of the perturbation function in terms of the mutial slopes of the orbits. We now consider the second parts of the functions (1). Since

$$
x x^{\prime}: y^{\prime} \quad \because y^{\prime}=r r^{\prime} \text { cos } H .
$$

the calcula ion of the second part is then reduced to the computation of expression

$$
R \quad \begin{gathered}
r \cos /! \\
r_{3}^{\prime},
\end{gathered} \quad R_{1}^{\prime} \underset{r}{r \cos / l} \text {. }
$$

Using equation (4), we obtain

In order to obtain the expansione of the perturbation function, given by equations (5) and (6), in a final form, it is necessary to express the
coordinates of the planets, $r, v, r^{\prime}$ and $v^{\prime}$, in terms of the orbital elements. For this purpose, it is recessary to again reconsider the particular case in which the eccentricities of the orbits are equal to zero, i.e. when the motion of planets $P$ and $P^{\prime}$ proceeds in a circle.
87. The Case of Circular Orbits:

If the eccentricities of the planets under consideration are equal to zero, then

$$
r=a, \quad w=i, \quad r^{\prime} \quad u^{\prime}, \quad w^{\prime}=\lambda^{\prime},
$$

where we denote by $\lambda$ anc $\lambda^{\prime}$. the average longitude in the crbit. Putting

$$
L=-i \quad r, \quad l \cdot, \quad,-Z^{\prime}, 1
$$

we obtain, instead of equation (5),

$$
\begin{equation*}
i \quad \Delta^{-1}=I-\| \cdot \mid I I I-N: . . . \tag{i}
\end{equation*}
$$

where

$$
\begin{aligned}
& I=\Delta_{0}^{-1} \quad \mid a^{-}-\vdots a^{\prime 2}-2 a a^{\prime} \cos \left(1 I^{\prime} \quad I .1 \mid\right.
\end{aligned}
$$

$$
\begin{aligned}
& \text {. . . . . . . . . . . . . }
\end{aligned}
$$

In this way, the problen of further expansion fis reduced to the expansion of a trigonometric series of the following type
where $n=1,3,5,7, \ldots$ and $S=L^{\prime}-L$. Moreover,

$$
\begin{array}{ll}
2 & \prime \\
\|^{\prime}
\end{array}
$$

Choosing our notations such that $a<a^{\prime}$, we can consider that and write

$$
\begin{equation*}
\left(1-2 x \cos S+z^{\prime \prime}\right)^{n} \cdot \frac{1}{2} b_{n}^{\prime \prime \prime} \cos i s \tag{1}
\end{equation*}
$$

since the function standing on the left-hand side can evidently be expanded in a Fourier series. The coefficjents $f_{n}^{(f)}$ are known as the Laplace coefficfents.

Furthermore, putting

$$
\begin{equation*}
c_{n}^{(1)}-a^{n-1} b_{n}^{(1)} . \tag{110}
\end{equation*}
$$

we obtain

We must note that

$$
b_{n}^{\prime} " \cdot b_{n}^{(\prime \prime \prime} \cdot \quad c_{n}^{\prime \prime} c_{n}^{(1)} .
$$

Substituting serf.e (1.1) in equation (7), we have to maltiply each of. these series (for $n=3,5,7, \ldots$ ) by one of the following expressions

$$
\begin{aligned}
& \left.2 \sin L \sin I^{\prime}=\cos 1 U^{\prime}-\quad \text {. }\right) \quad \cos \left(L^{\prime} \mid \quad\right. \text { I). }
\end{aligned}
$$

$$
\begin{aligned}
& +3 \cos \left(3 L+L^{\prime}\right)-3 \cos \left(3 L--L^{\prime}\right)+ \\
& +3 \cos (3 L+L) \quad 3 \cos (3 J . \quad 1)+ \\
& \left.\because \operatorname{con} 3 L^{\prime}-3 L\right)-\cos (3 L \mid 3 L)
\end{aligned}
$$

Th other words, we have to calculate a product of the type

$$
\begin{aligned}
& +{ }_{2}^{\prime} \searrow c_{n}^{\prime \prime \prime} \cos \left|\left(a^{\prime}-l\right)-v\right| .
\end{aligned}
$$

The two sums on the right-hand side of this equation are equal, and they are found to be so when $:$ fn one of them is replaced by - $i$.

Therefore

$$
\cos : \sum_{r}^{\prime(1)} \cos i\left(L^{\prime}-L\right) \cdots c_{n}^{\prime \prime} \cos \left|t\left(L^{\prime}-I .\right)+\eta\right|
$$

Using this rule, we can easily find
or, replacing $i+1$ in the first $\sin$ by $i$,

We calculate IIT, [V, ... by the sme nethod exactly.
Substituring the resulting expressions into equation (7), and then coliecting together the similar terms, we finally obtain

$$
\begin{align*}
& \left.+7^{i} \text { Da' } 1\right), \cos \left[(i+i) I^{\prime}-(i \quad \text { (i) } l \mid\right. \tag{12}
\end{align*}
$$

where the coefficients of these series are given by

$$
\begin{aligned}
& \text { | } \left.16 c_{4}^{\prime \prime}{ }^{3 \prime} \mid 18 c_{0}^{(1)}\right) \text { —... }
\end{aligned}
$$

$$
\begin{aligned}
& 350^{6}\left(r_{4}^{14}+f \cdot r_{4}^{(1)}, 4 c_{4}^{(1+1)}+c_{4}^{(1+3)}\right)+\text {. }
\end{aligned}
$$

$$
\begin{aligned}
& 35 \\
& a^{\prime} E_{1}=-\frac{3}{12} a_{y}^{(a)} .
\end{aligned}
$$

Formula (12) defines the expansion of the principal part of each of the perturbation functions (J) For the case of circular orbits. In order to actually perform this expansion, we only need to be able to calculate the coefficients $\mathrm{C}_{\mathrm{n}}^{(\mathrm{i})}$, or equivalently the Laplace coefficients for the values of $\dot{\alpha}=a / a^{\prime}$ under consideration. The way to do this calculation will be shown in one of the coming sections.

We now consider the secund part, of $R_{1}$ and $R_{2}$, of the perturbation functions, namely

$$
k^{\prime}=k^{\prime} m^{\prime}\left(د^{1} \quad R_{1}\right) \quad k^{\prime} \quad \& m^{\prime}\left(\Delta^{1}-\mu_{1}\right)
$$

Consulting formulae (6), we write

Comparing these expressions with the expansion coefficients of series (12) obtained here, we see that the influence of the second term of the perturbatic 1 function $R$ will be completely taken into account if we replace

Similarly, in order to obtain $R^{\prime}$, it is sufficient to replace in expansion (12)

$$
\begin{aligned}
& \text { di } \quad \text { リp: } \because A_{1}-3(1 \cdots:) \\
& \therefore i \quad \text { al, } 111 \cdots \\
& a \% \quad . \quad a^{\prime} \beta_{n} .
\end{aligned}
$$

In this way, the problem of the expansion of the perturbation function for the case of circular orbits is perfectly solved.

## Annotation:

It is important to note that the second part of the perturbation functinn consists entir€ 1 y of periodic terms. This can easily be seen from expressions (13). "ihe perturbation function does not finclude other secular terms except those which can be obtained from the sum of the expansion (12) for $i=0$.
39. Expansion of the perturbation fonction in powers of the eccentricities. Newcu ${ }^{\text {T }}$ 's method

We have seen in $S \in c .86$ that the gerturbation function $R$ is a function cf $r, r^{\prime}, W^{\prime}=\Pi+v$ and $W^{\prime}=\Gamma '^{\prime}+v^{\prime}$, and it is thus possible to write

$$
\Leftrightarrow \quad f\left(r, r, u^{\prime}, i r\right)
$$

In the previous section, we have put $e=0$ and $e^{\prime}=0$. Consequenily, $r, r^{\prime}, W$ and $W^{\prime}$ have keen respectively transformed into a, $a^{\prime}$ and

$$
1 \quad 11 \quad .11 \quad 1 . \quad 11 .: i
$$

where $M$ and $M^{\prime}$ are the average anomalies of the planets under consideration. In the corresponding case, the functions $f\left(a, a^{\prime}, l, L^{\prime}\right)$ have been given
by formulae (12) and (13) in the form of unfolded expressions.
We shall now consider that the eccentricities $e$ and $e^{\prime}$ have small values. We shall expand the expression $f\left(r, r^{\prime}, W, W^{\prime}\right)$ in powers of $e$ and c'. In order to simplify, the application of Taylor's formula, it is better to consider the perturbation function $R$ a $s$ function of $\log r$ and $\log r^{\prime}$ and not of $r$ and $r^{\prime}$. Actually, the transition from $\log a$ Lo 1 g $r$ will te performed by adding an increment, whereas the transition from a to $r$ is done by multiplying a by some correction factor.

Hence, we assume on one hand
and on the other hand
where we denote by

$$
j \because r . \therefore!\quad \quad \therefore \quad r-r
$$

the equation of the centre for the planets under consideration.
We have proved in Sec. 82 that $\rho^{\circ}$ and $f$ can be expanded in powers of $e$, in series having the form


Our task consists in expanding the expression
in powers of $e$ and $e^{\prime}$. For this purpose, we apply Taylor's formula,
which can be represented for the sase of expanding a function of several variables in the following symbolic form

To make this fnrmula more compact, we introduce the following notations


In these notations, we shall have the following operator equation

Since the operator $\exp \left(f^{\prime} D+\rho^{\prime} I^{\prime}+f_{1}+f^{\prime} D_{1}^{\prime}\right)$ is a product of the following two operators
then the operator equation under consideration is the product of each of these operators and the function

$$
\begin{equation*}
l\left(\underline{l}, 6,1 \leq a^{\prime}, l, A^{\prime}\right) \tag{16}
\end{equation*}
$$

Putting
we can write the function (16), reprosented by formulae (12) and (1:), in the following way

$$
\underline{\prime \prime}
$$

We multiply the arbitrary term

$$
R^{\prime \prime .}\left(1(1, s)+A^{\prime}\right.
$$

$1 .-$
of this series by thefirst of the two operators given by equation (15). Since
then

Let us now consider equations (14). Putting
whre $k_{0}, k_{1}, \ldots$ are functions of $D, s$ and $K$, and

$$
\because \quad \text { expll-1 }: 1
$$

we write equation (14) as

$$
\begin{aligned}
& +{ }_{8}^{1} e^{5}\left(-1+i a^{-1}+\int_{3}^{13} \mu^{s}-3_{3}^{1.3}\right)+
\end{aligned}
$$

Substituting these expressions into equation (18), - attein

$$
\begin{aligned}
& R_{0}=1 \\
& k_{1}=\| I_{1}^{\prime} \mu-1-1 I^{\prime}, \mu^{-1}
\end{aligned}
$$

where $\Pi_{m}^{m}=\Pi_{m}^{n}(D, s)$ is an $n$-order polynomial in $S$ and $D$. These symbolic polynomiais will be called operaturs. It is easy to see that

$$
\left\|\|_{m}^{n}(1),-s\right) \ldots \|_{m}^{n}(D, s)
$$

Thus, the term of expansion (17) that contains the factor $e^{n}$ will be of the form
where $m=n, n-2, n-4, \ldots,-n$, and

$$
f_{m}^{n}\left(x, v^{\prime}\right)=\|_{m}^{n} H\left(s, s^{\prime}\right)
$$

Multiplying expression (20) by thesecond of the operations (15), and denoting the corresponds-s polynomials by $\prod_{\mathrm{cm}^{\prime}}^{0 n}$, we obtain for the expansinn term of function (20) having factor $e^{\prime * \prime}$ in the following form

The result of acting first by $\prod_{m}^{m}$ and then by $\prod_{0_{m}}^{a_{n}}$ on the coefficient ( $H$ ( $s, S$ ) may be represented as the result of action of the composite operator

In short, assuming that
and separiting the real part in expressinn (21), we finally obtain the expansion of the perturbation function in the following form
where the indices $n$ and $n^{\prime}$ vary fron 0 to $+\infty$, while the indices $S, S^{\prime}$, $m$ and $m^{\prime}$ vary from $-\infty$ to $+\infty$. The coefficients $P$ depend on the semimajor axes a and $a^{\text {: }}$ and on the mutual slope of the orbit J. Equations (12) and (1.3) indicate that the sum $S+S$ is always an even integer. Moreover, taking equation (14) into account, it is easy to see that each of the differences $n-m$ and $n^{\prime}-m^{\prime}$ is also equal to a non-negative even integer.

Thus, the expansion of the perturbation function in powers of the eccentricities is finally reduced to the calculation of the operators. For the initial values of the indices $n, m, n^{\prime}$ and $m$ the calculation of the operators is simple. With increasing values of these indices, complexity of the calculation rapidly increases. In order to calculate the operators in these case it is advisable to use the recurrence relations existing between them ${ }^{(1)}$.

We have seen in the previous section that in the case
when $e=e^{\prime}=0$, the secular terms in the $p \cdot a n s i o n$ of the perturbation functions are entirely obtained from the expansion of the principal term $\Delta^{\mathbf{- 1}}$. The application of the operators can only give periodic terns, as we have already seen in our consfderation of the above-ment ioned method of expansion in powers of eccentricities, This enables us to formulate the following thenrem.
(1) The methods which have been suggested fir the calculation of the operators are given in detafi in the mnaugraph:
B.A. Orlov, Fxpansion of the perturbation functions by Newcomb's method, Transactions ot the istronomical Observatory of the Universfty of Leningrad (Razlozenie perturbacionnof Funkcii po metocud N'jocoma, 'Trudy Astronomiceskoj otsefvatorif Leningrads'zogo universiteta)
6, 1936, 82 - 125.

## Theorem:

The seculax terms in the expansion of the perturbation functions are obtained only in the expansion of the principal texm.

In conclusion, we note that in order to carry out the expansion given by equation (22), it is necessary to not only know the coefficients (10) but also their derivatives with raspect to $\log \alpha$. Indeed, since $\alpha=a / a^{\prime}$ then

Hence, we :an consider $D$ as the differentiation symbol with respect to $\operatorname{lo}_{\mathrm{g}} \alpha$ and write

$$
J_{i_{n}^{\prime \prime}} \quad d^{\prime} i_{4}^{\prime \prime}
$$

$$
\left(\begin{array}{ll}
i & -1, \therefore \\
, & .
\end{array}\right.
$$

## Annotation

For an arbitrary homogeneous function $\Phi\left(a, a^{\prime}\right)$, the order of which is -1, Euler's theorem gives
or

Thus, for any such function, the following symbolic equation lolds

$$
1): 1
$$

From this equalifor, it follows that

$$
\left.11_{1}^{:} \quad \|_{n}^{n}(-1)-1, s^{\prime}\right)
$$

Hence, the calculation of all operators

$$
\left\|\|_{m \times n}^{n} \cdot(1), s, v\right) \quad\left\|\left\|_{m}^{n}(l, s)\right\|_{m, 1}^{n}(1)-1, i!\right.
$$

is reduced to thecalculation of the simple operator $\prod_{\dot{m}}^{4}$
80. The Final Form of the Fxpansion of the Percurbation Function

In the previous seclion, we have become aquainted with the methods for obtaining an arbitrary number of terms for the expansion of the perturbation functions in powers of the excentricities and mutc. 1 slope of the orbits. We have shown that this axpansion has the following form
where $h, h^{\prime}$ and $f$ run over the values $0,1,2, \ldots$ while indices $s, s$, $m$ and $m^{\prime}$ take the values $0, \pm 1, \pm 2, \ldots$ The $\operatorname{sum} m+m^{\prime}$ should always be equal to an even integer. The differences $\mathrm{t}-/ \mathrm{m} /, \mathrm{h}^{\prime}-/ \mathrm{m}^{\prime} /$ and 2f-/sts'/ may be only equal to the even integers $0,2,4, \ldots$ The coefficients $K$ lepend on the indices $h, h^{\prime}, s, s^{\prime}, m$ and $m^{\prime}$ and are functions of the semimairo axes of the orbits $a$ and $a^{\prime}$. The power of each term of the expansion (23), i.e. the sum $h+h^{\prime}-2 f$, is equal to or exceeds by an even integer the following quantity

$$
s: s^{\prime} \mid-: m: m^{\prime}
$$

Expansion (23) represents the perturbation function in the form of a trigonometric series of four arguments. It is the simplest of all representations of the perturbation function as an explicit function of time.

In order to integrate in a simpler way the Lagrange euqations which define the osculating elements (sec. 13), it is recomended to slightly modify expansion (23). Noting that

$$
M=1 .-11, \quad A \quad 1-11
$$

then this expansion may be given the following form

Indices $p, p^{\prime}, q$ and $q^{\prime}$ run all the integral values from $-\infty+\infty$, where the differences

$$
h-q 1 . \quad n^{\prime} \rightarrow q^{\prime} . \quad \because \prime-j \mid-n^{\prime}+q: q
$$

are non-negative even integers. We thus write

$$
h \cdot a: \therefore / \cdot 0 \cdot p \cdot q: 4: q+4
$$

From this inequality, it follows thご

$$
i ; n \quad 2 y \therefore i+p .
$$

It is easy to see that thedffference between the left and right-hand sides of this inequality is always equal to an even integer. Hence, the power ofeach term of the expansion (24) is either equal to / p $+p^{\prime} /$ or exceeds this quantity by an even integer.

We cannot direct:ly apply the expansions (23) or (24) to the fategration of lagrange equations. The reason is that these expressions involve the elements $\Pi, \Pi$ and $J^{\prime}$ whick define the mutual orientation of the orbits, while the differential equations involve the elements $i$, $\mathcal{R}^{\prime \prime}, \ldots$, $i^{\prime}, \ldots$. In order to obtain the perturbation function $R$ in the form of an explicit function of these orbjtal elements, we use the following relations

$$
\begin{aligned}
& \text { I. }-11 \cdot \|=n t+=-1 \quad N, \quad I^{\prime} \quad n^{\prime} t+\varepsilon^{\prime}-M^{\prime}-N^{\prime} \text {. }
\end{aligned}
$$

Therefore, the argument of the expansion (24) may be replaced by

$$
\begin{aligned}
& p(n \prime-i) \therefore p^{\prime}\left(n^{\prime} t+\because^{\prime}\right) \cdot\left(-q-, q^{\prime} \therefore^{\prime}-(p \cdot q) \underline{Q}\right. \\
& -\left(p^{\prime}+q^{\prime}\right) \geq-(p-q q) V-\left(p^{\prime}+q^{\prime}\right) N^{\prime} .
\end{aligned}
$$

or

$$
\begin{equation*}
D+\frac{1}{2} 3\left(N+N^{\prime}\right)+\frac{1}{2} ;\left(N-N^{\prime}\right) \tag{25}
\end{equation*}
$$

where

$$
\begin{gathered}
D=p(n t+s)+p^{\prime}\left(n^{\prime} t-t\right)+q r+q^{\prime} q^{\prime}-(p+q) Q-\left(p^{\prime}+q^{\prime}\right) q^{\prime} \\
\beta=-p-q-p^{\prime}-q^{\prime} . \quad ;=-p-q+p^{\prime}+q^{\prime} .
\end{gathered}
$$

When we are unfolding the cosine functions having arguments of the type (25), we find that the factors of the expressions, which do not depend on $\Pi, \Pi^{\prime}$ and $J$, have one of the following forms
where $\sigma^{\circ}=\sin \frac{1}{2} \mathrm{~J}$. Using formulae (16), Sec. 68 , to express these quantities in terms of the elements and using Euler's formulae, we obtain from thesc equations

$$
\begin{aligned}
& \cdot 11_{1}^{2}, 111_{2}^{\prime} 1.11,1!!\quad \because 11-1
\end{aligned}
$$

Raising these equations to the power $\beta$ and making a transtion from the exponentional to trigonometric functions, we obtain the following expressions for the quantitias

La terms of the orbital elements. On the other hand, by applying the same formulse (16), Sec. 68, wi obtain after ralsing them to the second power and adding in pairs

Rlasing these equations respectively to the powers $f-1 / 2$ and $-\beta$ and multiplying them term by term by the expressions obtained for the quantities (28), we obtain the final expressions for the quantities (26). In this way, the perturbation function $R$ will be given the following form

$$
R^{\prime}=\sum^{\prime} A e^{\prime \prime} c^{\prime \prime}\left(\sin \begin{array}{l}
i  \tag{29}\\
2
\end{array}\right)^{k}\left(\sin \frac{1}{2}\right)^{k^{\prime}} \cos D_{0}
$$

where

The coefficient $A$ depends only on and $a^{\prime}$. It is easily scen that the findices $p, p^{\prime}, \ldots, s^{\prime}$ involved in the argument $D_{o}$ must always satisfy the relation

$$
\rho+\rho^{\prime}\left|4:-q^{\prime}\right| \leqslant \vdots \therefore 0 .
$$

Indeed, the perturbation function $R$ does not ev. .ently repend on the initial point for calculating the longitude. Yet, if this initial point is displaced by an angle $\mathcal{\Lambda}$, then the argument $D_{0}$ will be changed into
 depend on $\Lambda$, then this quantity should be set equal to zero. This means that the trigonometric series (29) is developed not by six indices, but only by five indtces.

The actual working up of expansion (29) is an extremely tedians job and it has never been actually dpplied. The theory of motion of
large planets, developed by Laverrier, which describes the most commonly applied method of variation of the elerents in the anayytical calculation of perturbations, is bascd on expansions occupying an incermediate place between formulae (24) and (29).

Laverrier imtroduced, istead of the longitudes $L$ and $A^{\prime}$ measured from the point of intersection of the orbits, the longitudes in the orbjes (using his notations):

$$
l n t-\mathrm{l}-\mathrm{l}: \div \quad l=\pi l+z \quad l .
$$

Assuming
yields

$$
\text { ) } 1
$$

$$
\cdots n
$$

$$
\text { L. }=\mathrm{i} .-:^{\prime} . \quad L^{\prime}+\gamma^{\prime}-\because^{\prime}
$$

$$
1.1
$$

Similarly, notir that the longitudes of the perihelions are equal to

$$
\because \cdots \quad 11 .^{\square} \quad ; \quad \div 11^{\prime}
$$

and assuming

$$
\begin{equation*}
\omega . \therefore \mid:^{\prime}-=. .11+\cdots \tag{1:2}
\end{equation*}
$$

we obtain

$$
\begin{equation*}
\|=\| \quad \because \quad \quad \|^{\prime} \cdot \pi^{\prime}-\because^{\prime} \tag{i3}
\end{equation*}
$$

Substituting expressions (31) and (33) irto the expansion (24), we finally have the following expansion for the perturbation function
'This expansion constitutes the basts of all Laverrter's work.
It is easily seen that the $j, j^{\prime}, k, k^{\prime}$ ans 2 g are related by
the fcllowing relation

$$
j-+j^{\prime}+k^{\prime}+k^{\prime}-2 g \cdots 0
$$

This relation is a consequence of the independence of the perturbation function $R$ on the initial point for calculating the longitudes.

The quantitios $\lambda$ and 4 , defined by equaiions (30) and (32), differ from $\ell$ and $\pi$ by the infinitesimal quantłty $\tau^{*}-\tau$. It easy to show (though not undertaken here) that

$$
\operatorname{tg} \frac{a^{\prime}-}{\underline{2}}=\frac{\operatorname{tg} \frac{i}{2} \operatorname{tg} \frac{i^{\prime}}{2} \sin \left(x^{\prime}-\infty\right)}{1-\operatorname{tg} \frac{i}{2} \operatorname{tg} \frac{i}{2} \cos \left(x^{\prime}-s^{\prime}\right)}
$$

Hence,

$$
\begin{equation*}
\frac{1}{2}\left(\tau^{\prime}-=\right)=\operatorname{tg} \frac{i}{2} \operatorname{tg} \frac{i^{\prime}}{2} \sin \left(x^{\prime}-x\right)-\frac{1}{2} \operatorname{tg}^{2} \frac{i}{2} \operatorname{tg} \frac{t^{\prime}}{2} \sin 2\left(x^{\prime}-x\right)+\ldots \tag{35}
\end{equation*}
$$

In order to make use of the expansion (34), Leverrier was obliged to introduce sone spccial modification fnto the Lagrange equations (Sec. 97).
90. The Initiai Tenns of the Expansion of the Perturbation Function

In order to carry out the expansion (24), it is necessary to express the coefficients $K$ in terms of the finctions $c_{k}^{(j)}$ of the ratio $\alpha=a / a^{\prime}$ that $h$.ve been introduced in Sec. 37 , and then to compute these functions for a given value of $\alpha$. We shall considur the latter problen in the next chapter. Here we shall consider the calculation of the coefficients $K$ in an explicit form.

We shall confine: ourselves to second-order terms with respect so e, $e^{\prime}$ and $\sigma$. The operations deseribed in the previous section will then apply in a quite simple manner. It l.s easlly seen that these operations will lead to the following expansion for the principal part of the perturtation function

$$
\begin{aligned}
& +!c \geqslant(-2 i-D) c_{1}^{\prime} \mathrm{cosi}(V+M,+ \\
& \dot{-1} e^{e^{\prime}}{ }^{7}(2 i+1 ; D) c_{1}^{\prime \prime \prime} \operatorname{coi}\left(V-+M^{\prime}\right)+ \\
& \left.+\frac{1}{8} e^{2}\left(1 i^{\prime}-\Gamma i+,: \therefore-3\right) D+D^{2}\right\}\left(1^{\prime \prime \prime} \cos (V \mid 2 n) \gamma\right.
\end{aligned}
$$

$$
\begin{aligned}
& \text { - } 1
\end{aligned}
$$

where, the following notation has been adopted

$$
\left.V-l\left(L^{\prime}-i\right), \quad l\right) \cdot \dot{j}(\lg 2)^{\circ}
$$

and the sumation is to be carried over the vaiues $i=0, \pm 1, \pm 2 \ldots$.
? … :...-f tern of the perturbation function can be taken into accounc ; meatis of the rethods given at the end of sec. 87. In Eact, the series-expansion of this term can be immediately written down using formula (36). Indeed, assuming that the eccentricities e and $e^{\prime}$ are equal to zero, the second part of the perturbation funstion

$$
R_{1}-\underset{r^{\prime}:}{r \cos H}
$$

will be defined, according to the Eirst of formulae (13), by the following relation

$$
a^{\prime} R_{1}-z\left(1 \quad \because^{\prime \prime}\right) \cos \left(L^{\prime}-1 .\right) \mid a z^{\prime \prime} \cos \left(l^{\prime} ; L\right) .
$$

Comoaring this expression with that given by equation (12), we see tiat. the former can be corsidered as a particular case or the Jatter, if re put in that latter expression

$$
c_{1}^{\prime \prime} \therefore c_{1}^{\prime-1 \prime}=2 . \quad \dot{G}_{1}^{\prime \prime}=12 .
$$

and equate all the other quantities $C_{k}^{(i)}$ to zero. Applying in a similar manner formula (36), to this particular case, we obtain in terms of the second power of $e, e^{\prime}$ and $\sigma$ :

$$
\begin{aligned}
& \text { "r'cus ( } \because!\text { ! } 1 .-11^{\prime} \text { ) ! }
\end{aligned}
$$

$$
\begin{aligned}
& \frac{1}{n} e^{\prime} \cdot \cos (1 .+1-211)+\frac{17}{-2} \cdot \cos (1 / 2-1 . \quad \text { - } 11):
\end{aligned}
$$

-ary, the secori $a$ formbe (13), thich yields
indicates that the second part of the perturbation function $r$ ' can be obtained trom expression (12) by putting

Again using formula (24., we obtain

$$
\begin{aligned}
& +-=2 \cos \left(L^{\prime}+L\right)+1 .
\end{aligned}
$$

Whis is the exparbion os the perturbutlon function in the case when terms of the third-order with respeat to the eccentricitics and slopes can be neglected. This expansion has alrcady been pareviousiy obtained by Lagrange and Laplace to uithin the same accuracy.

In order to chtain the perturbation of the cordinates of large planets with an accuracy corresponding to that of the recent observations, it is nceessary to carry out the expanston of the perturbation function w to terms of the 7-th orfer inclusively. Tre foscitility of doing such expansions was shown by Borckhardt. Mis calleulations were corrected and feveloped by Binet and de Porteocoulant, who gave expatisions tiat ancluded a considerable part with sixth crder teras. The complete exparision of ibrenturbation fanction up to s-order tams inclasive!y, was farst. obtained in Plerce's work in 1849. Finall: in 1855, Leverrier ${ }^{(1)}$ pablished the expansion of tie perturtation function up to the 7-order terms inclusivaly. The aceurate expressicas which he obtained for all the terns up to thels limit lid not iorioe their
(1) Amales de I'Observatoire le Paris, t.1, 1\&55.
values, even at the present time. Boquet ${ }^{(1)}$ included all the $\mathcal{G}-\mathrm{th}^{\prime}$ order terms into the expansion preser ing iaverrier's method and notations.

The most convenient way to obtain the expansion of the perturbation function ia the form given by equation (23) is by Newcomt's method ${ }^{(2)}$, based on the appifation of cperators as defined in Sec. 88.

## Annotation

Using the formulae obtained in the previous chapter, it is easy to obtain the expansion of the second part of the perturbation in a general fosm with coefficients expressed in terms of Bessel functions.
91. Numerical Method for the Expansion of the Perturbation Function

In the previous scetione, we studied the methods of the accurate calculation of the perturbation functions in the fcrm of a series. Each' term of such series is an explicit function of the abital dements and average anomalies $V$ and $M^{\prime}$ of the planets undes consideration. This form of expansions gives ile owst general solution to the probler. It allows us to obtain the pertirbstions as explicit. functions of the orbital elements. The methods of obtaining such expansions, where all the elements enter as letters (except the semimijo: exes, which are given nunerical valites in order to be able to someute the Leplace cerfficients). are known as the analytical methods of expansion of the perturtation function.
(1) F. Boquet, Developpmert de Lifonction porturbatrice, Annales de l'Observatoire de Paris, t. 1E, 1885.
(2) S. Nevcomb, A Development of the perturbation function pic., Astronomical Papers, Vol. V, 1895.

The analytical methods of expansion give the perturbation function, in the form of a series in powers of eccentricities. They can be applied practically for only small values of eccentricities (not exceeding 0.15 or 0.20 ). If this condition does not $9 p l y$, then the expansion can in practice te only obtained by means of numerical methods although it may converge quite rapidly (cf. the annotation to Sec. 84). In this case, one has to apply the numerical methods ofexpansion in winch the elements enter from the very beginning $u$ ing their numerical values.

If the orbital elements of the planets under consider: ion are given using their numerical values, the perturbation function may then be represented by the series
developed by multiples of the average anomalies $M$ and $M$. The enefficients $A$ and $B$ of this series are expressed by the following we11-known formulae

These coefficients can be obtained by means of the apficxinate formulae which replace each integral by a sum of the values of the integrand for virtuous values of the argument. These formulate ray be put together in a single formula in the following way.
where $R_{k, k^{\prime}}$ is the value of thefunction $R$ for $M=k \frac{2 \pi}{m}$ and $M^{\prime}=k^{\prime} \frac{2 \pi}{m^{\prime}}$ Praluatirg the Eunction $R$ for a sufficienliy iarge number of specific values of the average anomaly, we can compute the coefficients of the expansion (37) to an arbitrarily high accuracy.

Naturally, the expansion (37) with numerical coefficients cannot be used in calculating the derivatives of the perturbation function with respect to the elements. Hence, it is not possitle to apply this expansion to compute the perturbation of the elements by means of Lagrange's formulae (Sec. 15). However, this kind of expansion is quite useful for the purpose ofthe direct computation of the perturbations of the coordinates (Chapter XVI). In this case, it is sufficient to have the exprasiors of $\Delta^{-1}$ and $\Delta^{-3}$ if only first order perturbations are required. For higher-order perturbations, 1: is necessary to also have the expansions of $\Delta^{-5}, \Delta^{-7}, \ldots$.

Hansen was the first to pub:ish an application of the numeri.cal. method of expansion of the perturbatio: function, which he had been appiying to the study of the mutual pert:urbations of Jupiter and Saturn (1831). He used as an argument the dffference M-M' between the average anomalies and the average anomaly of Saturn M'. In order to obtajn the expansion corfficients by means of the harmonic analysis for Mulae, he computed $\Delta^{-1}, \Delta^{-3}, \ldots$ for all the combinations oi the following values of the argument

$$
\begin{aligned}
H-H^{\prime} & =111 j^{\prime} \cdot k, \quad k_{1} \quad 0,1, \ldots .+31 \\
. H^{\prime} & =2230^{\prime}, k^{\prime}, \quad 4^{\prime} \cdot 1, \ldots
\end{aligned}
$$

By this mathod he had to compute $32 \times 16=512$ particular values of the abors mentioned functions.

## 92. Hansen's Mechod

The numerical method mentioned in the previous section enables us to obtain the expansion of the perturbation function with an arbitrarily high accuracy by means of simple and easily mecianized computations. The only inconvenience in these computations is in their extensiveness.

In applying numerical methods, we do not make use of the properties of the function to be expanded into a series. It is quite natural that the following question crops up: can we reduce the calculation wo $k$ by the eppropriate use of the properties which we know on the analytical structure of the expanded function?

Cauchy was the first to apply a semianalytical method for the expansion of the perturbation function (1844). He carried out the expansion analytically by one argument and numerically by the other. This idea was further developed by Hansen (1857) who gave an expansion method wich had been widely applied. Hill, in particular, applied this method to the construction of the theory of motion of Jupi er and Saturi. The problem consists in expanding the quantities $\Delta^{-3}, \Delta^{-3}, \Delta^{-5}, \ldots$, wher:

$$
\begin{aligned}
& \therefore=:=r: r^{2} \text { 2rricull. }
\end{aligned}
$$

in a double tafonometric sfries. Considering again $a<a^{\prime}$, and putting $\mathcal{}\left(a / a^{\prime}\right.$, we write this equation as

$$
\binom{د}{a^{\prime}}^{\prime}=\binom{r^{\prime}}{a^{\prime}}^{:} \vdots\left(\begin{array}{llll}
1 \\
a & )^{2} & \ldots & 1 \\
a & a^{\prime} & a \cos 11 .
\end{array}\right.
$$

On the other hand, the function cos $H$ equals in an unfolded form, to

|  |  |
| :---: | :---: |
| C | - cos 11 sm $11 ;$ - $\sin \\|\operatorname{los}\\|^{\prime} \cos .1$ |
| $C_{1}=$ |  |
| Ci) | vall still'a cos Il cos II' cos. |

Introducing the auxiliary quantities $k, K, k_{1}$ and $k_{1}$ by means of the following relations

we easilv fand that

Substituting this expression firto equation (38), and using the following well-known formulae

$$
\begin{aligned}
& \text { r. a(l cast) r-ail erosti). }
\end{aligned}
$$

we obtain the following expression for 're square of the separation distance of two planets in terms of their excentric anom

$$
\left(\frac{1}{a^{\prime}}\right)^{\prime}: 11-f \cos (t-\mu)+\frac{1}{2}, \cos : t .
$$

where

$$
\begin{aligned}
& \text { 2.-2:1" }
\end{aligned}
$$

where the constant coefficients have the following values

$$
\begin{aligned}
& \text { (1) 2 (etcoskーか) }
\end{aligned}
$$

$$
\begin{aligned}
& \text { (i, = = ? R Anstico: } \\
& \text { ( } i_{1} \text { : } 2 h_{1}>11 h_{1} \text { con. } \\
& \text { ij } \quad \therefore \text { cos. } h_{1} \text { cur, con. }
\end{aligned}
$$

$$
\begin{aligned}
& \text { il 2lはいい } \\
& 11 \text { - 2i~IIncus : }
\end{aligned}
$$

he note that if the quantities $D, f$ and $F$ are lnown，it is then easy to compu＇e the functions $E$ and $M^{\prime}$ ．When $M^{\prime}$ is increased by 2 ，the angle $F$ also increases by $2 T$ as it is easily seen from the above equations．

The detailed examination of formulae（42）indicates that when the excentricities $e$ and $e^{\prime}$ are small，the differences $E-E \cdot$ and consequently F－n＇remain within sufficiently close limits whatever the change in $\mathrm{M}^{\prime}$ ．

In all practical cases，the last term of expression（41）is very small as compared to the sum of the first two terms．This situation enables us to write，on the basis of the binomial formuia，the following repidly converging expansion

$$
\begin{array}{ccccc}
1  \tag{111}\\
1 & \vdots & 1 & 1 & 1 \\
1 & \cdots & \cdots & 1 & 0
\end{array}
$$

where

$$
I_{0} \quad \mid 1 \quad / \text { comal: }-\left.i\right|^{\prime} \text {. }
$$

The present problem should be reduced to the expansion of the quantities $\Delta_{0}^{-1}, \Delta_{0}^{-3}, \ldots$ ．Each of these ruantities can be representca by a

## Fourier series

$$
د_{1} " \quad z_{n}^{\prime \prime \prime} \mid \because z_{n}^{\prime:} \cos (E \quad H), \because z_{0}^{\prime:} \cos 2(t-l) i \ldots \text { (ti) }
$$

The coefficients of this series can easily be expressed in terms of Laplace＇s coefficients．Jn fact，setting
() … 俥 (1-;
we obtain

$$
\begin{aligned}
& د_{0}^{-n}=\text { 虫 } \because \|+1,3-2 \prime \cos (\because-r) \mid:=
\end{aligned}
$$

conserfuently，

$$
\begin{equation*}
a_{n}^{\prime \prime \prime}-\frac{1}{2} \boldsymbol{y}^{\prime} \quad b_{n}^{\prime \prime \prime} \tag{1i}
\end{equation*}
$$

Replacing the argument of the series（45）by the difference

$$
t \quad r=t \quad \cdot r-\left(l \cdot-I^{\prime}\right)
$$

we will be able to transform this series in the following way
where

$$
\begin{align*}
& \therefore " \quad 3_{n}^{\prime \prime \prime} \cos 1(f-. \mid l)  \tag{14}\\
& \therefore ",
\end{align*}
$$

As we have already poir．亡ed out we can calculate the quantities $D, f$ and F for any values of $M^{\prime}$ ，and subsequently compute the coefficients and by means of formulae（46），（47）and（49）．Computiny each of these cuefficients for a series of equally－spaced values of $M^{\prime}$ ，and applyjng．
the conventional methods of harmonic analysis, we can expand each of the coefficients $\beta$ and $\gamma$ in a series of the type

Substituting such a series for each of the coefficients of expansions (48), and introducing these latter expansions into formula (44) and then unfolding the resulting, product of trigonometric functions, we finally obtain series of tide type
where $n=1,3,5, \ldots$, while the indices $i$ and $i$ are set equal to $i=0, \pm 1, \pm 2, \ldots, i^{n}=2,1,2, \ldots$.

The expansion (50) involves two variables, $E$ and $M^{\prime}$, and hence cannot be directly applied to the integration of the equations which define the perturbations. Hansen expressed $M^{\prime}$ in terms of the eccentric anomaly E, which le considered as an independent variable. Since,
then

$$
\begin{gathered}
11 \\
n!
\end{gathered} \quad . H_{\ldots}, .11 \quad n!\quad H_{1}{ }^{\prime}
$$

where $\mu=n^{\prime} / n$ and $C$ is some constant. Hence, using Kepler's equation, we obtain

$$
H^{\prime} \text { it ictintic } C
$$

Substituting this expression instead of $M^{\prime}$ in formula (50) and unfolding the functions

# viigginal page le OF POOR QUALITY 

involved in the trigonometric series by means of formulaf (15), Sec. 78, we obtain an expansion of the type

$$
\begin{equation*}
\left(\frac{u^{\prime}}{J}\right)^{n}-\quad|i, i, r|_{n} \cos (i t i-1 \mu t): \sum^{\prime}|i, i, s|_{n} \text { vn (it: itti). } \tag{i}
\end{equation*}
$$

In the theory described above on Jupiter and Saturn, Hill transformed expansion (50) finto an expansion of the type
which could easily becarried out by means of the methods indicated in Sec. 81.

## Annotation

In ordex to compute $\theta$ and $\Pi$ by means of formulae (16), it is recommended to introduce the auxiliary angle $\psi$ : defined by the relation

$$
\sin ; \quad \frac{1}{l}
$$

and the condition $0<4<90^{\circ}$

The equation

$$
1^{-\prime \prime} \quad \text { sur }
$$

has the following two roots

Taking the first of these roots, we will have the following equalities
I IA太, 将 I Jrus ir

## THE IAPLACE COEFFICIENTE

93. Calculation of the Laplace Coefficients by Means of Series

In order to end the question on the expansion of the pexturbation functions into series, it remains for us to consider the methods of calculating the quantities $C_{n}^{(i)}$. It follows from equation (10), Sec. 87, that the computation of these quantities is equivalent to the computation of the quantities $b_{b}^{(i)}$, defined by the following relation

$$
\begin{equation*}
\text { il } 2 \cdot \cos +2 \cdot{ }^{\prime \prime}=!_{2} b_{n}^{\prime \prime \prime}(1,4 \tag{i}
\end{equation*}
$$

and known as the Laplace coefficients. We shall prove that the Laplace coefficients can be computed by means of infinite series. We put

$$
\therefore \text { M1! } 55^{\prime}-11
$$

Then

Consequently, equation (1) may be replaced by

$$
\left.\begin{array}{ll}
1 & 2 z
\end{array}\right)^{\prime}\left(1 \ldots x 1 y^{n} \quad!\text { N } b_{n}^{\prime \prime \prime}\right.
$$

Sunce
then, we can easily obtain, by eouating the coefficients of $z^{i}$ on the right- and left-hand sides, the following formula

Taking $i=0$, we obtain

It is clear that these series converge for all positive values of which satisfy the condition

$$
<1
$$

However, the convergence of these series is very slow for all values of $\alpha$ even for the case in which $n=1$.

We make use of the conventional notation of the hypergenmetric function
and also introduce the symbol

We then write formula (3) in the following way
it is well known that the bypergeometric function, def-ned by series (5), satisfies the following relation

$$
H+, B, C, y)=\left(1 \quad \text {, }+\left(\lambda, C, B, C_{i}, \quad x\right)\right.
$$

Hence
or, in an unfolded form,
where

$$
\because \quad 1^{24}
$$

This se:ies converges wher $p<1$, j.e. when the vilue of $\alpha$ satisfies the rolluwing conc inn

$$
<\frac{1}{\sqrt{!}}=11 i 1
$$

The advantage of epplying series (7) rather than suries (3) is nore apparent for larger values of $i .$. Indeed, the ratios of the roreeponding coefficients in these two series are

This ratio tends to zero when i tends tc infirity.
If $a \geqslant 0.707 \ldots$, then series (7) diverges. However, even in tias case we can daply this series for compating $b_{r}^{(i)}$ if $i$ is suffieiently lacga. A.: a matter of fact, it is possible to show thet in this zase series (7) heromes an asymptotic sezies. The upplication of the divergeni series (7) for large values of $i$ will be ever: macre practizal itar the epplicaticin of the convergent series (3).

Series (7) mey be transformed into a more nratical and at the same han as a convergent series by ncans of an analyeical continuation. Tndeed, the branch of , he function (7) under consiteration has no singuiar points except $\alpha=1$ and $\alpha=\infty$ consequently the only singular points of the corresponding branch of the function
$1, \% 1 \quad \because \quad 1$
will be the points $p=-1$ and $p=\infty$. Hence, applying Tayior's
formula io the function $F(p)$,
where $F_{0}$ is real and posjtive, we obtain a series which has a circle of convergence with radius $1+p_{0}$.

We ehall now give the numerical coefficients of the series which may be weed for computing $b_{1}^{(10)}$ and $b_{1}^{(11)}$ in the cases, when $p_{0}=0$, $\frac{1}{2}$ and 1.

The first of these Laplace coefficients is equal to
where, for the function $F_{10}(p)$, we have the following expansion


For $b_{1}^{(11)}$, equation (7) yields
where one of the following expminsions may be used for the finction $F_{i i}(p):$

| p, - 1 | $p_{0} \quad 1$ | r 1 |
| :---: | :---: | :---: |
| $+\operatorname{lin} 0$ cound |  | + 1 M917: 227 |
| $0.11294 .6 .43 p$ | $-0.01907: m i p r l y)$ | - 00177ati20tp 11 |
| + $0.00180288 p^{\prime}$ | $+0.691+16880\left(\begin{array}{ll}D & 1 \\ 0\end{array}\right)$ |  |
| - 000026439 | $00001851(p-1)$ |  |
|  | : 0.0000 .311 (p.. 1. | OMNHOLTM, |
| -0.00001 in |  | Oncmo. $1.55(p-1):$ |
| 10.001010 | (f)0.00)2 $\left(\begin{array}{ll}p & 1 \\ 2\end{array}\right)$ |  |
|  |  |  |
|  |  |  |

The choice for the most convenient series may be obtained by consuiting the foilowing table

## 94. The recurrence relations between the Laplace coefficients

We consider equation (2), which may be used for the detcrination of the Laplace coefficients. This equation may be written as

$$
|1| x^{2}-a\left(\left.z!z^{-1} j\right|^{-n} \quad \vdots \quad \vdots n_{n}^{\prime \prime \prime}=\Xi^{\prime}\right.
$$

Differentiating this equation with respect to $z$, we obtain

Owing to equation (9), we may rewrite this previous equation in the following two forms

Equating : he coefficieats of $x^{i-1}$ in both sides of equation (10), we obtain

This relation enables us to know all the coefficients $b_{n}^{(i)}$ if tw, of chem, say $b_{b}^{(0)}$ and $b_{n}^{(1)}$, we known . Similarly, equating the coefficients of $z^{\text {i-1 }}$ or beth sides of equation (11), we obtain

On the other hand, it follows from equation (9) that if $n$ is replaced by $n+2$, then
from which we easily obtein

Eliminating $b_{n+2}^{(i-1)}$ from equations (1.3) and (14), we obtain

$$
\left.n\left(1+x^{\prime}\right) b_{n+=}^{\prime \prime \prime}-2 a n b_{n}^{\prime \prime \prime}, 1 \prime-1 n-2 i\right) b_{n}^{\prime \prime \prime} .
$$

Similarly, eliminating $b_{n+2}^{(t+1)}$ from both equations, we obtain

$$
n\left(1 \mid a^{\prime \prime}\right) t_{n}^{\prime \prime \prime},-2 n n t_{n}^{\prime \prime}, \cdots(n-2 i) b_{n}^{\prime \prime \prime} .
$$

Replacing here $i$ by $i+1$ and simultaneously solving the resulting: equations with the previous one, we obtain

$$
\begin{align*}
& n\left(1-a^{2}\right)\left(b_{n+2}^{(\prime \prime)}-b_{n 1}^{\prime \prime \prime},: ;=-(2 i+n) b_{n}^{(\prime \prime}-!(2 i-\pi: 2) b_{n}^{\prime \prime \cdot 11}\right. \text {. }
\end{align*}
$$

In this way, if we obtain coefficients $b_{1}^{(i)}$, calculate all the coefficients $b_{3}^{(1)}$, and subsequently find $a l l b_{5}^{(i)}$ etc. Comining this result with the result obtained from equation (12), we conclude that it is. sufficient to directly calculate oniy two of the Laplace coefficients, for exampic $b_{1}^{(0)}$ and $b_{1}^{(1)}$, and to find the other coefficients by applying the relations (12) and (15). Instead of $b_{1}^{(0)}$ and $b_{1}^{(1)}$, the computation of which will be considered in the following section, we can take as the initial quantities the coefficients $b_{1}^{(10)}$ and $b_{1}^{(11)}$ which can easily be found by means of the formulae of the preceding section.

The application of the recurrence relation (12) is not convenient for small values of $\mathcal{O}$ since in this case it is accompanied by a considerable loss of accuracy. The same may be said on the application of formulae (15) for large values of $\alpha$. However, it is necessary to point out that at present, one is rarely in need of computing Laplace coefficients for there are several published tables which give L : values of these coefficients. The best of these tahle is by Brown and Browwer ${ }^{(1)}$, Putting

$$
i_{n}^{\prime \prime} \cdot \frac{a^{\prime}}{11 i_{n}^{\prime \prime}} i_{n}^{\prime \prime}
$$

(1) E.W. Brown and D. Brouwer, Tables for the development of tho Alsturbing functions with schedules for harmonic analysis, Cambridge, 1933.
these authors compcied $\lg G_{\frac{n}{2}}^{(1)} \quad$ (for $n=1,3,5$ with eight decimals and for $n=7$ with seven decimals $)$ for the argument $p=\alpha^{2}:\left(1-\alpha^{2}\right)$, varying from 1.00 to 2.50. They took $t=0,1,2, \ldots, 11$.

## 95. The expression of the Laplace Coefficients in Terms of Definite

## Integrals

Applying the well-known Euler's formula for the computatione of the coefficients of the Fourier series to $x$ xpression (1), we obtain

$$
b_{n}^{d u}=\frac{!}{\pi} \int_{i=}^{a}(1-1 \cdot 2-2 \cos a) \quad \cos i x d x .
$$

(11.

This formula is not useful forcomputing the Lap? ace ccefficients with Large values of 1 , because in this case the function cos $1 \times$ changes sign many times. Moreover, for small values of $\alpha$ the coefficient $b_{n}^{(1)}$ behaves 1ike $\alpha^{\ell}$. This important property is not clear in formula (16). A more convenfent formula can easily be obtained from formula (16). The well knewn relation

$$
\int_{i}^{i} \cos ^{p} x \cos i x d x \quad p(p-1) \cdot .\left(p-i!11 \int_{i}^{0} \cos ^{p-1} x \sin ^{d} x d x\right.
$$

ensbies us to write for any function $f(t)$ which can be expanded by the following uniformly convergent series in the interval $-1<1 .<1$

$$
t(t) \cdot a t^{\prime \prime}
$$

the following relation
whlch has been indicated by Jacobi. Applying this relation to the integral (16), we obtain

This integral can be computed using the formula of quadratures, shown in Sec. 56. This is almost the best method of calculating the Lap?ace coefficients, especially when high accuracy is required.

We can apply Landea's transfolmation
which yields
to equatin in (17). We then easily obtain

In particular, we obtain for $n=1$
or, as we cen easily see,
whence,

$$
\begin{equation*}
b_{1}^{(1)} \int_{\pi}^{4} \int_{0}^{i} \ddot{V} 1-a^{2} \sin =\frac{4}{3} r\left(r_{1}, \quad 2\right) \tag{9}
\end{equation*}
$$

where $F\left(\alpha, \frac{\pi}{2}\right)$ is the complete elliptic integral of the first kind. For $1=1$, equation (19) yie.lds
where
is the complete elligtic integral of the second kind.
The existance of many detailed tasles on the complete elliptic integrals makes the application of formulae (20) and (21) particularly simple, However, the complete elliptic integrals can be easily caiculated In a simple manner. For example, the complete elliptic integral of the first kind can be evalrated by means of Gauss' formula
where $M(a, b)$ is the arithmetic-geometric mean of $a$ and $b$, i.e. the quantity defined by thefollowing limiting transition

$$
\begin{aligned}
& \text { (1) } \quad 1 \quad 10, b_{1}, \quad \therefore=1 a n \\
& a=\frac{1}{\vdots}\left(a_{1} \vdots b_{1}, \quad \therefore \quad V^{\prime} a_{1} b_{1}\right. \\
& a_{n} \frac{1}{I}\left(a_{n}: \mid b_{n}, b_{n} \quad V^{n}, l_{n},\right. \\
& M(a, i) \operatorname{iin}_{n} a_{n} \cdot \ln \| b_{n} .
\end{aligned}
$$

96. Calculation of the i)erivetives of the Laglace Coefficients.

Newcomb's Method
We have seen in the previous chapter that in order to calculate the perturbation function by a series expansion, we have to not only know the Laplace cocfficients but also their derivatives with respect to

These derivatives can be calculated by means of the series which results from differentiating series (3) term by term. However, these serfes converge more slowly than series (3). Hence, this method cannot be of any practical valise.

Differentiating equation (9) term by tern with respect to $\mathcal{X}$, we obtain
or

Consequently,

$$
d b_{n}^{\prime \prime \prime}=\frac{n}{2}\left(b_{n}^{\prime \prime},:+b_{n}^{\prime \prime: ~}!\quad \because x f_{n}^{\prime \prime \prime}\right)
$$

Differentiating this relation and combining it with formulae (1.2) and (15), we easily find a series of recurrence relations which enables us to define the derivative of any order. These formulae, which have been used by Leverrier ${ }^{(1)}$, are however not very practical.

On the other hand, as we have already pointed out in Scc. 88, what enters the expansion of the perturbation function is not the Laplace coefficients, but the quantities

$$
r_{n}^{\prime \prime}-3^{n} n_{n}^{\prime}
$$

and their derivatives with respect to $1 g \ll$, i.e.
(1) Annales de $1^{\prime}$ Ohservatoire de Paris, t. 2, 1856.

$$
\text { ! }{ }^{\prime} c_{n}^{\prime}-\begin{gathered}
!^{\prime}!_{n}^{\prime \prime \prime} \\
\left(d!h^{\prime} x\right)^{4}
\end{gathered} \text {. }
$$

$$
1 \therefore
$$

The most convenient and the most accurate method for the computation of quantities (23) and (24) is the method suggested by liewcomb ${ }^{(1)}$. It consists in the development and the improvement of the method of computing Laplace coefficients which had been previously suggested by Laplace. In the following, we give a brief iccount of this method.

It follows from equation (6) that

Newcomb introduced the following, more general function
so that

$$
c_{n}^{i \prime \prime} \cdots c_{n}^{i \cdot 0} .
$$

We note that equation (5) yields

$$
{\underset{d i}{d}}_{d(A, B, C ; x) \ldots}^{C}{ }_{C}^{A B} f(A: 1, B+1, C+1 ; x)
$$

Erom which it follows that
(1) S. Newcomb, Development of the Perturbative Function and its Derivarives in sines and cosines of multiples of the eccentric anomalies, and in powers of the eccentricities, Astronomical papers, 3, Washington, 1891.

$$
\begin{align*}
& F\left(\frac{n}{2}+j, \frac{n}{2}-i+j, i!j, 1,2\right) \text {, } \tag{25}
\end{align*}
$$

$$
D E\left(A, B, C ; a^{2}\right)=22_{2}^{4 B} \stackrel{A}{1}_{1}^{1} \cdot\left(A+1, B+1, C+1-1 ; A^{2}\right)
$$

Applying this formula, we easily represent the derivative of the function (25) with respect to $\log \alpha$ in the following form:

$$
1) c_{n}^{(i)}-\frac{1}{2}(n+2 i+4 j-1) c_{n}^{\prime \cdot 1}+c_{n}^{(1,1)} .
$$

Applying to both sides of the latter equation the operation $D^{k}$, we obtain the following relation

$$
\begin{align*}
& D^{n+1} c_{n}^{i, \prime} \quad \frac{1}{2}(n+2 i+1 j-1) D^{k} c_{n}^{i,}+1 b^{k} c_{n}^{i, 1+1},  \tag{26}\\
& (n \quad 1,3, \ldots, \ldots, j, k-0,1,2, \ldots)
\end{align*}
$$

which enables us to first obtain $D C_{n}^{i, j}$, then $D^{2} C_{n}^{i, j}$, and so on, if we know the quantities (25). In this way, the quantities (24) become known.

In this way the problem under consideration is reduced to the computation of the quantities (25). We divide this problem into three parts and solve them using the linear relations occurring between any three hypergeometric functions $F(A, B, C, x)$, the parameters of which differ by integral values. For example, using series (5) it is easy to obtain that

Formula (25) yields
where

Therefore, substituting into equation (27)

$$
A=\frac{n}{2}+j, B \cdots \frac{n}{2}+i+j-1, C=i+j, x-a^{3}
$$

we obtain

$$
\begin{equation*}
\left.(2 i+\cdots+n-2) c_{n}^{\prime-1,}\right)-2\left(i+j+a^{2}\right) c_{n}^{\prime \prime}+(2 t-n+2) x c_{n}^{(11 .} \cdots 0 . \tag{28}
\end{equation*}
$$

## Putting

$$
\begin{equation*}
p_{n}^{\prime, 1} \quad r_{n}^{\prime \prime, 1}, \tag{29}
\end{equation*}
$$

we can rewrite this equation as

$$
p_{n}^{\prime \prime \prime} \quad \begin{gather*}
\rho_{n}^{\prime \prime \prime} \\
1-i_{n}^{\prime} p_{n}^{\prime \prime \prime} .,
\end{gather*}
$$

where

Hence, if we know the quantity (29) for $i=k$, we can find its values for $1=k-1, k-2, \ldots, 2$, 1. In order to compute $p_{n}^{k, j}$, we can use the following continuec fraction which Immediately follows irom this same formula (30):
whern for simplification it has keen adopted that

Computing in this manner quantities (29) for $1=1,2, \ldots, k$ and knowing, the value $c_{n}^{o, j}$, we can easily use equation (29) to obtain the values of $c_{n}^{i, j}$ for .11 of values of 1 under consideration.

We sinall now calculate the quantity $C_{n}^{o, j}$ by first noting "hat the quantity

$$
i_{1}^{\prime \prime} i_{1}^{\prime \prime} t_{1}^{\prime}
$$

is very practically computed by means of formulae (20) and (22). Once we know $c_{1}^{(0,0)}, c_{1}^{(2,2)}, \ldots$ by means of the recurrence relations which we are going to deduce.

On the basis of formula (25), we obtain

$$
\begin{aligned}
& c_{n}^{1.1}=1 / 1:\left(\begin{array}{lll}
n \\
\because & !1, \frac{n}{2} & 1 \\
2 & 1, j: 2 ; x^{2}
\end{array}\right)
\end{aligned}
$$

where

$$
A 1=2^{1+1}\left(\frac{n}{2}, j\right) \frac{\left(\frac{11}{2}, j+1-1\right)}{11, j+11}+2+1 n+1
$$

Hence, putting

$$
A \quad \begin{aligned}
& n \\
& \because-j \cdot j,
\end{aligned} \beta \cdot \frac{n}{j}+j+1, \quad i \quad j: 2, x=a ?
$$

and, noting that

$$
C \mathscr{C}(A, B, C ; N-C H(A+1, B, C ; 1)+B, H(A, 1,1,1, C, 1 ;, 1=1 .
$$

we obtata
or
ORIGNAL PAGE !
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This is the required reiation, which enable: as to find $C_{1}^{0,1}, C_{1}^{n, 2}, \ldots$ when the quantities $C_{1}^{0,0}$ and $p_{1}^{1, j}$ are already computed.

It remains for us to consider the computation of the ouantitias in terms of $c_{1}^{0, j}$, we derive a new recurrence relation. From equation (25), we obtain
where

Applying the following properties of the hypergenmetric functions, which can be easily checked $t y$ means of equations (5),
we obtain
from which it follows that

$$
1
$$

$$
\begin{aligned}
& \because \quad \therefore 2 r, \\
& \text { ‘. } n|n \rightarrow(n,-j) \geq| \cdots 1 n \quad 1: p_{n+1}^{:}
\end{aligned}
$$

$$
\begin{aligned}
& r_{n}^{n \prime \prime} \operatorname{Na}\left(\frac{n}{2} ; 1 \cdot \frac{n}{2} \cdot 1, j: 1 ; 2^{2}\right)
\end{aligned}
$$

This formula completely solves the problem of computing the quantities $c_{3}^{0, j}, \varepsilon_{5}^{0, j}$, ... by the vaiues $C_{j}^{0, j}$ which have already been found.

Hence, in applying the Newcomb's method, we have to carry out the following operations.
(1) We compute the quantities $p_{n}^{i, j}$ for the largest of the values of $i=k$ and for all the required values of $n$ and $j$ by means of the continued fraction given by equation (31).
(2) We find all the other values of $p_{n}^{i, j}$ by means the relation (30).
(3) We find the qrantity $c_{1}^{0,0}=b_{1}^{(0)}$ by means of equations (20) and (22).
(4) We compute all the values $\mathrm{C}_{1}^{0, j}$ by using equation (32)..
(5) We obtain $C_{3}^{0, j}, C_{5}^{0, j}, \ldots$ by means of equation (33).
(6) We calculate all the $\mathcal{C}_{n}^{j}, j$ values by means of equation (29).
(7) We finally use equation (26) to fird the quantities required for
expanding the perturbation function, namely,

$$
b^{4} c_{n}{ }^{6} \cdot \begin{gathered}
d^{*} c_{n}^{\prime \prime \prime} \\
\left(1, l_{x}^{\prime} x\right)^{4}
\end{gathered} .
$$

Annotation:
The continued fraction (30) rapidly converges only for small values of $\alpha$. For this reasuns, Iansen, suggested that this fraction should be replaced by

$$
\begin{aligned}
& p_{n}^{\prime \prime} \quad \stackrel{1}{1-a_{1}} \\
& 111 \\
& 1 \text { ": } \\
& 11 \\
& 1 \text { " } \\
& \text { 1. . . . }
\end{aligned}
$$

where
ORIGETAL : : ;
OF POOR QUA...

$$
\begin{aligned}
& \text { i. } \begin{array}{llll}
11 & \because i & \therefore 1 & \because \\
& \because_{11} & 11 &
\end{array}
\end{aligned}
$$

and $a_{m+1}$ and $b_{m+1}$ are obtained $f r o m a_{m}$ and $b_{m}$ by the replacement of $j$ and $n$ into $j+2$ and $n-2$. This formula is a particular case of the following expansion which has been obtained by Gauss:

$$
\begin{aligned}
& \left(\cdot \left(i, H,(i ; 1) \quad 1 \quad \begin{array}{l}
1 \\
1
\end{array} \beta_{1}, x\right.\right. \\
& 1-a_{2} x \\
& \text { 1-䧄, } \\
& \text { 1- . . }
\end{aligned}
$$

where

$$
\begin{aligned}
& x_{1}=\begin{array}{cc}
A & C-B \\
C & C \\
-1 & 1
\end{array}, \\
& \beta_{1}=\begin{array}{ll}
A+1 \\
C+1 & C+1-A
\end{array}
\end{aligned}
$$

$$
\begin{aligned}
& B+2 C+2-A
\end{aligned}
$$

## ANALYTICAL METHODS FOR OBTALNING THE PERTURRATIONS

## JF THE ELEMENTS

97. Transformation of the Differential equations which define the Orbital

## elements

The perturbation functions that correspond to the case, in which the motion of two planets is being consilered, are given by

$$
K=k: m^{\prime} R_{0,1}, \quad R^{\prime}=R^{\prime}: m k_{1},
$$

where $R_{o 1}$ and $R_{o 2}$ are defined by equation (1.), Sec. 86.
Leverrier noted that

$$
n^{2} a^{3} k=(1 ; m) . \quad n^{\prime}-a^{\circ} \quad k:\left(1 \mid m^{\prime}\right)
$$

and, hence, expressed the perturbation functions as

$$
R=\begin{gather*}
m^{\prime}  \tag{1}\\
1: m^{n^{2}, 6^{3}} R_{0,1},
\end{gather*} k^{\prime} \cdot \frac{m}{1, m^{\prime} a^{\prime}: a^{\prime} R_{1,0}}
$$

We already mentioned in Sec. 89 that Leverrier had applied the Eollowing expansion of the pervirbation function ${ }^{(1)}$

$$
1
$$

where

[^1]where the difference $\tau^{\prime}-\tau$ depends only on $i, i^{\prime}$ and $\Omega-\Omega$ as can be seen from formula (35) in Chapter XIII, then

Therefore, the Lagrange equations (41) given in Sec. 13 may be written in the following form

where, we denote as usual the angle of eccentricity by 0 . Since the expressions given by equation (2) do not explicitly involve $i$ and $\Omega$, we then have to eliminate the derivatives of $\mathrm{B}_{0,1}$ with respect to these quantities by replacing them by derivatives with respect to $\tau^{\prime}, \tau^{\prime}-\tau$ and $\mathbf{G}^{-}$. Taking into acc unt equation (3), we obtain

In order to calculate the derivatives of $\boldsymbol{\tau}^{\prime}, \tau^{\prime}-\boldsymbol{\tau}_{\text {and }} J$ with respect to 1 and $\Omega$ it is sufficient to apply the differential formulae of spherical trigonometry to the triangle $\Omega \mathbb{N} \Omega_{\text {, (fig. 11) formed by three nodes. This yields }}$ the following set of equations
from which the required partial derivatives are easily obtained. For example,

Substituting these partial derivatives into expressions (5), and then substituting the resulting expressions into equations (4), we obtain a set of integrable differential equations in their final form.

Leverrier introduced the auxiliary quantities $\mathrm{I}, \mathrm{M}, \mathrm{P}_{1}, \ldots$. I and V by means of the following relations

This enables us to give to the final equations that define the elements the following forms

$$
\begin{align*}
& \frac{d e}{d t}=\frac{d P_{3}}{d t} \cdots \frac{1}{2 a} \operatorname{tg} \frac{?}{2} \cos \frac{d L}{d t} \\
& \cdot \frac{d=}{d t}=\frac{d P_{2}}{d t}+c t \frac{i}{2} \sin i d t  \tag{i}\\
& d i=-\sin (:-2) \frac{d P_{4}+\cos (:-2 t)}{d t}\left(\begin{array}{ll}
d T & d V \\
d t & d t
\end{array}\right) \\
& \sin i \frac{d:}{d t} \cdots \cos (T-U) \frac{d P_{1}}{d t}: \sin (t-\underline{U})\binom{d T}{d t+\frac{d V}{d t}} .
\end{align*}
$$

Because the slopes of the planetary orbits $1, i^{\prime}, \ldots$ are smal1, Leverrier introduced instead of $i$ and $\Omega$ the following elements

$$
p=\lg i \sin !, \quad 4 \quad \text { Hicovil }
$$

The differential equations which define these elements can be written in the fcllowing way

The use of the elements $p$ and $q$ is convenient not only because their perturbations:are small while the perturbations of $\Omega$ may increase as much as possible due to the presence of a small factor in the denominator of the last of equations (7), but also because these quantities can in particular, be easily expressed in texms of the perturbations of the heliocentric latitudes (Sec. 100).
98. The Perturbations of the Elements

Let us denote by $\delta_{i} \lambda, \delta, a, \delta_{1} \in, \ldots$ the first order perturbations of the mean longitude $\lambda$ and the elements a, e, ... . Assuming' that the elements involved in the right-hand side of equations (7) are constants and that they properly define the integration constants we obtain

Substituting expressions (2) for $\mathrm{R}_{0,1}$. into equation (6), and integrating, we obtain
where we denote by $\mu$ the ratio of the mean motions $n / n^{8}$. Introducing these expansions into equations (9), we obtain the first-order perturbations of the elements fo theform
where

Combining in these series the terms for which $j=j^{\prime}=0$, we obtain the secular perturbations. The remaining terms give the periodic inequalities. Denoting the secular part of each of the perturbations, say $\delta_{1^{a}}$, by $\left[\delta_{1} a\right]$, we obtain

Let the secular term in the mean longitude be equal to

$$
\left\lvert\, \begin{array}{ll}
\mid i n
\end{array}\right.
$$

Then, the mean longitude will be calculated up to within the Eirst powers in mass by the formula
, = nt:i: : al +periodic terms

In thls way, if we define the mean motion of the planets by means of the longitudes obtained from the obserytions in two epochs which are divided by a long intervai of tame, as it usually occurs in practice, we then do not odtain $n$ but the quantity

We calculate $a_{1}$ by means of the relation

$$
n: u_{1}^{3}-n_{1} 1 \cdot m,
$$

which is similar to equation

$$
n a \cdot: "(1)
$$

that relates the unperturbed mean motion to the semimajor axis. Since

$$
\because \quad\binom{n_{1} a i}{\left(n_{1}\right.}:
$$

then, disregarding errors in the order of $\mathrm{m}^{\prime 2}$, we obtain the following equation

$$
\text { a } \quad a_{1}\left(1 \cdot \frac{3}{3} \frac{1}{n_{1}} .\right.
$$

Thus, defining $n_{1}$ from the results of these observations, we determine $n$ and a by means of equations (12) and (13). These quantities should be substituted into equations (10), which determine the first order perturbations.

In the calculation of the second- and higher-order perturbations of the mean longitude, we have to use the quantity defined by equation (11). In other words, $n$ should be replaced by $n+\mathcal{L}=n_{1}$. Hence, it is better to write from the very beginning $n_{1}$ instead of $n$ in all the arguments of $D$ than to take into consideration a constderable part of the second-order perturbations in the first approximation. In doing this, we must be aware that the replacement of $n$ by $n_{1}$ can only be done in the argument of D. As far as the coefficient of equations (4) and (6) are concerned, the quantity a will always have values given by equation (13), while $n$ will only be a notation for the quantity
\& | 1 m d $\quad n_{1}$

It is useful to point out that the mean motion of a planet which is given by the tables of the elements is $n_{1}$. This meane that the tahular values of the mean motion includes the constant parts of the perturbations. Conforring with this, the value of the semimajor axi; that are given by the tables and whif are equal to $a_{1}$ in our notations should be replaced durfing the computation of the perturbations by

$$
A \quad u_{1} \cdot \therefore \omega
$$

where

Within the limits of the accuracy accepted,

Hence, the corresponding correction to $1 g a_{1}$ is equal to

$$
د 1 t_{1} u_{1} \quad \frac{2}{3} H_{i_{1}}^{\prime} .
$$

Carrying out the calculation of the first terms of formulae (10), it is easy to find that

Similarly, we obtain for the other planet

It is assumed in the derivation of these formulae that the ratio $\boldsymbol{o f}_{2}=$ $a_{1} / a_{1}^{\prime}$ is less than unity. For each planet, we should take the sum of all those corrections, which correspond to all perturbing planets.

In order to illustrate the influence and character of the periodic pertur'sations, we list the perturbations produced by Venus on the motion of the sun, and calculated by Leverrier by means of equation (10).



We have only listed in this cable the 44 argunents, for which at least one rerm exceeds $0^{\prime \prime} .05$. Leverrier conputed all the terms exceeding $0^{\prime \prime} .001$ and his table included 123 arguments. (1)

We note that in the case when Earth is one of the planets under consideration,

$$
\because \quad 11 \quad 1, \quad 1, \quad 0 \quad .-
$$

In this case, argument D becomes

## (1) The cuefficients whose valucs were less than $0^{\prime \prime} .001$ were reflaced by dashes.

$\%$ The perturbations of the Elementy $\therefore$ cond Order wf.th Respect to

## Masses):

It have been able to express ' ans of equations (10) the first order pe: under conside:atom. We nor, $\therefore$ tar the calculation of the second order corrections. For thes prirpora, we have to replace the elements a, e, ..., $E^{\prime}, e^{\prime}, \ldots, a^{\prime \prime}, e^{\prime \prime}, \ldots$ of $a^{\prime \prime}$ the planets in the expression (1) of the perturbation function by


In this way, equations (4) lead to the following equations for the calculation of the perturbations of the second order

$$
(10)
$$

Substituting here expressions (2) and (10), and doing the necessary multipitation of the series, we obtain on the right-hand side a serfes developed in cosines and sines of arguments of the type
or

$$
\left(j n\left|\cdot j^{\prime} n^{\prime}\right| j^{\prime \prime} n^{\prime \prime} \vdots . . .\right) f \mid \text { conat. }
$$

Consequentiy, the integration of equations of tie type (16) introduces
divizors of the form $j n+j^{\prime} n^{\prime}+j^{\prime \prime} n^{\prime \prime}+\ldots$. If these exists a group of integers $j, j^{\prime}, j^{\prime \prime}, \ldots$, the absolute vaíue of each is nor lange, for which the sum $j n+j^{\prime} n '+j^{\prime \prime} n "+\ldots$ is small, then the corresponding second-order perturbation will je particulariy large. The pariod of this perturbation, which is equal to $360 /\left(j n+j^{\prime} n '+f^{\prime \prime} n^{\prime \prime}+\ldots\right)$, will be quite great.

As a re-ult of the second approximation, we obtain for each element an expression of the type

The computation of the second-order perturbations of the elements is quite tedious since it involves a large number of terms in equations of the type (16). This difficulty beccnes more significant when we carry out tine calculation of ehtrd-order corrections. In the following chapter, we shall see that it is much easier to calculate second- anc higher-order perturtations in the coordinates.

When Leverrier studied tie motion of Mercury, Venus, Earth and Mars, be could confine himself to the calculation of a few second-order terms. For other planets, and in particular for Jupiter and Saturn, one has to take into consideration not only a large number of second-order perturbations, but also some third order perturbations. Leverrier's work on the ca\&vulation of the fercurbations of these planets nas becn continued by Gaillot.
100. The transformation of perturtations of the elcments into perturbations of coordiaates. Construction of Tables.

After calculating the pertuibations $\delta_{1} a_{1}, \ldots, \delta_{1} \lambda$ or the elements and the mean lungitude, $n=$ can derive general expressions for the coordinetes of the planet. Ne first considec the longitude in the nrbit $w$. It is equal
for the unperturbed motion, to
, ;
19
where we denote by $f$, the equation of the centre (Sec. 82)

Here,

The mean anomaly is equal to

$$
i 1, \quad \pi \quad \text {, } 19
$$

Hence, replacing in ecuation (17) $\lambda, \pi$ and e by

$$
\therefore \quad \therefore 1, \quad \pi \because r \quad r \quad \therefore \prime
$$

and confining nurselves $\ddagger 0$ quantities of the first order with respect to mass, we obtain in the first appısximation the following expression for the perturbation of the longitude

This formula is usually only applicd to the inclusion of the shortpertodic perturbation of $\lambda$ and the perfodic perturbations of $\pi$ and e. The other perturbations uf these elements will be best of ail taieen into account in the following way.

Formula (18) enables us to compute a table giving the equation of

 contigunus miday 0 January leod, waye

When we use this iable, we make the argument (19) out of the values of $\lambda$, that have already been corrected for the long-periodic perturbations, and the values of $\pi$, in which the eecular parts of the perturbations have been included. In this way, only the periodic perturbations of $\pi$ and the short-periodic pe:turbations of $\lambda$ renain to the share of the corrections $\delta_{1} \lambda$ and $\delta_{1} \pi$, invoived in formula (20). On the otner hand, the influence of the secular part of $\delta_{i} e$, whict we have denoted by $\left[\delta_{1} e\right]$, is expressed by the following relation

This can be taken into account quite separatcly by means of sperial tables, which give t'ese quantities in terms of the argument M. For Earth
where $T$ is time given in centuries and measured from the above-mentioned initial moment. Hence the influence of the secular forturbations of the eccentricity on the equation of the centre is taken into consideration by the quantity

$$
\begin{aligned}
& \text { - } 0^{\prime} .011 \% \text { - } 1.11
\end{aligned}
$$

In Newcomb's tables for the motion of the earth (Astronomical Papers, Vol. VI), the quotient obtained by dividing this quantity by $T+0.0030 \mathrm{~T}^{2}$ is given by the argument M . It is worthwhile noting that the term proportional to $T^{2}$ expresses the contribution of the second-order secular perturbations.

Thus, the quantity $\delta_{1}$ e appearing in equation (20) may be understood as the aggregite of only the periodic terms.

The computation of the sum of the periocic terms involved in equation (20; is simplified by constructing epecial tables, each of which give the sum of theterms that depend on a given argument $j \boldsymbol{\lambda}+\boldsymbol{j} \lambda^{\prime}$

- The most important terms wfill be found by these tables. The sum of the remaining terms of equation (20) may te obtained $k y$ means of a table with two entrances corresponding tc the arguments $\lambda$ and $\lambda$

Let us now consider the perturbstions of the logarithe of the radius .vector. For the perturbed motion, this logarithm is given by (Sec. 82)
where $A_{0}, A_{1}, \ldots$ are functions of e. Consequenty $y$,

Formula (21) onables us to construct: a table, which gives the value 3 of the logarithn of the radius-vector for given values of $e$ by the argument:

$$
11, \quad \therefore
$$

# OAIGINAL PAGS IS <br> OF POCR QUALITY 

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For Earth, such a table has been constructed by Newcomb on the basis of the following formulae

In constructing this table, we agree to include in $M$ the iong-periodic pare $\delta_{1} \lambda_{1}$ for which a special table may be constructed together hith the secular part $1: \mathcal{S} \pi$. The influence of the secular part of $6, \hat{\lambda}$ or $\boldsymbol{\ell}_{G r}$ can easily be taken into account by neans of a special table. If we take the $\operatorname{sun} a s$ an example, we find that the sum of the terms of equation (22), which correspond to the secular part of $\delta_{1} e$, are equal to

Newcomb constructed a table, fron which we find the quotient resulting fron the division of this quantity by $T+0.0030 \mathrm{~T}^{2}$, by the argument. M. The sum of the periodic terms, which remain in formula (22) after having made all the simplifications, is partially computed for each of the Ferturbing planets by means of tables having a single entrance. The remainder can be taken from a table heving two entrances.

We shall finally consider the determination of the heliocentric longitude $\ell$ and latitude $b$. We have seen in Sec. 85 that
where
is the reduction to the ecliptic, and shereby $u$ is devoted as the argument of latitude

$$
\begin{equation*}
u=u \cdot u \tag{2}
\end{equation*}
$$

In this way, taking an arbitrarily given value of $i$, we can construct two tables, which give by the argument $u$, the values of $R$ and $b$. Putting, as before,

$$
p \quad \ln i \sin !, \quad \varphi=x \operatorname{th} i(0, \therefore .
$$

we obtain

According to the generally accepted rules, we take within argument (25) the values of whini include all the perturbations. Therefore, when we compute $\delta_{1} b$, we can consider that in the latter equation, increments are only given to $1, p$ and $q$. Whence,

The secular perturbations of $\Omega$ are taken into account by including them in the argument (25). The influence of the secular perturbations of $i$ are are evaluated by means of the following formula
which can be reduced for all the large planets to the following form

$$
|i, n| \quad A, 110: r
$$

This formula can be replaced by a table having $u$ as an argument. It is hence necessary to includc by means of equation (26) the faflence of orl.y the pertodic perturbations of $i, p$ and $q$. This can be done in anclogy to the previous cases.

We can take into consileration the influence of the secrilar perturbation of $f$ on the reauction to the ecliptic, given by equation (24), in a very simple manner. The periodic perturbations of 1 do not significantly change R. In order to show this, we write down for Mars, the expressions of these quantities

These expressions constitute the basis of the corresponding tables given by Newcomb.

## Annotation

We have considered thederiv: 'ion of the formulae, that define the first-order terms in the perturbations of the coordinates. The secondorder terms can be obtained exact-; in a similar way. However this $+$ requires a great deal of tedious work.

The tables constructed by Leverrier give the perturbations of the coordinates $r, \ell$ and $\dot{b}$ for Mars, Earth, Venus and Meirury, As regards the other plaze:s, for which second-order perturbations play an important role, Leverrier has only given tables which enable us to find the osculating elements for these planets at any moment. In order to find their coordinates, we have to apply the conventional iormulae of the elliptic motion.
101. The computation of the secuiar terms by Gauss' methoc'

The coefficients of the secular terms in the equations, which define the perturbations of the elements of a given planet, must be calculated more accurately than the coefficients of the other terms because the influence of the scular terms increases with time. The method of computaticn, developed in the previous sections, produces the coefficients the greatest powers of eccentricities and mutual slopes of the orbits which are kept in the expansion. Gauss suggested (1818) an alternative method for calculating the first-crder perturbations, which enables us to find them independentiy from the other perturbations. This method does not imply the expansion of the perturbation function in a series. Hence, it can be equally applied for any eccentricity and for any slope of the orbit.

In section 12, we obtained equations (37) which expiess the derivatives of the elements in terms of the components of the perturbing accelerations. These fomulae are of the components of the perturbing accelerations. These formulae are of the form
where

$$
S_{1} \quad \begin{gathered}
1 \\
\text { alin }
\end{gathered} \quad T_{1} \quad \begin{gathered}
1 \\
\end{gathered}
$$

and $\underline{S}, \underline{T}$ and $\underline{W}$ are the components of the perturbing accelerations.
When the perturbing acceleration is caused by the gravitation of a single planet, $P^{\prime}$, the components $\underline{S}, \underline{T}$ and $\underline{W}$ are equal to the dexivatives of the perturbation function in the direction of the radius vector and in the two difections perpendicular to the radius vector. One of these perpendicuiar directions is taker in the orbital nlane and the other in the normal direction to this plane (Sec. 11).

We have already pointed out in section 88 t⿵冂⿱一口䒑寸t the second part of the perturbation function dees not produce secular terms．We can therefore replace the perturbation function by its principal part

$$
m^{\prime} s^{-1}
$$

where $\Delta$ is the distance between $p l a n e r s ~ P$ and $P^{\prime}$ and $m$ the mass of planet $\mathrm{P}^{\prime}$ ．Using the series expansion of the perturbation function which we have already studied，we can write each of equations（27）in the following way

$$
\begin{align*}
& d i  \tag{29}\\
& d l
\end{align*}\left[\begin{array}{l}
d i \\
d l
\end{array}\right] \quad A_{, j} \cos \left(j H \cdot j^{\prime} M^{\prime} \mid O\right)
$$

where we denote by $\left[\frac{d i}{d t}\right]$ the constant term of the expansion．Assuming as before that the mean motions $n$ and $n '$ are incomensurable，we can set the quantity $j n+j^{\prime} n$＇equal．to zero oniy in the case in whick $j=j^{\prime}=0$ ．After integrating，we obtain

This means that the computation of the first－order secular perturbations is equivalent to the computation of the constant terms in expansions of the type（29）．

It follows from equation（29）that

$$
\left|\begin{array}{l}
d i \\
d i
\end{array}\right| \quad 4 \pi=\int_{i}^{0} \int_{i}^{0} d i d d d d M ;
$$

In other words，the unknown constant term is obtained by averaging the quantities（27）over the varlables $M$ and $M^{\prime}$ ．The variable $M^{\prime}$ which appears fin expressions（27）depends orly on $S_{1}, \eta_{1}, W_{1}$ ．Hence，we shall first integrate with respect to $M$ and then with respect to $M$ ． Putting
and similarly defining $T_{c}$ and $W_{0}$, we finally obtain

$$
\begin{align*}
& \left|\begin{array}{l}
d i \\
d t
\end{array}\right|=\frac{1}{2 r} \int_{i}^{:} r \cos u W_{n} d, I t
\end{align*}
$$

We shall first of all consider the computation of integrals (30). On the basis of the above-mentioned arguments concerning the replacement of the perturbation function by its principal part given by equation (28), we can consider integrals
as components of some force of gravitation, which corresponds to the pntential

$$
\because \int_{i}^{\pi} 1 /
$$

This potential has a quite simple mechanical interpretation. As a matter of fact, let us imagine that the mass of planct $P^{\prime}$ is distributed over the orbit of this planet in such a way, that each element of mass dm" which will be distributed over one of the linear elements of the orbit, will be proportional to the time interval ct during which the planet passes this linear element. Accordingly,

$$
\begin{array}{cc}
d / n^{\prime} & d t \\
m i^{\prime} & 1
\end{array}
$$

where we denote by $P^{\prime}$ theperiod of rotation of planet $P^{\prime}$.
However

$$
d t=\frac{n^{\prime} d t}{n^{\prime} t^{\prime}-\frac{d u^{\prime}}{2 \pi}}
$$

and, consequently

This is nothing eise but the potential of the elliptic ring produced by distributing the mass of planet $P^{\prime}$ along its orbit in the abovementioned way. In this manner, our problem is reduced to the wilculation of the components of the force of gravitation induced by a material elliptic ring, the density of which is defined by Kepler's law. We shall not do these calcuigtions here. We only point cut that the unknown integrals (32) can easily be expressed in terms of ellfptic functions;

We have thus found the way to calculate the values of integrals (32) and consequently the quantities $S_{o}, T_{o}$ and $W_{o} i_{n}$ any arbitrary point of space. We shall now consider the computation of the quantities given by equations (31). The integrals involved in these equations can be calculated numerically. We calculate each of the integrands, for example $r \cos u W_{o}$, for different values of $M$, and then take the mean values of the given quantities. Let us, for example, consider the expansion

We spply to it the conventfonal methods of harmonjc analysis. We denote by $\Phi_{0}, \Phi_{1}, \ldots, \Phi_{k-1}$ the values of the functions (33) that correspond to


In this case,

$$
\left|\begin{array}{l}
d i \\
d t
\end{array}\right| \text { a } \quad \frac{1}{k}\left(u_{1}, \|_{1}|\ldots . .| u_{k}\right)-i_{k}-u_{.} \quad a_{3}-\cdots
$$

The serles (33) converges so rapidly, that even for small values of $k$, the secular perturbations are obtained with a high accuracy. It is easy to show that the error obtained in calculating the eccentricities and mutual slopes of the orbits will we of the order of $k-1$ for the secular perturbations of $\lambda, \Omega, e$ and $\pi$, and ,f the order of $k$ for the secular perturbations of the mean longitude of the epoch.

Instead of tne variable $M$ in the integrals (31), the eccentric anomaly is often introduced by means of the following relation

$$
\text { d:1 } 11 \text { c enl. d!. }
$$

That is to say, the averaging over $M$ is replaced ry the averaging over E. As we have already pointed out in Sec. 69, pofnts corresponding to equidistant values of $E$ are distributed along the orbit more uniformly than points (34) in the case in which the value of the eccentricity e is significant. We should note, however, that the advantages of applying E as an independent variable are not above reproach. Considering the unirorm distribution of points along the nrbits, the parts of the orbit over which the planet rapidiy passes and those over which it passes slowly, have equal weights. Hence, we should not conclude that the replacement of $M$ by the variable $E$ will significantly reduce the value: of $k$.

An interesting generalization of the restricted three bory problom was given by facou ${ }^{(1)}$. He considered the motion of the material point $P$,
i1) P. Fator, Sur le mouvement d'un point materiel dans un champ de gravitation fixe, Acta Astronomica, Ser a, Vol. 2, 1931, 401,462.
having an infinitesimally small mass, in the field of gravitation of the central body $S$ and some material elifp:ic rings having lhe abovementioned density distributions.
102. Lagrange's Differential Equations for the Determination of

## Secular Pertur-bations.

When we study the notion of a planet during a relatively short period of time, in the order of a few centuries, we can confine ourselves to first-order secular perturbations. In this case, the secular perturbations are best of all calculated by means of Gauss' method. This method enables us to obtain the secular perturbations of the eccentricities and of the slopes in a simple way with an arbitrarily high accuracj. If we are interested in longer perinds of time, we have to apply the methods developed in sections 98 and 99. By these methods, we are able to obtain the second-, third-, ... etc. order secular perturbations. Naturally, the amount of krk required by these calculations sapidly increases with increasing order of pertuxbations. riPractically, we can hardly calculate the perturbations of orders higher than the third. Finally, if it is necessary to describe the motion of a planet during a considerably large interval of time, the decisive role will then be played by zero-rank terms. These are the terms in which the perturbing masses are raised to some puwer appear as multiplying factors (Sec. 15). Lagrange suggested an alternative method for calculating the secular perturliations of the eccentricities and the mutual slopes of the orbits. The main point of this method is that it can immediately lead tc us to just the zero-rank tems but will a relatively low accuracy. In the following, we shall give a brief account of the method.

Lagrange investigated the possibility of integrating equations (41)
of section 13 under the condition of neglecting all the periodic terms In the perturbation functinns appearing on the right-hand side of thesc equations. It is difficult to saywithout special considerations, that the integration of these dininished equations could lead us to the precise values of the perturbations, which would have been obtained in the form of secular terms if the exact equations were integrated by the method of successive approsimations. However, we still believe that the result of integrating these diminished Equations will satisfactorily elucidate the character of motion of a planet during an extremely long period of time. The results obtafned will thus be of particular interest for cosmological investigations.

Replac ang the perturbation function involved in the right hand si'e of the Lagrange equations by its secular terms, we can confine ourselves to the series expansion of this function in which terms involving higher second powers of the eccentricities and of the slopes are neglected, Only under this limitation, can we exacily carry out the integration of the Lagrange equations.

Let us now consider the expansion (35), Sec. 90. Noting that the second part of the perturbation fuaction does not lead to secular terms, we obtajn within the linits of accuracy, that the secular part part of the perturbation function will be given by

Since
and, as formula (35) of section 39 indicates, the difference $\tau^{2}$ - $\tau$ is a small quantity being of the seconc-order with respect to the slopes, we then may replace $\cos \left(\Pi^{\prime}-\Gamma\right)$ by $\cos \leqslant \pi^{\circ}-\Pi$ ) alyays remaining within the limits of accepted accuracy. On the other band, and within the same accuracy,

Hence, we can write
where we denote by $M_{C, 1}, N_{0,1}$ and $P_{0,1}$ coefficients which only deperd on a and $a^{\prime}$. According to the formilae developed in sectior 88 we can constder that these coefficients are symatric functions of $a$ and $a^{\prime}$.

Acccrding to Lagrange, we replace the elements e, $\pi$ and $i$, d 2 by the following variables

We can then rewrite expresion (35) in the following manner
where the var:ables $h^{\prime}, \ell^{\prime}, p^{\prime}$ and $q^{\prime}$ are defined by equat:-ons similar to semations (36) and (37).

We now deduce tue differential equat.jons that define the nex elements given by equations (36) and (37). Since
then, using equation (41) of section 13, and noting that
we easily obtain

Differentiating expressions (38) with rospect to $h, l$ and $i$ yields first-order powers of those small quantities, in terms of which the expansion is carried out. Hence, neg1ecting the terms involving third powers, we obtain

$$
1: 11
$$

When we are interesteü in the secular perturbation produced by the interaction of the two planets $P$ and $P^{\prime}$, we have to replace the quantity $R$ in equation (40) by the expression given by equation (38). We shall however consider the general problem which involves an arbitrary number of interacting elements and try tc find the secular terms of their elements. We denote by $m_{0}, m_{1}, m_{2}, \ldots$ the masses of planés $p, p^{\prime}, p \prime$, ...., and by $a_{o}, a_{1}, \ldots ; e_{0}, e_{1}, \ldots$ their elements.

Appiying equation (40) to planet $p|\mu|$, we obtain

$$
\begin{array}{cccccc}
d / r & 1 & 1, & 1:! & 1 & 1 \\
\text { if } & m m & \therefore, & 1: & \ddots & 1 .
\end{array}
$$

where, due to equations (38),

$$
\begin{array}{lll} 
& -419- & \text { ORIGNAL PAGE IS } \\
\text { OF POOR QUALITY }
\end{array}
$$

In order to avoid making exclusions during the sumation, we consider that

$$
\because \quad \therefore \quad / 1 \quad \text { U }
$$

Denoting, in short,
the result of substituting expression (42) in equation (41), is rewritten in the following way:
where

$$
\begin{equation*}
A_{,}, \because(!, 0) \cdots(!, 1) \quad(:, \because)-: \cdot \tag{44}
\end{equation*}
$$

Since each of the coefficients $N \mu, v$ and $P \mu, v$ is a symnetric function of a $\mu$ and a $\nu$, then the following relations hold

$$
\begin{array}{ll}
m, n, a,(a, i) & m, n a(n,!)  \tag{0.3}\\
m, n_{n} a_{i}|n, v| & m!a^{\prime}|v,!|
\end{array}
$$

In this way, the determination of the secular perturbations of the eccentricities $e_{0}, e_{1}, e_{2}, \ldots$ and perihelion 1 gitudes $\pi_{0}, \pi_{1}$, $\pi_{2}, \ldots$ is reduced, within the accuracy accepted to the integration of equations (43).

We now consider the variables defined by equation (37), which dctermine the position of the orbits. We, first of all, find that

$$
\begin{aligned}
& \begin{array}{l}
\text { w rec isme } \\
\text { d }
\end{array}
\end{aligned}
$$

Taking into account equations (41) in section 13, we easily obtain

In the present case, since the equantity R is expressed by equation (42), then $\frac{\partial R}{\partial \epsilon}=0$ and $\frac{\partial R}{\partial \pi}$ is a second-order quantity. Fence, neglecting terms involving third order powers, we oitain

Substituting expression (42) into similar equations that define $P_{\mu}$ and $a_{m}$, we finally obtair the fnilowing system

$$
\begin{array}{ccccccc|}
\ldots: & 1 & & 11 & . j & . & . \\
\ldots & 11 \\
\ldots & i & f & 1 \cdot .11: & 11.111 & . & 11
\end{array}
$$

We shall consider the solution of systems (43) and (43) in the next section. We here confine ourselves to obtain the outstanding first integrals of these systems that were first discovered by laplace. We multiply equations (13) by $\quad m_{\mu} \quad \eta_{\mu} a_{\mu}^{2} h_{\mu}$ and $m_{\mu} n_{\mu} a_{\mu}^{2}$ respectively, add them term by term, then sim the results. This yields,
due to relation (45),
or
where $C$ is an arbitrary constant which may be defined by means of the initial conditions of motion. Taking equation (36) into account, we finally obtain

Simflarly, we obtain the following first integral of equations (48):

$$
\begin{equation*}
\text { Nm, " } \operatorname{lin}_{i=}^{*} \tag{il}
\end{equation*}
$$

At present, the excentricities and the slopes of the orbits have small values. Hence, constants $C$ and $C^{\prime}$ are also small. Due to the fact that all the terms which constitute the sums (49) and (50) are positive, Laplace considered it possible to conclude that in the future, $e_{\mu}$ and $i_{\mu}$ will always remain positive quantities. This conclusion is only valid as far as it concerns planecs, whosw mass constitute a large part of the sum of the planetary masses. If the mass of a planet is very small, its excentricity and slope can be sufficiently large without violating equations (49) and (j0).

Since

$$
\text { n. il : nt. a } \quad \stackrel{i}{i}
$$

then, neglecting second-order quantities relative to the masses, we can replace equations (49) and (50) by

$$
\begin{aligned}
& \sum_{1,} a_{.} e_{1} \text { cont } \\
& \min _{1} a_{n} x_{1} \text { connt. }
\end{aligned}
$$

103. Trigonometric Expressions of the Secular Perturbations

We shall now consider the solution of equations (43). According to the general theory for integrating systems of linear differential equations with constant coefficients, we search for the particular solutions in the forn of
where $s, \beta$ and $L(\mu)$ are constants. Substituting these expressions Into equations (43), we obtain the following system for deternining $S$ and $\mathrm{L}:$
where $m+1$ is the number of planets, and

Equations (45) show that

$$
d, \quad A_{1}
$$

Hence, the determinant of the previous system

$$
D(s)=\left|\begin{array}{ccccc}
A_{1} . \cdots s . & i_{0,1} & \cdots & A_{1, m}  \tag{5.1}\\
A_{1,0} & & i_{1.1}- & n_{1} & \cdot
\end{array} A_{1, m}\right|
$$

is symmetric with respect to the main diagonal. Let us denote by $S_{o}, S_{1}, \ldots, S_{m}$ the roots of the equation

$$
11(s) \cdot I .
$$

(5)
where $L_{i}{ }^{(\mu)}$ are the values of the coefficients, which will be obtained if we put $S=S_{i}$. Gne of these coefficients will ramain arbitrary. Hence, we can put
where $C_{0}, C_{1}, \ldots, C_{m}$ are arbitrery constants, and $q_{\lambda}^{f+1}$ are known numbers.

Since the parameter $\beta$ involved in each of equations (51) remains arbitrary, then denoting by $\beta_{0}, \beta, \ldots, \beta_{m}$ another $m+1$ arbitrary constant, we can write the general solution of system (43) in the following way

$$
1.1
$$

where we have put

$$
\begin{array}{ll}
!^{\prime} & \text { a.l }_{1}^{\prime} n
\end{array}
$$

$$
1.1
$$

These equations are known as the secular equations. The properties of their solution depends strongly on the nature of the roots of equation (55).

Lagrange confined himself to the calculation of the roots of the
secular equations using the values of constants that are obtained for the solar system. He found that these roots were real and nequail. Laplace was able to prove by means of integral (49) that equation (55) could not have any complex roots whetever the values of the constants were. Indeed, if there were any complex roots amongst the roots of equation (55), then the correspond term of equation (56) would include an exponential function. In thiscase the sum $h^{2}+e^{2}$ would tend to infinity when $t \rightarrow \pm \infty$. This would violate relation (49).

Laplace tried to prove the absence of equal roots of the secular equations by using similar argments. In doing this he made an error. He considered that in the presence of equal roots, there should be in the general integral (56), certain polynomials of $t$ multiplying the trigonometric functions, which would certainly violate equation (46). However, it was almost simultaneously pointed out by Weierstrass (1858) and Somov (1859) that for equal roots, it was not at all necessary for $t$ to appear outside the signs of the trigonometric functions.

When only two planets are considered, the absence of equal roots of th the secular equations is established quite simply by means of direct verification. The impossibility of the presence of equal roots was shown by Seelinger in 1878 for the three body problem.

In order to obtain the limiting values of the eccentricity e $\mu$, we square expressions (56) and add. This procedure yields.
from which it is clear that

We note that if one of the coefficients (57) in formulae (56), say $M \rho_{\rho}^{(m)}$, exceeds by an absolute value the sum of all of the other terms, then the perihelions of planet $\mathrm{P}^{(f)}$ will. have a translationaj motion with an ayerage velocity S , Indeed, combiring together equations (56), we obtain

By condition,

$$
\mid M I_{:}^{(i)}, \sum M,
$$

Hence, $\cos \left(\pi_{\mu}-S_{\rho} t-\beta_{\rho}\right)$ will never be zero, and thus

$$
\nabla_{2} s t \mid p \quad k \cdot 180+i_{p}(t) .
$$

where $k$ is an integer, and the last tenn satisfies the condition

$$
-90 \quad i_{1,}(t) \quad \vdots 90
$$

Hence, the perinelion of the planet under consideration will never be displaced by an angle more thar $10^{\circ}$ from a point moving with a uniform velocity S $\rho$.

Considering equations (48) which define $\mathrm{p}_{\mu}$ and $\mathrm{q}_{\mu}$, sirce they have the same form as equation (43), we can then imediately Write their general solutions in the folloring way

$$
\left.\begin{array}{l}
p_{i 4}=\sum N_{1}^{\prime, 1} \sin (3, t-1 \cdot \%)  \tag{59}\\
q_{i A}=\sum N_{r}^{\prime \prime \prime} \cos (3, t \mid \eta)
\end{array}\right\}
$$

where, we denote by $\sigma_{0}, \sigma_{1}, \ldots, \sigma_{m}$ the roots of the secular equation

$$
\begin{equation*}
D^{\prime}(j)=-0 . \tag{60}
\end{equation*}
$$

which has the same structure as equation (55), with the orly exception that in the present case

$$
\begin{equation*}
\lambda_{1, n}-a_{1,}^{a_{i} m_{1,} n_{1,}}(\mu, v) \tag{6!}
\end{equation*}
$$

for $\mathcal{M} \neq \nu$.

Annotation:
Since all the quantities $A_{\mu \mu}$ ard $A_{\mu \nu}$ involved in the secular equations (55) and (50) are of the order of the planetary masses, then the roots of these equations will be first-order quantities relative to these massts. Hence, expanding expressions (56) and (59) in powers of $S_{\lambda} t$ and $\sigma_{\lambda} t$, we obtain zero-rank-terms.
104. Secular Perturbations of Large Planets

The numerical values of the coefficients of formulae (56) and (59) were for the first time obtained by Lagrange himself. His resulty are only of historical interest because Uranus could ...st be involved in these computations, and moreover, hypothetical values were taken for the masses of Mercury, Veuus and Mars, obtained by means of multiplying the valumes of these planets by some assumed density.

In 1839, Leverrier repeated these computations with better values for the constants and taking Uranus as an example. However, the influence of Neptune, which was subsequently discovered by Leverrier in 1846 , was not yet taken into account. The most con lete results obtained in this field are those obtained by Stockwel1 ${ }^{(1)}$, who suggested
(1) J.N. Stockwell, Smithsonian Contributions to Knowledce, Vol. 18, 1870, Washington 1873.
on the bases of niscomputations the following values for the masses, mean anual motions and semimajor axes

## Table 1

| planet | $1: m$ | $n$ | a |
| :---: | :---: | :---: | :---: |
|  |  | - |  |
| Mercuriy | 4805751 | 5.381016 .2 | 0.3570987 |
| Nenus. | 3910000 | 2100614.138 | 177.3:3\%; |
| Earth | 368889 | 12.5177111 | i 10100100 |
| Mars. | 2680637 | 68900.0.902: | 15236788 |
| Jupater. | 1047.879 | 1092.56 .719 | 6.202798 |
| Batamen. | 3511.6 | $4: 35127$ | 0.53885 |
| Itranus. | 24905 | $1.542450 \cdot 4$ | 11183581 |
| Nepture. | 18780 | , 873.643 | 31103380 |

Furthermore, he used the values of the required elements for 1.0 january 1850 and used as a basis the exiptic and equinox of 1850.0 . Thesf reines are given in table 2.

Table 2

| 1 | Planet | $e$ | $\pi$ | $i$ | U |
| :---: | :---: | :---: | :---: | :---: | :---: |
|  |  |  |  |  |  |
| 0 | Mepcuny | 0.20j617 | 75 T'0'0 | 7.088 | 16',33'3'2 |
| 1 | Venus | Ouni8118 | 129) $2 \mathrm{~K} \quad 51.7$ | 32331.1 | 7.519129 |
| 2 | Earth. | 0.0107712 | 1002111.1 | $0 \quad 0 \quad 11.0$ | 000.1 |
| 3 | Mars. . . . . | 00931.324 | $\begin{array}{llll}333 & 17 & 17 & 8\end{array}$ | 1503 | $\begin{array}{lllll}48 & 2.3 & 36 & 5\end{array}$ |
| 4 | Jupiter. . . . | $0.0182: 888$ | 1151533.1 | 18189 | SK 5420.5 |
| 5 | Saturn. | 0.055 995 | 90 if 12.0 | $\because \quad 3124$ | 112 19:0 |
| 6 | Unanus. | $0.016: 1.15$ | $\begin{array}{lllllllllllll}170 & 17.0\end{array}$ | (1) 11.20 .9 | 7.314111 |
| 7 | Neptumc. . . . | 0.0091740 | 50163 | 1.17011 | $\begin{array}{llll}130 & 7 & 15.3\end{array}$ |

He obtained values for the parameters involved in equations (56) and (59) and which are given in tables 3 and 4.

The values obtained by Stookwell enables us to determine the limics, within which tine mean values of the eccentricities and slopes vary. By the term, mean values, we mean as usual the values of the elements
Table 3

Table 4

| 2- | 0 |  | 2 | 3 | 4 | 5 | 6 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $a_{j}=$ | $-5^{7} .126112$ | $-6 " .592128$ | $-17 \cdot 3933=0$ | $-18^{\prime \prime} .468014$ | $-0^{\prime \prime} .661$ ofe 6 | -2".916 188 | - 5 5-307 |
| $\gamma_{2}=$ | $21^{\circ} \mathrm{E}^{\prime} 26^{\prime \prime} .8$ | 132-4, $57^{\prime \prime} 8$, | $292^{-} 49^{\prime} 55^{\prime \prime} .4$ | $251^{-45^{\prime}} 8^{\prime \prime} .6$ | $20^{\circ} 31.24^{\prime \prime} 6$ | $135^{\circ} 56 \cdot 100.8$ | U6 152: 2 |
| $N_{4}^{* *}=$ | $\therefore 0.1210760$ | +0.928 3520 | $\div 0.0015240$ | + 0.0036775 | +0.014788 | $\div 0.0031283$ | -0.00202 |
| $A_{a}{ }^{17}=$ | $\cdots 8.0148570$ | -0,0078364 | -6008483 | -0.622 4278 | $\div 0.0013565$ | $+0.0018: 08$ | -90.0:2 |
| $x^{\prime \prime}=$ | -0.01065\% | -0.0663230 | - acos 9546 | + 0.0247708 | +000132.1 | $\div 0.601928$ | - 20285 |
| $N_{*}^{* *}=$ | A. 0.0021250 | -cten 3250 | - 0 rs005:2 | -0.037 5051 | - 20012586 | $-0.6011557$ | -60.2.9 |
| $N^{4 .}=$ | -0.600 0252 | $-0.000045$ | -0.60j 0025 | -0.000 0001 | +0.0011993 | + 0.000875 | - Wixes is |
| $A^{t_{1}}=$ | -0.0:00320 | $+0.000050$ | -0.0:00214 | -0.000 0005 | $+0.0011577$ | $+0.0007150$ | - \%i*. $\because$ |
| $N_{n}^{(s)}=$ | $\div 0.0000280$ | -0.00000:0 | $-0.0000 .212$ | +00000:60 | - 0.0011238 | -001765:2 | -0.couse |
| $\mathrm{N}_{6}=$ | -20000068! | -0.060r004 | -0000002 | -0.ccosose | $-0.0117 .82$ | - ocisoriz | - C.'.: |

in which both the secular and the periodic perturbations are taken into consideration. These iimits are given by the following tabie

| Phanet. | ; Eccentricity. |  | Slope |  |
| :---: | :---: | :---: | :---: | :---: |
|  | Mili. | Mix. | Min | Mrs |
| Merc uny | 0.1 | 0. ${ }^{1 / 2}$ | 411 | $\because 11$ |
| Kenus | 0 | $0.1 / 1$ | 0 | 311 |
| Eawith | 0 | 0.168 | 0 | 36 |
| Mains | . 0018 | 0.111 | 0 | 55 |
| Jupitar | . 0.025 | 00101 | 014 | 023 |
| \$atum | 0.012 | $00 \times 1$ | 10. | 11 |
| Mrains. | . 0.1911 | 0.1078 | 0.51 | 17 |
| Áptune. | . 0.004 | 0.015 | ()31 | $\cup 47$ |

In this table, the slope is measured relative to the invariable plane. On the basis of the above-mentioned argimerts, Stockwe ${ }^{11}$ defined this plane by thefollowing values of the elements

$$
i=1^{\circ} 3 J^{\prime} 19^{\prime \prime} .376, \quad \because=10 i^{\prime} 14^{\prime} 6^{\prime \prime} .00
$$

which are measured relative to the ecliptic and equinox of 1850.0. For all the planets except Venus and Earth, one of the coefficients $M_{\lambda}^{(\mu)}$ exceeds in absolute value the sum of the ab olute values of the other coefficients. Thus, the perihelious of all of these planets have mean motions. It is interesting to note that the mean motions of tho perihelions of Jupiter and Uranus ine equal and the longitudes of these perthelions differ by exactly $180^{\circ}$. nithe perthelion of Jupiter vibrates about its mean position within the lim'l:s $\pm 24^{\circ} 10^{\prime}$, while the perihelion of Uranus vibrates within the linits $\pm 47^{\circ} 33^{\prime}$. Therefore, the distance jetween the perihelions of these planets at their closest point of approach is given by

$$
180^{\prime}-\left(0.110^{\prime}+1733^{\prime}\right)=10817^{\prime}
$$

## Annotation

The solution of equations (55) and (60), the right hand side of

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which are expressed by determinants of the type (54), has always be a considered as a difficult problem. For this reason, Leverrfer (1839) used a special method for solving these equations. Later on, Jaccbi (1845) suggested an alternative method also basod on the properties of determasiants of the type (54). The best of such methods is that sugges ed by Kryiov ${ }^{(1)}$, which makes use of , we specific properties of the secular equat ons. We shall not ron- aue. these methods here, s.ace the convential methods of unfoldig determinants together with the LobacevskijGreffe method for numerical solution of equations leads to the required solution in a sufficiently simple way.

## 105. Secular Perturbations of Small Planets

We consider the case in which pianet $D$ has an infinitesinal mass sich that the influence of this planet on the other planets $\mathrm{P}^{\prime}, \mathrm{P}^{\prime \prime}, \ldots$ $p^{(m)}$ can be neglected. In this case, the systen of equations (4.3) is devided into two systems, one consist:ng of ths equations obtained by putting $\mu=1,2, \ldots, m$ rith do not involve quantities that concern $P$, and the other consists of the following two equations:

$$
\begin{aligned}
& \frac{d h}{d t} \quad, \sum_{( }^{\top}(0, n)-\sum_{i}|0, n| \\
& \frac{d}{d}=-h \sum_{i}(1, n)
\end{aligned}
$$

which define the secular perturbations of the small planet under
(1) A.N. Krylov, On the numerical solution of the equations, by which the frequencies of small vibrations of raterial systems are defined. Transacitons of the Avademy of Science of the USSR, 1931 (Ocis.1ennom resenil uravnenija, fotorym $v$ tenniceskih voprosah opredeljajutsa castoty ualyh kolebanija materialrył, oistem, Jzvestija Akarlemif nauk SSSR, 1.931).
investigation. When the first system which defines the mutual perturbations of the large planets is soived, we substitute the resulting expzessions into equations (56) and (61). This yields a system of two non-homogeneous equations of the type
where $A$ are constant coefficients. The solution of this system is given by the following equations

$$
\begin{align*}
& l=M \cos (U(0, \mu)+\beta)+\therefore \quad \therefore\left(0, \frac{A}{\mu}\right)-s_{1}, \cos \left(s, t+\beta_{2}\right), \tag{16}
\end{align*}
$$

in which in and $\beta$ are arbitrary constant s. Similar results can be obtained for the elements $p$ and $q$.

The nuestion on wnether the sua $\sum(0, \mu)$ can be equal to one, of the quantities $S_{\lambda}$ or net, which would simplify the simplification of the variables of integration has been investigated by Charlier (1).

The constants $N$ and $\beta$ characterize the motion of the small planet better than the variable elements e and $\pi$. For this ceasou, Hirayama called them the proper eccentricities and the proper peril:elion longitudes.

Computing these quantities for a large number of small planets, Hirayama could separate several fainilies of small planets. To each family, he reiatec the planet:s having ciose values for the proper elenents if and and semimajor exes.
(1) C.I. Chariler, Die Mechanik des Hinmels, $1,424$.
(2) K1yotsugu Mirayamit, Families of Astroids, Japanese Journal of Ast ronomy and Cenplysicis, Vol. T, 1923; Vol. V, i928.

## ANALYTICAL HETHODS FOR OBTAINING THE PERTURBATIONS

OF THE COORDINATES
106.

Equa'ions of the Perturbed Mntion in Hansen's Coordinates
Let us consider a fixed heliocentric system of rectangular coordinates, e.g. the ecliptic system of a given epoch, and denote by $x, y$ and $z$ the coordinates of planet $P$,the motion of which is being investigated, and by $x^{\prime}=y^{\prime}$ and $z^{\prime}$ the coordinates of the perturbing planet $P^{\prime}$. Let $m$ and $m^{\prime}$ be the masses of these planets, $r$ and $r^{\prime}$ their radjus-vectors and $\Delta$ the distance between them. The equations of motion of planet $P$ are thus given by (Sec. 3).

$$
\left.\begin{array}{l}
x+R^{2}(1+m) \frac{x}{r}=  \tag{1}\\
\ddot{y}+R^{2}(1+m) \frac{y}{r^{3}}=\frac{\partial R}{\partial y} \\
\ddot{z}+R^{2}(1-m) \frac{z}{r^{i}}=\frac{\partial R}{d z}
\end{array}\right\}
$$

where

$$
\begin{equation*}
R:=k^{2} m^{\prime}\left(\frac{1}{\Delta}-\frac{x x^{\prime}+1-y y^{\prime}+z^{\prime}}{r^{\prime}}+\right. \tag{2}
\end{equation*}
$$

is the perturbation function.
We introduce a new moving rectangular system of axes by means of the following equat ions

$$
\left.\begin{array}{l}
x=a x+-a_{1} y+a_{2} z \\
\gamma-\beta x-1 \cdot \beta_{1} y+\beta_{2} z  \tag{3}\\
\gamma=\gamma s-1-\gamma_{1} y+\gamma_{2} z
\end{array}\right\}
$$

The angular coefficients $\alpha, \alpha_{1}, \ldots \gamma_{2}$ are functions of time. They satisfy the following relations

$$
\begin{array}{ll}
a^{2}+a_{1}^{2}+a_{2}^{2}=1 ; & a \beta+a_{1} \beta_{1}+a_{2} \beta_{2}=0 \\
\beta_{2}^{2}+\beta_{1}^{3}+\beta_{2}^{2}=1 ; & a y+a_{1} \gamma_{1}+z_{2} \gamma_{2}=0  \tag{4}\\
T^{2}+\gamma_{2}^{2}+\gamma_{2}^{2}=1 ; & \beta_{y}+\beta_{1} r_{1}+\beta_{2} y_{2}=0 .
\end{array}
$$

from which it follows that

111

Using these relatlons, we obtain from equations (3) that

$$
\left.\begin{array}{l}
x-a X+1, y+1 Z \\
y=x_{1} X+\frac{1}{y} y \\
z=a_{2} X X, y_{1} Z+\gamma Z
\end{array}\right\}
$$

(i)

The nine angular coefficients are aiready connected by means of six relations. We imply tlat these coefficients sarisfy another auxiliary conditions, namely

$$
\left.\begin{array}{cc}
x \dot{a}+y \dot{x}_{1}+z \dot{z}_{2} & =0  \tag{7}\\
x_{i}^{\prime}+y_{\dot{3}}^{\prime}+z \dot{\beta}_{2} & 0 \\
x \dot{y}+y_{1}+z \dot{\gamma}_{2} & =0
\end{array}\right\}
$$

Because of these conditions, the derivatives of the new moving coordinates are expressed by means of the following formulae

$$
\begin{align*}
& x-z x \\
& \gamma-a_{1} y+\alpha_{2} z  \tag{8}\\
& \dot{\gamma}=\gamma \dot{x}-\beta_{1} \dot{y}+\gamma_{1} \dot{y}+\beta_{2} z
\end{align*}
$$

exactly as if this coordinate system was not moving. By virtue of equations (6), ore of equations (7) is a consequence of the other two.

Therefore, the nine angular coefficients $\alpha, \alpha_{1}^{\prime}, \ldots$ are related only by eight equations. Accordingly, there is an infinite numer of moving coordinate systems which satisfy all the condltions that we have imposed here.

Hansen calied the rectangular coordinates ( $\mathrm{X}, \mathrm{Y}, \mathrm{Z}$ ), tiue derivatives of which satisfy equations (8), the ideal coordinates. As we have already seen, these coordinates are expressed in terms of the fixed coordinates by means of formulae (3) together with the subsidiary conditions (4) and (7).

The choice of the ideal coord mate systen is still not completely defined. We remove this arbitrariness by inplying that these coordinates should satisfy condition

$$
\begin{equation*}
\because x \cdot 1 ; y:-x=0 \text {. } \tag{9}
\end{equation*}
$$

This condition prodices a coordinate system in which $Z$ is always equal to zero. In other words, the plane $\overline{X Y}$ vill always pass by the radius vector of planet $P$. We shall call such a coozdinate systex a $\because$ :ansen cooriinate systern. It is easy to see that the Hansen coordinate system is completely defined by conditions (4), (7) and (9), if we disregard the two arbitrary constants resulting from the integration of equations (7).

Our problem now is to obtain the equations of motion in terms of the Hansen coordinates. For this purpose we introduce a couple of sumidiary conditions. We multiply equations (3) by $\dot{\alpha}, \beta$ and $\dot{\gamma}$ and add. This procedure gives

$$
\begin{aligned}
& 1=(1,1: \dot{B} ; \quad ; \quad 1)
\end{aligned}
$$

Taking into account the relations which can be obtained from equictons
(5) by means of differentiation, we rewrite this equation as

$$
\begin{aligned}
x \dot{x}-y-y \dot{\beta}= & -y\left(2 \dot{x}_{1}+\left\{\dot{x}_{1}+f-\dot{r}_{1}\right)-\right. \\
& -z\left(\dot{x}_{2}+\dot{\dot{B}_{2}}+\dot{y_{2}}\right)
\end{aligned}
$$

The expression found on the right-hand side of this equation disappears. This can be proved by multiplving equation (7) respectively by $\alpha, \theta$ and $\gamma$ and adding the resulting equations, term by term. In this way, we obtain the first group of relations

$$
\left.\begin{array}{l}
X \dot{x}+f-Y_{3}=0  \tag{10}\\
X \dot{x}_{1}+1-Y_{1}=u \\
X \dot{a}_{2}-1-Y_{\beta_{2}}=0
\end{array}\right\}
$$

Another two groups can be obtained in a similar way. Hence, equations (o) will have for the Hansen coordinates the following form:

$$
\begin{equation*}
x=a X+3 Y ; \quad y=-a_{1} X-1-3_{1} Y ; \quad z=a_{2} X-+\beta_{2} Y \tag{11}
\end{equation*}
$$

Differentiating these cquations twice, we obtain

$$
\begin{aligned}
& x=a \dot{X}+\dot{B} \dot{\gamma}-\dot{a} \dot{X}+\dot{i} \dot{y} \\
& y=a_{1} \ddot{x}-f \dot{y}_{1} \dot{y}+\dot{x}_{1} \dot{x}-\dot{\beta}_{1} \dot{y} \\
& \dot{z}=\alpha_{2} \dot{X}\left|-\beta_{1} \dot{y}\right| a \leq X-+\beta_{1} \dot{y} \text {. }
\end{aligned}
$$

Multiplying these equations firstly by $\alpha_{,} \alpha_{1}$ and $\alpha_{2}$ secondly by $\beta_{1}, \beta_{1}$, and $\beta_{2}$ and thirdly by $\gamma_{1}, \gamma_{1}$ and $\gamma_{2}$ and adding after each multiplication, we finally obtain

$$
\begin{align*}
& a \ddot{x}-1 a_{1} y-1-a_{3} \ddot{z}=x  \tag{12}\\
& \beta \ddot{x}+\bar{s}_{1} \dot{y}-\dot{\beta}=1 \\
& \gamma \dot{x}+r_{1} \dot{j}+f r_{2} z \quad\left(i \dot{a}+f \gamma_{1} \dot{a}_{1}+\gamma_{2} z_{2}\right) \dot{x}+ \tag{1;3}
\end{align*}
$$

Substituting expression (11) fato the first of equations (7) yields.

$$
\left(x \dot{x}-1-a_{1} \dot{x}_{1}+x_{2} \dot{x}_{1}\right) \times-1 \cdot\left(\dot{x}_{1} \dot{x}-f-\dot{x}_{1} \dot{x}_{1}+\dot{\beta}_{1} \dot{x}_{2}\right) Y=0,
$$

Taking equation (4) into consideration, we obtain

$$
\dot{a}_{0}+\alpha_{1} \dot{a}_{1}+a_{3} \dot{x}_{2}=0, \quad a_{1}^{j}-1-x_{1} \dot{\beta}_{1}+\alpha_{2} \dot{\beta}_{2}+\dot{x}_{i}^{3}+\dot{\alpha}_{3} \beta_{1}+\dot{\alpha}_{2} \beta_{2}=0
$$

Comparing the last two equations, we obtain

$$
\beta_{1}+\beta_{1} \dot{x}_{1}+\beta_{2} \dot{a}_{2}=0, \quad a_{1} \dot{\beta}+a_{1} \dot{\beta}_{1}+\alpha_{2} \dot{\beta}_{2}=0 .
$$

These equations yield

$$
a_{1}^{1}=x_{2} \quad x_{1} \quad p_{1} a_{1}-y_{1} \quad p_{1} a_{1}
$$

However, since it follows from conditions (4) that

$$
\stackrel{Y}{\beta_{1} x_{2}-y_{2} x_{1}}=\stackrel{\gamma_{1}}{f_{2} x-\beta_{1}:}=\stackrel{\gamma_{2}}{\beta x_{1}-\gamma_{1} x},
$$

we finally obtain

$$
\frac{a}{\gamma}=\frac{a_{1}}{\gamma_{1}}=\frac{a_{1}}{\gamma_{3}} .
$$

We can in a similar way prove the following relations

$$
\frac{\dot{\beta}}{\gamma}=\frac{\beta_{1}}{i_{1}}=\frac{\xi_{2}}{i_{2}} .
$$

Ey taking all these relations into account, we can replace equation (13) by

$$
\gamma \dot{x}+\gamma_{1} \ddot{y}+\gamma_{2} z=1\left(\alpha_{2} \dot{x}+\gamma_{2} \dot{\gamma}\right) .
$$

We shall now consider equations (1). Multiplying them firstly by $\alpha, \alpha_{1}$ and $\alpha_{2}$, secondly by $\beta_{1} \beta_{1}$ and $\beta_{2}$ and thiraly by $\gamma$,
$\gamma_{1}$ and $\gamma_{2}$, and adding after each multiplication, we obtain considering equatione (12) and (14)

$$
\begin{align*}
& x-1-R^{2}(1+m) \frac{x^{\prime}}{y^{\prime}}=\frac{d R}{d X} \\
& \ddot{\gamma}+R^{3}(1+m) \begin{array}{l}
\gamma \\
\gamma
\end{array}  \tag{15}\\
& \dot{a}_{2} \dot{x}+\dot{\beta}_{2} \dot{\gamma}=\quad \begin{array}{l}
d Z \\
d Z
\end{array} \tag{16}
\end{align*}
$$

In fact, it follows froin equations (6) that

$$
\begin{aligned}
& a \frac{\partial R}{\partial x}+a_{3} \frac{\partial R}{\partial y}+a_{3} \frac{\partial R}{\partial z}=\frac{\partial R}{\partial X} \\
& \beta \frac{\partial R}{\partial x}+\beta_{1} \frac{\partial R}{\partial y}+\beta_{2} \frac{\partial R}{\partial z}=\frac{\partial R}{\partial Y} \\
& \gamma \frac{\partial R}{\partial x}+\gamma_{1} \frac{\partial R}{\partial y}+\gamma \frac{\partial R}{\partial z}=\frac{\partial R}{\partial Z}
\end{aligned}
$$

and, moreover,

$$
r^{2}=x^{2}+y^{2}+z^{2}=x^{2}+y^{3}
$$

Equation (16) and the last of equations (10) yield

$$
\left.\begin{array}{l}
\dot{a}_{3}=-h^{-1} y  \tag{17}\\
T_{2} \\
d R \\
\dot{H}_{2} \\
\frac{T_{2}}{T_{2}}=+A^{-1} X^{d K} \\
d Z
\end{array}\right\}
$$

where

$$
\begin{equation*}
n=X^{d \prime} \quad \cdots \gamma^{d i} \tag{18}
\end{equation*}
$$

since,

$$
\because \quad \gamma \quad 11 \quad a_{2}-\frac{6}{i}
$$

then equations (15) and (17) perfectly dezine $X, Y, \alpha$, and $\beta_{2}$. After having done this, we can now obtain $\alpha, \beta, \alpha_{i}$ and $\beta_{1}$ by means of equations (4). In this way, we are able to find all the quantities, whit ch euter expressions (11) defining the coordinales of planet P.
107. Transformation to the Polar Coordinates in the Plane of Osculating

Orbits
Let us first prove that the coordinate plane of the Hansen coordinate system is the plane of the osculating orbits. For this purpose, it is sufficient to prove that the product

$$
Z-\gamma \dot{x}+\gamma_{1} \dot{y}+\gamma_{2} z
$$

is equal to zero. Using relations (9) and (11), this product can be written in thefollowing way

$$
\begin{aligned}
& Z=-x_{i}-y \gamma_{1}-z \gamma_{1}= \\
& =-X\left(a \dot{\gamma}-1 \cdot a_{1} \dot{\eta}_{1}+\dot{\gamma}_{1}-a_{2} \dot{\gamma}_{2}\right) \cdots \\
& -y^{\prime}\left(\beta \gamma+\rho \cdot+-\beta_{2} \gamma_{2}\right) .
\end{aligned}
$$

Substicuting expressions (11) jrto the last of relations (7), we see that this quantity is really equal to zero.

We now introduce the polar coordinates in the plane of osculating orbits by putting

$$
\begin{equation*}
X=r \cos w^{\prime}, \quad \gamma=r \sin w . \tag{19}
\end{equation*}
$$

since,

$$
\begin{aligned}
& \partial R=\begin{array}{l}
\partial R \\
\partial w \\
\partial X \\
\partial \sin w+\frac{\partial R}{\partial Y^{2}} r \cos w \\
\partial R=+ \\
\partial r
\end{array} \frac{\partial R}{\partial X} \cos w+\frac{\partial R}{\partial V^{\prime}} \sin w
\end{aligned}
$$

then, instead of equations (15), we cobtain

$$
\left.\begin{array}{l}
d  \tag{20}\\
d t\left(r^{2}-\frac{d w}{d t}\right)=\frac{\partial R^{2}}{d u} \\
\begin{array}{l}
d: r \\
d t^{2}
\end{array}-r\binom{d w}{d l}^{2}+k^{2}(1+m)=\frac{\partial R}{r^{2}}
\end{array}\right\}
$$

Equation (18) yields

$$
\begin{equation*}
h=r^{2} \frac{d w}{d t} \tag{21}
\end{equation*}
$$

from which it follows that the first of equations (20) may be replaced by

$$
\begin{equation*}
\frac{d h}{d!}=\frac{d h_{!}^{\prime}}{d!!} . \tag{22}
\end{equation*}
$$

108. The case of unperturbed motions.

If the perturbation Eunction $R$ is equal to zero, then equations (20) are reduced into the well-known equations of the two-body protieul. In this case, their general solution w: 11 be given by
which includes the four arbitrary constants $a_{0}, e_{o}, M_{o}$ and $X_{o}$ Equation (7) can in this case be considered separately from equations (20). They lead to constant values for $\alpha_{2}$ and $\beta_{2}$. The remaining angular ccefficients required, $\alpha, \beta, \alpha_{1}$ and $\beta_{1}$ can be deifined by either equations (4) or (5). It is clear that one of these coefficients will remain arbitrary. The final equations of motion of planet $P$ are

$$
\begin{align*}
& x=a r \cos w-1-\beta_{r} \sin w \\
& y==a_{1} r \cos w-\beta_{1} r \sin w \\
& z=a_{2} r \cos w+\beta_{2} r \sin w,
\end{align*}
$$

They involve seven arbitrary constants. $t$ is, however, fasy to see that two of these constants, nanely $x_{o}$ and one of the coefficients $\propto$, $\beta, \alpha_{1}$ and $\beta_{1}$ whiteh remains arbitrary, define one and the same one thing, namely the position of the $X$ axis in the $X Y$ plane. Hence, the value of one of these two coefficients may be fixed.


Fig. 13

Instead of the integration constants $\propto_{2}$ and $\beta_{2}$ defining the position of the cxbital plane, we introduce another pair of more convenient parameters, namely the longiture of the ascending node $\Omega_{0}$ and the $\varepsilon$ lope $i$ (Fig. 13).

The position of the $X$ axis in the ortital plane may be defined by the angle $\sigma_{0}$ between this axis and the ascending node of the $X Y$ plane relative to $x y$ plane. Denoting by $l$ and $b$ the helfocontrir longitude $x Q$ and latitude $Q P$ of $p l a n e t P$, we obtain

$$
\begin{align*}
& x=r \cos b \cos l  \tag{}\\
& y \cdots r \cos b \sin l \\
& z=r \sin b .
\end{align*}
$$

Referring to triangle $\Omega \mathrm{Q}$, we see that since $\Omega \mathrm{P}=\mathrm{xp}-\mathrm{x} \Omega=$ $\omega-\sigma_{0}$, then

$$
\left.\begin{array}{rl}
\cos b \cos \left(l-Q_{0}\right) & \cdots  \tag{26}\\
\cos b \cos \left(I I \prime-o_{0}\right) \\
\sin \left(l-Q_{0}\right) & -\sin \left(m-\sigma_{0}\right) \cos i_{11} \\
\sin b & =\sin \left(I I-a_{11}\right) \sin i_{41}
\end{array}\right\}
$$

from which it follows that

Comparing quations (24) and (25), we obtain


Equations (23), (2.4) and (27) establish the complete solution of the differential equations (1) for the case in which $R=C$. This wolution involves seven arbitrary constants, two of which are equivalent to a single constant. Whe fact lies in that euquations (26) deperd only on

## ORYGTival Page IS <br> OF POOR QUALITY

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$$
w-\tau_{0}=u+\%_{0}-\tau_{0}
$$

i.e., on the difference $x_{0}-\sigma_{0}$ which refresents the distance between the perrihelion and the node. The presence of two constants instead of one in the abore formilae is made use of in Hansen's method for the calculation of perturbations.
1.09. The Laplace-Newcomb Method. The perturbation of the Radius Vector

Laplace vas the first to develop analytical methods for calculating the perturbations of the semimajor axis. He developed metnods which made it possibie to calculate the perturbations of the radius vector $r$, the longitudes $w$ in the osculating orbits and the sines of the longitudes of the planets relative to the plane of the unperturbed orbits. Laplace confined himself to the calculation of the perturbations of only the first order and involving no more than the third powers of the eccentricities and the slopes.

In the Qighties of the last century, Newcomb aimed to develop a theory for the motion ofali the layge planets. The revier 3 d Laplace's method, which consiciered to be the rost practiral method. All the tables computed according to Newcomb are to-day the basis of all of the annuals on astronomy. The construction of these tables is via the computation of the perturbations of the coorcinates. The perturbations, of Jupiter and Sat'rn are computed by Hansen's method, while the perturbations of the other planets are calculated by methods developed by Laplace. This shows that Laplace's methods still do not loose thes.r partical value.

In the following, we give a brief description of laplace's methods taking into account the improvements made by Newcomb in order to simplify the computation of the second- and righer-order perturbations.

Let us reconsider equation (20). Puting

$$
\begin{equation*}
p \ldots \ln r \tag{28}
\end{equation*}
$$

and, noting that

$$
r=\frac{d R}{d r}==\frac{1 / R}{d /}
$$

we san rewtite these equations in the following manner

Multiplying the first of these equations by $2 \frac{d P}{d t}=\frac{2}{r} \frac{d r}{d t}$ and the second by 2. $\frac{d w}{d t}$, adding and integrating, we obtain

$$
\binom{d r}{d l}^{2}+-r^{2}\binom{d u}{d l}^{2}-k_{1}^{2}=2\left(C+f_{1}^{\prime} d^{\prime} R\right),
$$

In order to simplify we have put

$$
\left(\begin{array}{cc|cc}
\partial R & d \rho & d R & d w \\
d, & d t & d w & d t
\end{array}\right) d t=-d^{\prime} R, \quad R^{2}(1: m)=k_{1}^{2},
$$

aca by $C$ we have denoted an arbtrary constant. Adding tinis equation term by term to the first of equations (29), we obtain

We denote by $r_{0}$ the radius vector of the unperturbed motion which satisfies the following condition

$$
\left.\begin{array}{lll}
1 & d^{2}\left(r_{0}^{2}\right) \\
2 & d t^{3}
\end{array}\right) \quad \begin{aligned}
& k_{1}^{3} \\
& r_{0}
\end{aligned}=2 C .
$$

Subtracting this equation tarm by term from equation (30), we obtain
where the integral is defined by the condition that both the right- and left-hand sides are equal to zero when $R=0$.

Newconb considered it more useful not to determine the perturbation of the radius vector but of its logarithn. Hence, puting in enology with equation (28)

$$
\rho_{0}=\ln r_{0}
$$

and introducing the following notation

$$
i_{\beta}=\beta-p_{0}
$$

we obtain

$$
\begin{aligned}
& r^{3}=\exp \left(2 p_{0}+2 i_{p}\right)=r_{u}^{2} \exp \left(2 i i_{p}\right)=- \\
& =r_{u}^{3}\left(1+2 i p+\frac{4}{1.2} i p^{2}+\ldots\right)
\end{aligned}
$$

from which it follows that

$$
\begin{aligned}
& \text {, 1, 1 } 1,{ }^{-13} \left\lvert\, \frac{1}{!}\right., \cdots, \ldots
\end{aligned}
$$

We substitute these expressions into equation (31). We keep in the right-hand side only terms of the first order tugether wfth the second order terms whilest neglecting cerms of the higher orders. We then obtain
which will be the basic equation for calculating the perturbations of the radias vectur. In this way, our tasi. in turned into the integration of a differenclal equation of the folluwng type

$$
\begin{array}{c:cc}
d^{2} q  \tag{33}\\
d i- & k_{1}^{2} \\
r_{u}^{2}
\end{array} a \quad Q
$$

where $k_{1}^{2} r_{o}^{-3}$ and $Q$ are unknown functions of $t$. Indeed, in order to obtain the Efrst-order perturbarions, it is necessary to neglect on the right-hand side of equation (32) the terme which involve $\mathcal{S} \rho^{2}$ and to calrulate the perturbation function $R$ using the unperturbed values of the coordinates of the planets. This wili lead us to the expressior of the quantity $Q$ by a well defined function of time. Similyry, by means of the computation of the second-order perturbations, we find the value of the right-hand side of equation (32) within any required accurany in terms of the already known first order perturbations, and so on.

It remains for is to consider the integration of equation 133 ), which is a linear non-homogeneous second-order differential equatior. Thstead of applying the conventional form of the variation method of artitrary, constants, we proceed in a different manner. in suppose that the two lincari?y independent soiutiors $q_{1}$, and $g_{2}$ of the correspondtrg hemogereous equation

$$
\begin{equation*}
q-1 k_{1}^{\prime \prime} r_{0}{ }^{3} q=0 \tag{34}
\end{equation*}
$$

are known, ro that.

$$
\begin{equation*}
\dot{q}_{1}+k_{1} r_{0}^{3} q_{1}=0, \quad \ddot{q}_{2}+k_{1}^{2} r_{0}^{\prime} q_{2}=0 \tag{35}
\end{equation*}
$$

El iminating $k_{1}^{2} r_{c}^{-3}$ fron these equations, we obiadr

$$
q_{1} 4:-4: 1 / 1=0
$$

from which it follows that

$$
4, \dot{4}_{1} \quad-\quad a_{i n}=\text { runvt. }
$$

On the other hand, eliminating the same quantity from equations (3) by meane of each of equations (33), we obtain:

$$
q_{1} \varphi-\varphi \varphi_{1}=Q \varphi_{1}, \quad q_{2} \dot{\varphi}-4 \varphi_{3}=Q \varphi_{2},
$$

Denoting tie arioitrary constante by $K_{2}$ and $K_{2}$, these equations become

$$
\begin{aligned}
& q_{i} q-q \dot{q}_{2}=-x_{i} \mid \int \|, \downarrow d t
\end{aligned}
$$

Ifultiplying the first equation by $q_{2}$ and the second by ... $q_{1}$ and arding, we obtair

In order to obtain the requared particular solutions $q_{1}$ and $q_{2}$, we note that the orbital coordinates

$$
\xi-a(\cos E-r), \quad r_{1}=a \cos \% \sin t
$$

satisfy condition (34). Consequently, it is possible to write

$$
q_{1} \cdot \cos t--c, \quad q_{2} \cdots \sin t i
$$

These solutions can be expressed in tix fnm of explicit furctions of time âs follc

$$
q_{1}=: \begin{align*}
& 1  \tag{37}\\
& 2 \\
& 2
\end{align*} c_{1} \cos \text { int, } \quad q_{2}=\frac{1}{2} \sum_{1} \text { s, sin i.17. }
$$

where, using the results obtained in section $8 \varepsilon$,

$$
\begin{aligned}
& c_{1}=\frac{1}{3} a^{-}-\frac{6}{15} e^{2+}+\ldots ; \quad c_{3}=\frac{125}{381} a-\frac{1375}{9216} i^{c}+\ldots
\end{aligned}
$$

$$
\begin{aligned}
& s_{t}=\frac{16807}{400800^{6}-\ldots}
\end{aligned}
$$

The particular solutions, which we nave chosen, satisfy the following relation

$$
\varphi_{1} \dot{\varphi}_{2}-q_{2} \varphi_{1}=(1-r \cos l) \frac{d t}{d f}=n
$$

which is inplied by the Kepler equation.
We now aprly formula (36) to the solution of equation (32), the right-hand side of which we denote by $O$. We obtain

$$
\therefore n^{\prime} 4: \int 40^{\prime} d^{\prime} t \quad n^{\prime} / 1 / 0_{1}^{\prime \prime} d t^{\prime}
$$

where $q_{1}$ and $q_{2}$ are defined by equations (37). The cuefficient $K_{1}$ and $K_{2}$ will in this case be equad. to zero liecanse the right-kand side should be the same order relptive to the perturbing nass as the left-hand side. When the expansion of the perturbation function F fy multip es or the mean anomalies is miown, the application of equation (38) to the calculation of the first-as well as the second order perturtations is
quite simple. In evaluating second-order perturbations, the last
two terms of equation (32) should be taken into account.
We nots that it is better to evaluate the secular perturbations of the radius vector separately, by means of the secular perturbations of the elements which for example can be computed by Garss' nethod, and then included in $r_{0}$. In this way, $6 \rho$ will consist of o:ly periodic terms. This w111 simplify the calculation of the second order perturbations.
110. The Lariace-Newcomb Method. Computation of Longitudes. Computation

## of Heliccentric Coordinatec

According to Newcomb, we use the second of equations (20) to find the pertarbed longitudes. We obtain

$$
d u \prime=r:\left\{(<) \int_{d u}^{*} d l\right\}
$$

We put

$$
u^{\prime}==\omega_{0} \mid+i, \omega_{0},
$$

and denote by $w_{0}$ the longitude wich corresponds to the ecliotic motion. Since
and, conseçuert $1 \%$,

$$
C \quad a n c o s p \text {. }
$$

thes

$$
\frac{d i t v}{d t} \cdots r: \int_{d: 4^{4}}^{a} d t| | r=-I_{,}, 1 u^{2} n \operatorname{cu}: \varphi .
$$

"sing the following expansion:

$$
r:-r_{0}^{-2}\left(1-2 i_{i}!2 i\right.
$$

and confining ourselver, to second-order terws, we obtain

$$
\begin{equation*}
r_{0}^{2}{ }_{d t}^{d i, u}=(1-2 i p) \int_{d!}^{0} d R \tag{40}
\end{equation*}
$$

From this, we conclude that once we know the perturbations of the radius vector, we can obtain the perturbed longitudes by means of a quadrature.

Particular attention must be paid to the long-periodic terms of d'R. They yield for both $\delta_{\rho}$ anill $\sigma_{w}$, terms having coefficients involving the sjuares of small jivisors appearing as a result of the double integratioi.. Taplace suggested that these terms mignt be taken into consideration by calculating the elliptic coordinates $r_{0}$ and $w_{0}$ In terms of the mean anomaly, which could be found by

$$
n \quad \int_{0} n_{0} d l: \varepsilon_{v}-\ldots
$$

where the Jong-periodic expressions

$$
h_{i}^{\prime} \because(1, m) \quad \text { fa! }
$$

had isen added to $n_{0}$. We are not going to consider here the development of this approach ${ }^{(1)}$.

After obtaining first-order perturbations of the $2 . \mathrm{ri}^{\prime}$ ius vector and latitude jy means of equations (32) and (40), Newcomb calculated the perturbations of the elements $i$ and $\Omega$ by using convential formulae. This encbles us to perform the final steps in the calculation cf the heliocentric coordinates by the method developed in section 1.00 . Having calculated the first-ozder perturbations for all the three coorcinates,
(1) P.S. Lapalce, Traite de mechanique celeste, 1, 1799, 292.
we can find the secnnd order perturbations in equations (32) and (40) by means of the same formulae. Only the right hand sides are changed.

## 111. The initial form of Taolace's method

We have already pointed out that Newcomb introduced some important modifications in the Laplace method whilst studying the motion of large planets. However, the initial form of this method, as described by Haplace in the volume $I$ of "Mecanique Celeste", is not watl:out
interest. This method is even advantageous when we conf!ne ourselves to first order perturbations. Laplace gave

$$
r \quad r_{0}+i \text { ir, } \quad \text { u } \cdot w, \quad \text { ! i,ur, }
$$

which allows us to write equation (30) and the first of equations (29) in the following manner
where we denote by $G_{2}$ and $H_{2}$, the second- and higher order terms. Excluding quantity $k_{1}^{2} r_{0} \delta r / r_{0}^{3}$ from these equations, Laplace obtained on the basis of equation (39) the following equation for calculating the perturjations of the longitude:
where $\ell_{2}$ denotes the aggregate of the second- and higher-order turms.
Equation (41) is equivalent to equation (32). It allows us to obtain Eirst-order perturbations of the radius vector and subsequent ly to find the second-order perturbations by means of equatior: (43). Equation (43) has the advantage over equation (40) thet it doe- not involve the derivative $\frac{\partial R}{\partial \omega}$. Hence it does not require the caiculations
needed for obtaining this derivative. However this equation becomes less interesting when we wish to evaluate serond-order perturbations. In this case we alsc have to calculate not only $\frac{\partial R}{\partial \omega}$ but the second derivatives of $R$.

Laplace suggester to simplify the integration of equation (41)
in the following way. Since (Sec. 32)
then, equation (41) may be written as
where

$$
n_{1} \quad n_{i}\left(1+\frac{3}{2},: \frac{10}{x} e^{1}, \quad,\right)
$$

The latter equation can easily be integrated by the method of successive approximations.

If we denote by
icosin): il

$$
1 . .1
$$

one of the terms of the right-hand side of this last equation, then the corresponding term in the expression of $r_{0} \delta r$ will be

$$
\begin{equation*}
\left.n_{i}-\cdots: \cos (x), y\right) \tag{111}
\end{equation*}
$$

if $\nu \neq n, \varepsilon r{ }^{2}$

$$
\begin{array}{ll}
d \prime \\
2 n_{1}
\end{array}: 111(n, s: 1)
$$

if $\mathcal{\nu}=n_{1}$. We have already pointed out that in order to avoid secular terms in $\delta r$, we have co calculate $r_{0}$ by means of terms which already include secular perturbations. In this case, the quantities $A$, and $\beta$ involved in the terms (45) behave as variables consequently, we shall have an expression of thetype
in $r_{0} \delta r$, instead of the terms (46).
In calculating the perturbations of the longitude, we will similarly have, instead of the integrals

$$
\begin{aligned}
& \int \operatorname{Acos}(d ; \beta) d t \quad \operatorname{ann}(1 \mid \mathrm{a})
\end{aligned}
$$

the following integrals

$$
\begin{aligned}
& \int A \cos (v t+\beta) d t={ }^{A}{ }_{v} \sin (v t,-\xi) ;
\end{aligned}
$$

$$
\begin{aligned}
& \left.+\left.\cos v\right|_{1} ^{1} \begin{array}{ccc}
d(A \cos 3) & d & d \cdot(A \sin \beta) \\
v & d t^{2}
\end{array} . \quad \right\rvert\, \\
& \int d t \int A \sin (v t+9) d t-\frac{A}{v}=\sin (v t-19)-1 \\
& \left.\left\lvert\, \sin v l^{2} \begin{array}{lll}
2 d(A \sin \xi) & 3 d^{2}(A \cos \beta) \\
v^{3} & d t & v^{\prime} d t
\end{array}\right.\right]
\end{aligned}
$$

## 112. Calculation of the Coordinates of the Rectangular Coordinates

The conventional form of the eguations of motion of the perturbed planet $P$ whose mass is $m$, is the fcllowing
where $k_{1}^{2}=k^{2}(1+n)$ and $R$ is the perturbation function defined by

$$
R=\sum^{\prime} k^{\prime} m^{\prime}\left[\frac{1}{2}-x x^{\prime}+\underset{r^{\prime}}{y^{\prime} y^{\prime}}+z^{\prime} z^{\prime}\right]
$$

where the summation is over all the perturbing planets. Let us replace the masses of the perturbing planets $\mathrm{m}^{\prime}$, $\mathrm{n}^{\prime \prime}$, ... by

$$
m^{\prime}=!\eta_{0}^{\prime}, \quad m^{\prime}=a m_{u}^{\prime}, \ldots
$$

where $\pi_{0}^{\prime}$, $r_{0}^{\prime \prime}, \ldots$ are constants and is a perameter varying fror: 0 to 1 . If we take the initial values of the coordinates $x, y, z, x^{\prime}$, $y^{\prime}, z^{\prime}, x^{\prime \prime}, \ldots$ at the moment $t=t_{0}$ in that field, in which the righthand side of the equations of motion are finite, there then exists such numbers as $\tau$ and $\mu_{0}$, such that in the field

$$
i!\quad \therefore 1 \quad \therefore \quad \because \quad!\quad \text { (4n) }
$$

functions $x(t, \mu), y(t, \mu)$ and $\dot{z}(t, \mu)$ will exist, which satisfy equations (47) and can be expanded in series of the tyre
which converge inside the field (48). The terns $x_{0}(t), y_{0}(t)$, ... are responsible for the unperturbed motions. The terms

$$
i_{n} x \quad 1^{n} i_{n}(t), \quad i_{n} y=-n^{n} r_{n}(t), \quad i_{n} z \cdots \mu^{n} \zeta_{n}(t)
$$

represent thr a-order perturbations.
The order to obtain equations for the determination of first-order equations, we substitute into equations (47) the following expressions

$$
\begin{aligned}
& x=x_{0}-f i_{1} x+i_{2} x+. . \\
& y=y_{0}-1 i_{1} y+-i_{2} y+. . \\
& z=z_{0}+i_{1} z+i_{2} z+. . \\
& r=r_{0}+\cdots+i_{3} r+. .
\end{aligned}
$$

and then equate those terms having a first power with respect to $\mu$. We obtain

$$
\begin{gathered}
x_{0}+i_{1} x+\ldots \\
\left(r_{0}+i_{1} r+\ldots\right)
\end{gathered}=\frac{x_{0}}{r_{0}^{i}}++_{r_{0}}^{i_{1} x} \cdots 3{ }_{3}^{x_{0} i_{1} r}+\ldots,
$$

On the other hand,

$$
\begin{aligned}
& r=\left[\left(x_{0}+i x+1 . . \cdot\right)^{2}+\left(v_{0}+i y+. . .\right)^{2}+\left(z_{0}+i z+f . . .\right)^{2}\right]^{\frac{1}{2}}= \\
& =r_{11} \left\lvert\, \begin{array}{c}
x_{0}{ }^{2} x+y_{0} i_{1} y+z_{0} i_{1} z \\
r_{0}
\end{array}+\ldots\right.,
\end{aligned}
$$

from which we find that

$$
i_{1} r \cdot \frac{1}{r_{0}}\left(x_{0} i_{1} x: y_{0} i_{1} y+z_{0} i_{1} z\right) .
$$

Therefore, we finally obtain
as well as cro similar equations for the other two coordinates.
Similarly, by equating termsinvolving the f:sctor $\mu^{*}$, we obtain the following equations:
for the calculation of the n-order perturbations. Here, we denote by $X_{n}, y_{n}$ and $Z_{n}$ expressions involving $x_{0}, \ldots, \delta_{1} x, \ldots$,

$$
\delta_{n-1} x, \delta_{n-1} y \text { and } \delta_{n-1}
$$

Using the series-expansion of the perturbation function, we can successively obtain the first- and second- order perturbations by means of equations (49).

Enke improved this metiod by suggesting the calculation of the perturbations of the radius vector by means of equations (41) iu parallel. with the calculation of the perturbations of the coordinates, when $\delta_{n} r$ is found, then following equation:
in which $R_{n}$ denotes the aggregate of cerms not higher thar the ( $n-1$ )-th order, enables us to compute the first terms in the right-hand side of equations (49). Subsequently, these equations are reduced to the form (33), the integration of which is done quite easily as we have already seen.

The present method immediately gives the coordinates $x, y$ and $z$ required for the calculation of these ephemeride. In spite of this, the method was not widely recognized since perturbations in rectanguiar coordinates are large and their computation is ted*ous. foreover, the

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perturbations of all the three coordinates obtained hy this method are equally large. On the other hand, other methods based on the application of polar coordinates, thecalculation of relatively small perturbations of the third coordinate is quite simple.

113. Hill's Method.

We have just seen that the calculation of ti.a perturbation of any order of the radius vector and the rectangular coordinates is reduced to the integration of the following equations

When the first of these equations is solved, each of the other three equations is integrated independentiy fror the other two. According to section 109 on the integracion of such equations, we can write

In order to represent these equations in a more convenient fora, we agree to lable by a dash the functions of $t$ fn which $t$ is replaced by Putting

$$
N=q_{2} \varphi_{1} \quad q_{1} \psi_{2}
$$

we obtain

$$
\begin{align*}
& r_{0}^{i r} \\
& i_{n}=/ V \varphi_{r} d t  \tag{:1}\\
& V u_{1} d, \ldots .
\end{align*}
$$

if we again replace $\mathcal{C}$ after each integration by $t$. Taking once more ir accordance to section 109

$$
\begin{aligned}
& q_{1} \quad \cos E-e=\frac{r}{a} \cos u \\
& q_{i}-\sin t \quad=\frac{r}{a} \sin u
\end{aligned}
$$

we then obtain

$$
\begin{gathered}
N \cdots \sin (t-i)-c(\sin E-\sin E)=-r_{a} \operatorname{con} \varphi \sin (i-v) \\
q_{1} \dot{q}_{i}-q_{2} \dot{q}_{1}=11 .
\end{gathered}
$$

Hill :ook as an independent variable, the true anomaly v. Since

$$
d t=\begin{gathered}
r^{2} d u \\
u^{2} \| \cos \dot{9} \cdot
\end{gathered}
$$

formulae (51) finally is reduced to the following form

$$
\begin{aligned}
& \text { ir }=\begin{array}{c}
1 \\
n-a^{\prime} \cos \because
\end{array} \int^{0} O, r_{1} \sin (\bar{v}-v) d v \mid
\end{aligned}
$$

$$
\begin{align*}
& \left.\therefore y \cdot n=\frac{r_{n}}{\cos }-\int_{0}^{0} O_{0}^{3} \sin (\bar{u}-\cdots) \text { d }\right) \tag{52}
\end{align*}
$$

After integration, $\bar{v}$ should be replaced by $v$. System (52) involves one equation more than what is required for the crlculation of the perturbations. For this reasnn, H. 1.1 replaced the second and third equations by a single equation. This was easily done by the iritroduction of polar coordinates. Putting
$x \quad r \cos , \cos \beta, \quad \nu=r \sin , \cos i t . \quad=r \sin i$,
we obtadn

$$
H_{1}=\frac{y}{x}
$$

from which it follows that

On thiather hand, it follows from equations (47) that

$$
x \frac{d y}{d l}-y_{d t}^{d x}=n_{0}+\int\left(x_{d y}^{d R}-y_{d x}^{d R}\right) d t
$$

The calculation of the constant $h_{c}$ is done by considering the case of cmperturbed motions in which

Hence, assuming
we finally obtain

$$
\left(r^{\prime}-:\right)^{d} d t \quad \text { h } \int_{0}^{\prime}!d^{\prime}
$$

In this way, putting $\lambda=\lambda_{0}+\delta \lambda$ and noting that

$$
\begin{aligned}
& \begin{aligned}
(r!r a r & (z: ~ \\
r_{i} & \because!
\end{aligned}
\end{aligned}
$$

we cbtain

Finally ending with a transition to the variable $v$, and obtaining the following equations for thedeternination of the perturbations of the coordinates $r, \lambda$ and $z$ :

$$
\begin{aligned}
& \text { ir }=\frac{1}{r i, p} \int_{0}^{0}(1, r, i n(i n-i) d i
\end{aligned}
$$

0.51

Hill paricularly stressed the fact that these equations were accurate and that the, could be applied for the calculation of the perturbations of any order.

Hill used the pane of the ellintic orbit of the planet under crasideration as the $2 y$ plane. Th: y yelds an accuracy of the thirdorder inclusively

$$
z==i z \quad 1, i 4
$$

Putting
we obtain

$$
\begin{aligned}
& i \text { i. } 1.1 \\
& \text { |1.1.', ', i }
\end{aligned}
$$

Hill introduced the foliowing auxiliary quantity

By neans of this quantity we can reduce the equations obtained above in 1h. Eom:

$$
\begin{align*}
& \text { i.f } \int\left|f \int^{\prime} d u \quad 2 \frac{i r}{r}\right| d n \tag{5}
\end{align*}
$$

It is :mportant to note tiat the application of these formulae require the computation of only three derivatives of theperturhation function, namely

$$
\begin{array}{ccc}
\partial R & \partial R & d R \\
\partial r & d, & d z
\end{array}
$$

Indeed, pẹuition (53) shows that

$$
\vartheta=\frac{d k}{d}
$$

We shall not go further than thfs derfvation of eguatinios (it) anc (5r) which . . astitute the basis of $\mathrm{Hi} 11^{\prime} \mathrm{s}$ method. For detafls on the adp-ica ion of these equations to the computation of ferturbatior, we sufta he readur to Hill's original work. ${ }^{(1) \text {, winere the gueries on the }}$ application of the method to the computation of first- and second-order perturbations, are thorclighly eraritined.

We finalyy point out that ia hill's method the perturbation function
(1) G.W. Hill, A Hechod of Computing absolute Ferturbations.

Astr. Nachr. 83, 1874, 209-22\% = Works, 1, 151-166;
G.W. Hill, Jupiter rerturiations of ceres of the First order and the Derivation of the Mean elements, isistr. Jouraa], 16, 2896, 57-62 = Works 4, 111-122.
is expanded in a series by multiples of the true anomalies $v$ and $v^{\prime}$, $v^{\prime}$ being expressed in terns of $v$. The first operation is easily carried cut if the coefficjents of expansion are to be found by means of numerical methods, which considerably aimplify the calculations when $v$ 's are chosen as independent variables. (1) It is clear that the expressions of the perturbations as fun:tions of $v$, obtained this way, are less convenient than the expressions of the perturbations in terms of explicit functions of time.

## 114. The Main Ideas of Hansen's Yethod

The cholee of the soordiates, in terms of which the perturbations are calculated, is of significant value. We have already seen in the Frevious sectior itat the serturbat lons are more easily calculated in terms of folur coordinates rather than in terms of rectangular ones. One naturally raises the ouestion on whether it is posisible to finc another system of variailes in terms of which the calculation of the perturbations would be eveu ezsfer. This question was investiqated by Hansen.

We have already nointed out in section 11C that iapla:" suggested to include long-pericdic perturbations into the rean anomaly serving in the calculatiors of the radius vectors ard the longitudes. Hansen
(1) In the case of and ?ytical expansion of the perturbation f. ction, the choice of $v$ as an independent variable leads to more complis teu results, than does the application of Leverrier and Newcomb's urethod developes in setions 86-39. On theexpansion in multiples of the true anomaly, we quote $H$. Gylden, Trajte analytique des orbites absolues des huit planetes principales, 1, Stockholm 1.893; G.K. Hill, Development in Terms of the True Anomaly of odd Wegative Powers of the Distance between two Planets Movine in the Same Plane, Vorks, 4, 398-4.0).
developed this idea further and suggestec to tale the mean anomaly as one of the variables in terms of witich ine perturbations would be evaluated.

Let us first consider the metion in ar orbital plane, d-fined in the absence of perturbations by equation (23). According to Hansen we assume that the perturbed values of the orb tal coordiaates $r$ and $w$ are defined by the simi"ar formulae:
where $z$ and $v$ are the corresponding functions of time $t$. The equations which define the unknown functions 2 and $v$ are deduced by substituting expressions of $v$ and $r$, given by formulae (56), intc equations (20). Se ria'l not pive here the method for the deduction of these equatirns.

For the unperturbed motion $z=t$. It is hence natrual to search for $z$ in the form

$$
\because: 1==
$$

where $\delta_{z}$ is a suall quantity $\varepsilon_{-v i}$ g the same order of magnituac as the jerturbing masses. When $\delta z$ and $v$ are obtained, he can caiculate the coordinates wind $I$ that determine the position of the planet in the $X Y$ plane. What remains after this is tc show how the positions of the axes $S X$ and $S Y$ couid be specified at an arbic. ary monent of time.

In consfilerlag the perturbed motion, we have to reflace in formulae (27), the constant elemeate $i_{0}, \Omega_{0}$ and $\sigma_{0}$ ly iniv osenlating elements at the moment $t$ uncer consideration. This yields

$$
2:=-\operatorname{sill} 3 \sin i, \quad \dot{\beta}_{2}=\cos 0 \text { vil. } i, \quad \therefore=\operatorname{co}, i .
$$

Hence, returning to equitions (17) and taking eçuations (19) into consideration, we cbtain

$$
\begin{align*}
& \frac{d i}{d!}=h^{\prime}, \cos (u-z) \frac{d R}{d /} \tag{7,7}
\end{align*}
$$

On the other hand, sunstituting into the following eqiation (Sec. 106):

$$
\therefore d_{d}^{d} \quad 4 . \quad \vdots \quad d .
$$

the valuf 3 of the anguiar ccefficirats defined by equition (27) of section 106 , we oitain

$$
d^{-} \mathrm{c}, 1 . . \therefore \quad(.)
$$

Consequently

Integrating equations (57) ar.d (58), ze obtoin $i, \sigma$ and $\Omega$. Since we have obtainei a scaperflucus irtegration constant, then according to Hansen we require, that the initial values of the element- coreesponding to the moment $t=0$ satisfy the following condition

$$
\because \quad \because
$$

In this case, lic quantity $x_{0}$ will be aothirg else but the longitude of the perthelion.

It is easy to find the heliocentric coordinates $l$ and $b$ when the integration of the equations that define the quantities $\delta_{z}, V, \Omega, i$ and $\sigma$ is already carried out. For tinis purpose, we apply formulae
which are similar to equations (26) corresponding to the unperturbed motion.

Formulae (59) completely solves the probler of ohtaining the perturbed motion. They are only convenient when we are interested in the calculation of a smali number of separate positions for a planet. Nevertheless, the determination of the perturbatic:s by naiyticai methods usually ends by constricting tables for the motion of the planet under consideration which simplifies as much as possible the computation of its coordinates. In this case, it is better to avoid the uce of tables having two entronces and for this reason, Hansen suggested tc replacr iomman (59) by other formulac fourd mare convenient on tabulating. Hansen preved the transformation these formulae to the following Eozm
wheree the first tews on the right hand side could br conveniently tabulated by the arguricnt $w-\Omega$ owile the small quantities $\Gamma, \psi, \psi^{\prime}$ and $s$ could be detemined without a special effort. Vithout going into their proof, the final results are given by
where $\Gamma, P$ and $Q$ are defined by the Eollowing differential equations:


If we confine ourselves to first-order perturbations, then

$$
\text { i } 11, \quad, \quad 1: 1 . . \quad \because 11 .
$$

In this case, the application of formulae (60) requirer the construction of only one tabale with twe entrances giving the yalues of $S$. Due to the smallness of this quantity, the construction of such a tatie is quite simple.

It is usefui to note, that in the differential equations that we have to solve in the application of Sansen's method, the perturbation function appezrs only in the form of the partial derivatives

$$
\partial K . \quad \partial R \quad \partial R
$$

When Fansen gave the final account of his method ${ }^{(1)}$, he clcsely relatad his method to the expansion $r^{-}$the perturbation iunction(or more exactly, the partial derivatives winich we have just mentjonec) b; multiples of the eccentric anomalies, using the eccentric anomaly of the ferturlid pinnet as an independent variable. This methoù howevnr, depends on neither the chojce of the iniepencient: variable nor the way by which the perturbation
(1) P.A. Hansen, Auseinandersetzung einer zwecknassigen isethode zur Berechning der absoluten bcorungen der kleinen Elaneten, Abt. I-II-ITi, Lelpzig 1857-1861.

## function is expanced.

Hanser.'s method has been widely applied to the calculation of the perturbations of small planets. This is explained, on one hand, by the practicality of this wethod which artually redaces the magnitude of the perturbations to a minimum, and on the otion hand, by the fact that Hasen explained lis 'nethod with the fullest of details.
115. The calculation of the Derivatives of Ferturbation Functions wich

## Respect to the Courdinates

In the calculation of pertur?ations in the coordirates, it is necessary to expand the partial deriwative:;

$$
\begin{array}{llll}
\partial K & d R & d! & d R  \tag{6:}\\
d u & d u & d r & d Z
\end{array}
$$

of the perturbation function in serics. We shall now indteate the way by which this expansion could be performed. We have already found that for each of the perturbinu planets

$$
R-k^{2} m\left(\begin{array}{cc}
1 & r i o r l \\
1 & r
\end{array}\right)
$$

where

Differentiating this perturbation function with respect to v, we obtain

Tt in ciear that

$$
\begin{aligned}
& { }_{11}^{\prime \prime}, 1 /
\end{aligned}
$$

On the other hand, the identity
and the well-krown formulae
allows us to write
from which we can casily obtaic

$$
\partial\binom{1}{\Delta}-\begin{array}{cc}
1 & d  \tag{65}\\
a^{2} \cos = & d^{\prime}
\end{array}\binom{r^{2}}{\Delta}-\frac{e \operatorname{sinl}!}{p r}\left|\begin{array}{cccc}
3 & r^{2} & 1 & r^{\prime}\left(r^{\prime 2}-r^{2}\right) \\
2 & د^{2} & r^{2} & د^{3}
\end{array}\right|
$$

Formulae (63), (64) and (65) Iead is to the first of the derivatives (62). We then consider the calculation of the derivarive with respect to the radius vector. Evilently,

Since

$$
2 r r^{\prime} \cos H=r^{2} \cdot \mid r^{\prime}-1
$$

then

$$
\begin{gather*}
\partial R^{\prime}  \tag{66}\\
\partial r
\end{gather*} k^{3} m^{\prime}\left|\begin{array}{ccc}
1 & r^{\prime \prime}-r^{i} & r \cos l l \\
2,1 & 2 د^{3} & r^{\prime}:
\end{array}\right|
$$

The last of the derivatives (62) can be regarded as the component of the Frturbing acceleration aing the nornal to the orbital plane. Therefore, as we have already seen in section 67,

$$
\left.\begin{array}{lll}
\text { W } \\
\text { W } & \text { K } & 1 \\
j & 1 \\
r^{\prime}
\end{array}\right):
$$

where we denote by $\dot{\zeta}$ the $z$-coordinate of the perturbing planet. Evidently,

Consequently

It is thus suffictent for the calculation of the partial derivatives (62) to expand the quantities $\Delta^{-1}$ and $\Delta^{-3}$ in double trigonometric serfes by multiples of the mean anomalies. For all the other quantities involved in formulae (64) $\cdot$ (67), we have already obtained in section 82 in a general form their expansion series.

We have aeveloped a series expansion by multiples of the mean anomalies. The same formulae can be used viner the seires expansion is carried out iay multiples of the eccentric anomaly. We only have to substitute in equation (65)

$$
\left.\begin{array}{cc}
1 & r: \\
d .11 & 1
\end{array}\right) \quad d t: 1 \quad 1 \quad\binom{1}{2}
$$

We rote that it is more useful to expand the quantity $r^{2} \Delta^{-1}$ and not the quantity $\Delta^{-1}$ in $A$ do:hle series, since the former cuantity is the one involved in the final equitions.

PART FOIR

## THEORY OF LUNAR MOTION

## CHAFTER XVII

## PRIICIPLES OF THF THEORY OF LUNAR MOTION

LAPTACE'S THEORY
116. General Properties of the Junar Motion

The position of the moon is always determined relative to the centre of the earth which, in this case, is chosen as the central body, The motion that the moon would inave in the absence of celestial bodies other than earth is considered to be the basic unperturbed motion. The modifications that thesun and the other planets introduce in this motion are called perturbations or inequaif.ties.

The perturbations produced by the sun into the motion of the moon are of farticular interest. These perturbations are çife dirferent in character as compared to those we deal with in the study of the planets' motions. The perturbations produced by all the planets, exceft earth, onto the lunar motion are small due to the smallness of the perturting masses as compared to the mass of the sun, although these planets are ofter much nearer to the moon than to the sun. On the other hand, the sun is considered as a perturbing body in the theory of lunar motion inspite the fact that its mass is 331950 times lerger than the mass of the earth, which is considerer as the central body. The reason for this chotce is that the sun is at a distance almost 400 times larger than the distance from the earth. Taking the mean distances of the moon and the an from the earth tn je 384400 km anc 149450000 kn respectively, we find that the ratio of these distances is equal o 0 1/389. Tt t. en follows that
the acceleration induced by the sun to the moon is on the average

$$
\begin{gathered}
3.11150 \\
389:
\end{gathered}=2.2
$$

times greater than the acceleration caused by earth to the moon. However, we usually study the motion of the moon relative to earth, we are thas finterested in the difference in accelerations caused by the sur on the motions of the moon and the earth. It is easy to see that the perturiing acceieration, occurring thisway, is equal on the average tc

| $\frac{331!3(i)}{354}=\frac{1}{1,7}$ |
| :---: |
|  |  |

of the acceleration induced by earth. Taking the coccentricities of the terrestial and lunar orbits into sonsideration, it is easy to show that the previous ratio can at most reach the value $1 / 80$. Hence, we can couclude that the perturbations produced by the sun on the motion of the moon is by two orders of magnitudes larger than those we ordirarily deal with in the theory of planets.

The aearness of the moon to the earth, on one hand, simplifies the investigation of lunar motion by rendering the perturbatioins produred ty all of the other planets, quite small. On the other han i, due to thls n zarness, one has to take into accuant the influerce of the deviation of the earth's structure from the spherical symmetry upon the lunir notion. Taking these reasons into consideration, we find that the tieory oi lunar noticn fs naturally divided into the following items
(1) The investigation of the motion of the threc material points $T$, $L$ and 3 , one of which $S$ (Sun) moves along a Keplex ellipse around the centre of gravity $G$ of theother two pointe $I$ (Earth) and $L(T)$ This is the tasic probler of the theory of lunia motion.
(2) The calculation of the perturbations, that the deviation of the earth and moon's structure firm a spherical symnetry, causes to the lunar motion.
(3) The calculation of the perturbations produced by the direct attraction of the planets.
(4) The calculation of the percurbations produced by tie deviation of the mction of the sun $S$ around point $G$ according to Kepler's law. 1.e. the perturbations winch indirectly depend on the lateraction of the other planets.
(5) The calculation of thesecond- and higher-order perturbatiors occurring due to the combine: infiuence of the factors indicated in items 2,3 and 4.
(6) The calculation of the contributions of all of the other factors that can influens. the lunar motion (e.g. sea tides, increase in the noon's and the earth's masses due to the accumulation of meteorites; Only the first of these items involves serious difficulcie of the first magnitude. The methode apolied to this example are of general interest and the remaining five prohlems, which sometimes require a great amcuat of work, can always be solved by applying tine method of successive approxjmations in itr convencional form. Taking this fact Into consideration, we sha 11 in thefuture confine ourselves entirely to the consideration of the basic problems, which is sometimes called the solar theory of the Moon's motfon (theorie solaire du mouvement de la Lune).

We note that the sharp division of the theory of lunar motion into the items given above is not alrizys recomended. For example, il nay be useful to take into account part of the perturhations of the sun while solving he basic problen, such as thesecular motion of the perihelion and
the secular decrease of the eccentricity of the terrestrial orbit.
We conclude these introductory remarks by eiting the most inportani perturbations produced by the sun onto the lunar rotion. The unper= turbed orbit of the moon may be takun to be an ellipse with an econtricfty equal to 0.05490 , lying in a plane inclined to an ecliptic of 1850.0 by an angle of $5^{\circ} 9$ '. The perihelion of tine mon's orbjt has a cranslational motion. Tt ferforms a full rotation in 8.8503 years on the average. The influence of the sun consists, firse of all, in adding periodic inequalities to the uniform motion of the perinelion of the moon. The largest of these inequalities has an amplitude of $8^{\circ} 41^{\prime}$. The eccentricity is slightig changed and oscillates around the abovementioned real value. On the other land, the line of nodes mores backwards making a full rotation in 18.5995 years on the average. The most signtficant of the periodic inequalities, which adied to this uniform motion, will have an amplitude of $1^{\circ} 26^{\prime}$. The slope of the orbit will have a periodic inequality as a consequerice of which, it will vary within the 1imits of $4^{\circ} 5,7^{\prime}$ to $5^{\circ} 20^{\prime}$.

Now, considering the perfodic fnequalities of the longituce, the folluwing formula gives en estimate of themst significant quantities:

We denote here by $v$ the true longitude of the mon, by $\lambda$ the mean magnitude, ty $M$ and $M^{\prime}$ the mean anomaties of the moon and the sun respectively, where $: 1$ is onlculated from the mean position of the periielion. and finally, by $D$ the difference between the mean longitudes of the mour and thesun. The terms whose argurnai: are $M, 2 M$. ... are called the elliptic terms. Their sum define: ribe quation of the centre. The term having the argunent $2 \mathrm{D}-\mathrm{M}$ is called the evection. It 1 s easy to see that
the period of this perturbation is equal to 31.3 days. Ther induality produced by the term, the arrment of which is 2 D , is called it variation. The period $c \in$ the variation is evidently equal to one half of a synodic month, i.e. to 14.76 days. The variation does not charge the position of the moon in the syzygies or quadrant i, it produces a 1arge displacement of the moon in the octant:. The term having the argunent $M^{\prime}$ produces a perturbation, the period of which is one ysar. This perturbation is cailed the arrual irequality. It is caused by the ellipticity of the terrestrial orbit, leading to sone changes in the distance to the sun and consequently in the magnitis ie of the perturbing force. Finally, the terms of arguments $\bar{D}$, 3D, ... are re ponsible $f$ r the parallactic inequaliti三s. The amplitude of each of these inezualitins is proportional to the ratic a/a' of the mean distances of the moon and the sun. Since the pacallax of the mooin is sasily obcainex from the results of the ouservations, the comparison between the observed and theoreti, al values of the parallactic ineolalities makns it possible to determine the paralinx of the sun. This is one of the most accurate methods for the determination of the parallax of the sun.

The: are te:ns having similat argumente in the expanstons of the radius vector and latituce of the moon. The series-expanzion of the pairsllax of the mon is easily deduced from the ser ifz-axpansion of che radius vector. It has the folioving form
where thefjrst 1 ine involves tise mean value of the parallar and the t:11irtic terms, while the second line invoives the nost fin ortant serturbations.

At the present tine, each of the above-mentionpd names denotes a group oi terms which are similar to the corresponding principal term given above, For xanple, the following superposition of terms
are called in evection. The superposition of terms having arguments LD, 4D $\quad \therefore, \ldots$, i.e.
are called a variation. The annual inequality is the name of the group of terms that depend on the mean anomaly of the sun, namely

Finally, the parallactic inequality is given by
117. A Brief Historical Survey of the Development of the Lunar Motion

Theory
The modern lunar motion theory began after the discovery of the universal law of gravity, Newton proved that the variation, the motion of the perihelion, the motion of the node and the obsparad changes in the slope and eccentricity can be interpreted within the framework of the universal law of gravity. Newton did not adm to develop a complete theory which could reproduce Lunar motion. Nevertheless, he could determine a mummer of s:rparate inequalities with sufficiently high accuracy. We have the right to think that Newton obtained his results by means of a general method, namely the methods of variation of elements, although he published his results in the for of fragmentary then ems.

The way to construct a theury, capable of descr bing all the particular:tips of the mon: 's motion was tadicated by Kepler. He wes able to express this problen in terms of a differential equation syster and s: rted in li4? to soive this systen by usirg the method of successive approximations. Clero was first to suggest that the first spproximation to be maje on the lunar motion should be tine takinf of an ellipse having a uniformly rotating line of ipars instead of a fixed ellipse as Kepler suggested. $D^{\prime}$ Alembert (1754-1756) developed a method similar to that of Clero's, which was much more systematic. While Clers's theory adopted fror: the very begimiry :umerical values for the parameters, D'Alemtert gave the first example of an aigetraic thecry in wheth the pazameters are allowed co have arbitrary values. The common factor in both Cioro and D'Alembert's works was the choice of the true longitude of the woon as an independent variable.

An alternative arthod, based on the same ideas, was later developsc by laplace, who studied lunar motion for more than thirty years. The results ne cbtajned were fncluced th the ihird volume of his book 'Mecanique Celeste': fublished in 1802 . Apart fron working out a general method for obtainiag til the porturbations produced by the attraction of the sun, Laplace could for the first tine deternine $t$ : i (itequalities produced by the nonsphericity of tine earth and the attraction $c f$ tre other planct-. "ha latter problem is concerned with one of Laplace's most outstanding discoveries, ramely, the finterp:etallon of the secular acceleration of the monn's mear motion (See section 125). Fe wis also able to prove that simblar accelerations depending on the ecular decrease of the eccentifity of the terrestrial crb: took place into the motion of the perihelton and the node. Laplace calculated the lunar perturbations upto the sicond-and

## ORIGNAL PAGR OF POOR QUALIIT

partically, third-order powers of those parameters, by which the serjes expansions vere developed. Later, in 1827, Damoiseau appied Lapiace's metiod for obtaining the numerical values of lie inegralytien to a much higher accuracy. In 183?, Plana reneated the same work algebraically, but his results involved several errors. In 1846, de pontecoulant publtshed a now theory on lunar motion. In analogy with the theories cited above, de Pontecoulant's theory was based on the application of the polar coordinates. The only dfference was that he those time as the independent variable. The corresponsing differential equations (Sec. 7) were given by Laplace. Thesane method was simultaneousiy developed by lubbock, who published his results in 1834. However, this autlor only confined hinself to the calculation of the second-order approxinations.

A new apfroach to the theory of lunar nc:ion was introduced fr 1753 in Euler's book entitled: "Theoria motus lunae prlílens omm, ᄅjus inequalitates". The extensive "Additamentum", ty whici he conciuded his book, actually inciuded a me'hoc for the variation of the elliptic elements. A further, yat very rough development of Fuler's idiss was given by the method of integration or the perturbed motion equations suggested by Delanay ln 18is. Fy this nethod, Delaunay developed a most. perfect analytical theory for solar inequelitfes. After fouly years of work, he succeeded in obtaining gencral cxpressions for all the perturbations in the perturbing forces up to the seventh order inclusively. Delaunay's thenry was reconsidered by Racau and Auduyer. The revised version of this theory servce as a basis for the extensive tables on lunar theory which have been constracted by Radau.

The problcm ralsed b. Luler on the determinatior of the osculating elements was further developed hy Poisson (1835), Pufseaux (1.864) and Vil'ev (1919). Tt shculd, howver, be pointed out that the deternination
of the osculating elements is not very useful for achieving the matio target of lunar theory, namely the construction of tables for the roon's motion.

Fuler sugptst: d another important idea in the book, he published in 1772 under the titie: "Theorfa not.um lunae nowa mílod" pertractataa una cum tabalif astion micis, unde ad quodvis tempus loca lunae expedite comprifari possunt" ${ }^{(1)}$. This idea consists in expanding the unkno:m functions of the lunar coordinates into a serfes of the yous
where " ans e' are the eccentrisities of the luner and solaf orbits, $i$ the slope of the iurar orbjt anc $A,{ }^{\mathrm{P}_{10}, \ldots, G_{f}, \ldots \text { are periodic }}$ functions. Euler obtained systems or difierential euquations for the consequent deterninstion of the cofiticients.

Anongst the numerous interesting improvenents introduced by Euler in the lunar theary, wecord the application of uniformly-rotating rectangular conrdinate systems. This idea did not find any application for a flong time, in contrast to Fuler's other ideas. Duly after more' than a hundred years, in 1377, did Hき11 show in his well know work ${ }^{(2)}$ the advantage of combining this thea with the atoverientioned mechod for the successive calculation of inequalities of different powers of eccentricities
(1) These exists a Russian translation for the most inportaint sections of this book, which has heen made by Academician A.N. KryJcv, and conplerented with soyer mi interesting counents and addenda. This translation constitutes the materfal. of the locok: Lronard Euler, New theory of lunar motion (Jeonarà Ejler, Novaja Teorija dvizenija Luny) Leningrad 1934.
(2) G.W. Hill, On the Part of Lumar Perlgee which is a Function of the Mean Notion of the Sun and Moon, Acta Math., 8, 1.886, 1-35 (Works, 1, 243-270); Researches on Lunar Theory, Amertcan Journal of Math., ㅍ, 1877 (Works, 1, 281-335).
and slopes, which may be considered as the start of modem celestial mechanics. The theory of huns motion developed by Hill in this work, as well as in other subsequent works, was inter developed by $B=\frac{w^{n}}{}{ }^{(2)}$ Lo its final stage. At the proser time this theory is considered to be the best theory available on lunar motion sinpfete that it is mainly based on Euler's very old ideas.

It is worthwhile mentioning that the develcpaert of Euler's ideas in the ahove-mentioned direction was at the same time started by Hill and ifmultanenusly by Adams who studied the behaviour of separate luriar inequalities.

A 31 fLightily different approach to the study of the moon's motion was suggested by Hansen which consisted in applying his method for the study of perturbed nations (Chapter XVI). The work of this author which continued during the period from 1829 tc 1864 led to the constriction ar the tiles of lunar not lon, which wore published in 1857. Until recently, these tables have been considered as one of the most accurate tab? es available.

Th this chapter, we shall give an account of the theory developed by Laplace. This theory gives us a rapid but sufficiently thorough aquaintance with the main features of the lunar motion. The methods developed by Laplace are also quite interesting by themselves for they can be successfully applied to other problems, such as the study of the
(2) W.E. Brown, Investigations on lunar theory, American Journal of Math., 1.7, 1895, 318-358, The Theory of the motion of the moon, etc., Memoirs of the R. Astr. Society 53, 54, 57, 5: (1807-1908).
motion in the syatens of triplet stars, (1)

## 118. Differential equations for the basic problem

Let us adopt tide centre of the earth as the orifin of a rectancular enlipilc coordinate systen. Denoting by $v$ the largitule of the noon, by $S$ the tangent to its latitude and by $u$ the profection of the radius vector on the ecliptic plaue, ve o'stain
from which it follows that

$$
r \quad 11 \quad \therefore
$$

conequently, we can writc the equations of wotion in the following way (ree section 8):
where $h$ is a constant of integration.
Denoting by $T, I$ and $m^{\prime}$ the masses of the earth, moon and sun, and
(1) A detailed bibliograf'y on the theory of the motion of the moon is given in the paper: E.W. Brown, Theorie des Erdmonces, Encyklopedie der Mathem. Wieisunschaften, Bd VI, 2, 1515.
by $x^{\prime}, y^{\prime}, z^{\prime}$ and $r^{\prime}$ the heliocentric coordinates and radtus vector of the $\operatorname{sim}$; we obtain the following expreseion for the force function (Section 3):
where $A$ is the distance between the moon and the sun.
Furthermore, denoting by 1 the angle letween tie ralus vectors $r$ and $r^{\prime}$, we write

$$
\cdots \quad 31+: \quad 1011
$$

In order to obtain the expanston of ti. Eunction V in powers of the ratio $r / I^{\prime}$, we appiy the following weli-inown fomula
in winch, we denote by
the legandre polynomials. Substituting this expansion into equation (4), and Aropping the term $k^{2} \mathrm{~m}^{\prime} / \mathrm{r}^{\prime}$ hhich: does not depend on the lunar coordinates, and consequertily does not affect the partial derivatives which we are intercsted in, we obtain

This series converges rapjaly since the rath $r / r^{\prime}$ is of the order of 1/400. In order to reduce the size of these formulae, we choose the units of time and mass in sucll a way, that

$$
k=\quad 1, \quad \% \quad l . \quad 1 .
$$

Coufining ourselves to only the necessary terms, we finally obtain

Denoting by $u^{\prime}, v^{\prime}$ and $s^{\prime}$ the coordinates of the sun Ir the adopted coordinate system, we wife

Accordingly, neglecting the coordirate $s^{\prime}$,


Substituting these expressions into $U$, and then replacirg the powers of $\cos \left(v-v^{\prime}\right)$ by the cosines of multiples of this arc, we finally obtain
where we have dropped the term $\mathrm{m}^{\prime} / \mathrm{r}$ ' which wili vanish after differentf.ation.

We note that the quantity $s^{\prime}$ is very small because the position of the orbit of the earth changes very sligidiy and very slowiy. Laplace always put $s^{\prime}=0$ and considered that this would not cause any considerable violation to the motion of the moon. In 1848, while Airy was observing the latitude of the moon, he discovered a smali periodic deviation from the theoretical values. Hznsen proved that this deviazjun suld ber ramely explained by the influence of the term involving $s$ ' in the expansion (7). Indeed, this term produces in the parameter s a percurbation equal co



#### Abstract

All of the $t: r m^{\prime} s$ observable influence involving $s^{*}$ is erhausted by this perturbation. Taking this inco consiceration, we shail alrivs set the guantity $s^{\prime}$ equal to zero in our following discuesions.

We note that the first group of terms, causing perturbations in ! he expansion (7), have the multiplying factor $u^{3}$. If we denotc by $n^{\prime}$ the semimajor axis of the terrestrial orbit, then, since $r$ ' sproportional to $A^{\prime}$, the group of terins under considerations will have the multiplying factor


$$
*^{\prime} \quad \begin{gathered}
n! \\
\\
\quad 11!
\end{gathered}
$$

where $n^{\prime}$ is the mean wiction of the sun. Hence, those terms, witich cause tine largest perturbations of the motion of the moon, depend essentially not on $a^{\prime}$ but on $n^{\prime}$. This latter quentity can be very icocurately determined by combining different observations of the sm, separated by sufficiently large intervals of tire. On the other hanc, the terins of the second group that include the multiplying factor $u^{5^{4}}$ will have the following factor after the expansion in powers o. the eccentricity e':

$$
\begin{gathered}
\text { ' }: \\
\text { 1. } \quad m i d
\end{gathered}
$$

The comparison between the theoretical values o.e the perturbations, predicted by these expressions, with the values obtained by the roservations gives is the possibility of cetermining, the parallax $a^{\prime}$ of the sun. For this reason, the correspording perturbation is saljed the parallac:ic Inequality.

The next terme, which we rave dropped in the expansion (7), have a negligible influence on the morn's motion, This ran be judgen by the suallness of the amplitudes of the inequelities produced by these terms.

After these general corsicerations, we start the integratifn of equations (1), (2) and (3), in which $U$ is replaced by expression (7). We shall use the method of successive : approximations.
219.

## The First Approximation

If the perturbations caused by the an were absent, or equivalently, fi. the mass $m^{\prime}$ of the sun could be set equal to zero, then equations (1) and (2) wotld change into the fullowing forms


The second equation gives

$$
\therefore \quad ; \sin (6-1)
$$

where $\gamma$ and $\theta$ are integration constants. It is easy to see that the general solution of the firsi equat lon can be written as

$$
\begin{gathered}
11+\operatorname{siccos}(11 \pi) \\
n=(1+3)
\end{gathered}
$$

where $e$ and $\pi$ are new axbitrary constants.
We are dealing in the present case with a two-hody problem. Rence, we could have obtafned exprossions (9) and (10) stauting with the wellknown elliptical motion formulat. Tet look at figure 14 , which illustrates the heliccentric celestail sphere. Let $x$ NL' be the ecliptic, ard NL the lunar orbit. Denoting by $i$ and $\theta$ the slope and longitude of the node of the the lunar orbit, we ohtain from the trfangle NLL'.
L. 8.1.
or


Comparing with equation (9), we find that
1 15. :ic. 9.4
while the constant $\theta$ involved in equation ( 9 ) is nothing else but the longitude of the rode.

In croler to ohtain the relation between the constants e and with the ellipitcal elements, we rewrite equation (10) in the following mamer

The Eriangle NLL' then gfores

Denoting by w the lungitide in the oroit $x N+N L$, we: cttair
1.

Consequenlily,

On the other haind, we 'ave

$$
\begin{aligned}
& \text { ( } 11 \text { - }- \text {.) }
\end{aligned}
$$

where we denote by $a, e_{0}$ and $\pi_{0}$ the semi-major axis, che ecentricity and the longitude or the perinelion. Comparing these expressions, we find
fron which it follows that

The first two of these equatiors indicate that the quantities and $e$. differ only by a quartity having the urder of magnitude of $\gamma^{2}$. The last equation yields

$$
r^{r}:=11 r . \quad 1 \quad 1 \| 1
$$

when fouttil orear quantities are neglected.
It is useful to note that

$$
\therefore \because!\quad \therefore \quad: 11
$$

The motion of the perihelion and node of the lunar arbit proceeds so rapilly that it is not useful to adopt expre: sions (9) and (10) as first approximations. The perihelion and apogel or the lumar orbit interchange their positions each four and u haif years. Therefore, if we wish to fuvestigate the mot fon of the mon during a long intervai of time, the fixed ellipse will te as bad an approxination to the real orbit as the rircie is.

These arguments led Clers to take, as a first approximation, dr Invariable ellfpse, rotating in its owil plane. Laplace developerl this idea by takfrg, as a first ipproximation, an orbit defined by

where $g$ anc $c$ are constanis sighty differing from unity. since
then the longitude of the node and the peribelion will be equal to $\theta+(1-3) v$ and $\Pi+(1-c) v$ respectigely. With each rotation, they will be changed by ( $1-\mathrm{g}$ ) $360^{\circ}$ and (1-c) $360^{\circ}$ respectively.

Naturaily, expressions (i2) and (13) cannot exactly satisfy equations (12) and (13) if $g$ and $c$ are not equal co unily. Howeve it will be proved furcher on that when expressions (12) anc: (13) are substitutec into equation (8), second-order quantities reiptive to the small quantities $\gamma$, e, $3-g$ and $1-n$ are obtained. Thus, adopting expressions (12) and (13) as a first approrimation, we already take into consideration sone part of the perturbtitions.

In order to find the dependence of the coordiantes on time in the orbit, defined by equat ion: (12) and (13), we consider equation (3) whirh gives fur the case $m^{\prime}=0$.

$$
\because \quad \therefore \quad \therefore \quad d r
$$

Neglecting the fourth powers of the :mal. quantities e and $\gamma$, we write exprescion (13) as follow:
from which it foliows that
or, after integration,

Denoting by $n$ the mean motion of the moon, related to the semi-najor axis by

$$
1 .
$$

wa write equation (1i) in the following manner

$$
\therefore \quad{ }^{\prime}\left(\begin{array}{lllll}
1 & 1 & 1 \\
1 & \vdots & \vdots & r^{\prime} & \ldots
\end{array}\right)
$$

from which we obtain

The coefficient or $v$ in the raiations between $t$ and $v$ just obtained, must be evactiy equal to $n^{-1}$. Therefore, taking equation (15) into consideration, and neglecting, the third order term, we rewrite these relations as followi

$$
m: \quad i \quad \therefore \quad \because \cdots i \quad \pi+:
$$

where $E$ Lis an incegration constant.
We finally note that equation (14) may be written fin th- following mamer


If we substitutc for $h$ its expression given by equation (i5).
120. Galcuiation of the coordinates of the sun

Since we alopt the longitude $v$ of the moon as the indeperdent varlable, we have to express the coordinates of tie sin by explicit functions of $v$. The rotion cf the sun :elative to the centre of gravity of the system earth-moon can be considered as an elliptical motion if we noclect the influence of the cther planets and oniy study, ws we are going to ao row, the solar inaqualities of the motion of the noon. Ori the contrary, the motion of the sun relative to the centre of tre earth sigaificantiy difeers from an elliptical motion. However, the corresponding corrortion can easily be introduced, as it will be shown at the end of this section.

Hence, we shal: assume that the sun moves relasive to the earth acecrdirg to Fepiex's Iaws. Accordingly, we represent the motion of the sun by mears of the formulde dorlved in the preceding section.

Since for the sun $s^{\prime}=0$ and $X=0$, then we obvata f. for iowing equations of rotion

Hore, we kept the coefficient $i^{\prime}$, depending on the secular motion of the periheiion of the terrestrial orbit inside the arounents. By this method, we lave taken a part of the planetary periub bations into account whthout introducing additional complications to tie aiculation. Since $c^{\prime}$ differs vely slighily from unity, we can then set in the soefficients $c^{\prime}=1$. Elininating $t$ from equations (16) and (19), we obeain
where

$$
\begin{array}{ll}
1 \\
\\
n
\end{array}
$$

Since tiiss ratin is a small quantity in order of $1 / 10$, and since third-order terms are dropped from the left-hand side of the last equations we then neglect in the right-lana side, the terms which have the tioltipiying factcrs $\mu \mathrm{e}^{2}$ and $\mu \gamma^{2}$. For this reason we have replaced in the previous equations the quantity $\mu e^{-1}$ ty $\mu \varepsilon$.

Tie equations obtained abuve can be snlved with respect to $v^{\prime}$ by means of the successive-approximations metiod. Writing $\mu v$ instead of $\mu v+\epsilon^{2}-\mu^{2}$, we obtain

Substituting this value for $v^{\prime}$ into equations (18), we obtain, within the taken accuracy,

$$
\|\quad\| \quad \| \quad .1 . .
$$

since,
where .' is not multiplied by on indefinitely increasing quantity and, therefore can be replared by unity.

## Annotation I

We wish to know, how accurately we can eppicximate the motion of the sun relat ive to the centre of gravity $G$ of the system earth-mon by an elliptical motion. We denote by $x_{1}, y$, and $z$, the cocrilnates of the sun in a coordinete system, the axes of wich are parallel to the previous axes. and the origin is in point $C$. We chen ontain (cf. sec. 4).

Therefore
or, denuting by $H_{1}$, the angle between the vectors Gan and GL and putting $r_{1}^{2}=x_{1}^{2}+y_{1}^{2}+z_{1}^{2}$,

Due to the formulae derived in Sec. 4 , the equations of motion of the sun relative to $G$ read
where

$$
\begin{aligned}
& \mu^{\prime} \quad \begin{array}{lllll}
\Gamma & f & \ell & 1 & m \\
m & (l & 1 & l .)
\end{array}
\end{aligned}
$$

The first tern will not contribute to the derivatives and can this be neglected. Using expansion (5), we obtain

In tins way, we can even drop the second lerm of this expansion which will have an upper 1 init equal to

$$
\begin{array}{ccc}
i & 1 & \cdot \\
\cdots i \cdot & 1 & \cdots, 11,11 ;
\end{array}
$$

## Annotation II

We have applied the formulae, derived for the elliptical :ootion to describe the motion of the sun. In urder to be quite objecrive, re have (1) stazt by the equations of motion deduced in section 4 in order to study the motion of the moon. In other words, instead of using the force function l , defined ly equation (4), we have to use the force funclía $\ddot{u}_{L}$, given just above. Takirg into consietration Factor $\frac{m_{0}+m_{1}}{m_{0} n_{1}}=\frac{T+L}{T L}$, which is multiplied by the force function in the lunar theory, we see that function $U$ has to be replaced 'v

Exanding the 1ast two terms in serics, we obtain
where

$$
\cdot \quad 1,1, i^{i} \quad 1 \quad 1
$$

and the term $k^{2} m^{\prime} / r_{1}$, which does not depend on the coordinates $c f$ the monn has been neglected.


#### Abstract

We now compare this exact expression for the force function, with the expression given dy equation (6). This expression will not be very accurate if we make use of the elliptical motion formulae for the determination of the heliocentric coordiantes of the sun. Sfnce the difierence between $\cos \mathrm{H}$ and $\cos \mathrm{H}_{1}$ is not significant, then the transformation from equation (6) to equation (22) can be carried out by multiplying the terms of the former expansion by the followirg correcting factors



which differ slightly from unity. 121. On the integration of the equations of perturbed motion in the second and highar approximations

Let us substitute the expressions of the coordinates of the roon and the sun, which we have found to b? given by equations (12), (if) and (20) in the firsi approxination, into equations (1) and (2) in which the force function $L$ is replaced by expression (7). Ia this ananer, we obtained the following equations for calculating more accurate values of the coordinates $u$ and $s$ :
where the sumation on the right-hand side consists of a finite number of tems, and where $p$ is a constant not necesscrily having lategral values. Integrating equation (23) as well as a similar eçuation for the coordinate $s$, we obtain the values of $u$ and $s$ in the second approximation. Substituling these values Into equations (1) and (2), we obtain equations for finding the

## - 493 - <br> JRIGINAL PAGE IS <br> (IF POOR QUALITY

third approximation, and so on. It is import ant to note that in each of these successive approximations, we have to leal entirely with euqations of the form (23).

If one of the constants $p$ is not equal to unity, then the general solution of equation (23) will be
where $C_{1}$ and $C_{2}$ are arbitrary coustants.
If these existed a term of the type $P \cos (v+K)$ on the right.-hand side of that equation, then the corresponding term in the general solution would be

$$
\therefore / \because 110 \text { 人). }
$$

The presence of such a term in the expansions of $u$ and $s$ must be prohibited, since the coordinates of the moon always remain finite, while a term of the type (25) can adopt indefinitely large values. In order to prevent terms of this rype from appearing, it is necessary to set In each approximation, the constants $C_{1}$ and $C_{2}$ equal to zero in the feneral solution (24).

Formula (24) shows that the soefficients $P$ of the term for which the constant $p$ is close tu unity must be calculated with a nuch higher accuracy than actuali used in calculating the other cocfficfents. Evidently, this is due to the presence of the small divisor $1 \cdots{ }^{2}$ which will lead to a loss in accuracy on integrating these terms. Care shoula also be paid to the computation of terms for vhich the cuefticients $p$ are near zero. These terms vary slowly when we calculate coordinates is and $s$, but can lead to a considerahle loss in decuracy on salculating $t$ using equation (3). Indeed, a term of the type $P \cos (p v+K)$ involved in the

$$
\begin{aligned}
& -494- \\
& \rho^{\prime} \\
& p^{\prime \prime}
\end{aligned}
$$

If this term is involved in theexpression, found in equation (3) behind an integration sign, we then obtain after the second integıation a term of the type

$$
\because \quad \because \cdot H \cdot k i
$$

It is thus clear that the coefficients of these terms must be computed
with a much higher accuracy, especially in the second case. Such terms will be called critical.
122. Equations for the nost periodical incoualities

Let us first of all consider the inequalities, caused by terms having the multiplying factor $u^{3} u^{-2}$. For this furpose, we put
in equations (1) and (2). We then obtain

$$
\begin{aligned}
& \begin{array}{l}
4 \prime s: s \\
d u=
\end{array} \\
& \text { (27) }
\end{aligned}
$$

incre

$$
\begin{aligned}
& \Gamma \because \cdots \begin{array}{c}
3 m^{\prime} u^{\prime 3} \\
2 / n^{2} \\
u^{\prime}
\end{array},\left|1+\cos 2\left(v-u^{\prime}\right)\right|
\end{aligned}
$$

$$
\begin{aligned}
& \text { III' }
\end{aligned}
$$

We denote the values of $u$ and $e$ that correspond to the second approxImation by

$$
u=u_{0}+\dot{\partial} \|, \quad s==s_{0} \mid-\partial s,
$$

where $u_{o}$ ard $s_{o}$ are the quantities cintalned in the first approximation, i.e.

$$
\begin{align*}
& s_{0}= \gamma \sin \left(g^{r} v-1\right) \\
& u_{0}=-a^{1}\left[1+e^{2}+\frac{1}{i} \gamma^{2}+\left(e \mid e^{\prime}\right) \cos (c v-n)-\right. \\
&-\frac{1}{t} \gamma^{2} \cos \left(2 g u-2^{\prime}\right)+\ldots .1 . \tag{28}
\end{align*}
$$

where the quantity a has been defined in section 119 as the semi-mafor axis, related to the observed mean motion by the relation

$$
\begin{equation*}
n: a^{\prime \prime} \quad 1 \tag{29}
\end{equation*}
$$

By the influence of the perturbing action of the sun, the constant part u will be changed and will no longer be equal to the expression involved in equation (28) (cf. Sec. 95). Following Laplace, it is understood that in the future, symbol a would denote a number which would render the constant part of $u$ have the satie expression as $u_{0}$ in each approximation, namely

$$
=11:=i \frac{1}{1} \ldots
$$

This new constant a will no longer be related to $n$ by equation (29), and for this reason equation (15) will no longer hold irue. Thercfore, we lefine the new quantity $a_{1}$ by

$$
\begin{equation*}
n^{2}, u_{1} \quad 1, \text { or } h \quad a_{1}^{1}\left(1 \frac{!}{2} e^{n} \frac{1}{3} ; \quad .\right) \tag{.50}
\end{equation*}
$$

in analogy with equation (15).

Wie now calculate the quantities, dencted by $I$, II, ... . and start by calculating the expression

$$
\begin{array}{cccc}
m^{\prime} & u^{\prime 3} & m a^{\prime} & \left(a^{\prime} u^{\prime}\right) \\
\hdashline h^{\prime} & u^{\prime} & 2 h^{\prime} u^{:} & (u)^{\prime}
\end{array}
$$

Since

$$
n^{\prime}=a^{\prime} \cdot-1: m^{\prime}=m^{\prime}
$$

then, this expression will be equal to

$$
\begin{array}{ccc}
n^{\prime}: a^{3} & \left(a^{\prime} u^{\prime}\right) \\
(a!)^{\prime}
\end{array}
$$

## Furthermore,

$$
\begin{aligned}
& (a)^{3}=1-\frac{3}{4} \gamma^{2}-3 c\left(1-\frac{1}{2} e^{2}-r^{2}\right) \cos \left(c l^{\prime}-\pi\right)+3 e^{2} \cos 2\left(c l^{-}-\pi\right)-1 \\
& \gamma \frac{3}{4} \gamma^{2} \cos 2\left(x^{\prime}(1)-1+\right.
\end{aligned}
$$

We confine ourselves tc second order terms.
However, on the basis of the arg'ants of the previcus section, also keep the third-order terms in the coefficients of the criticai terms. Putting

$$
n^{\prime \prime} u \cdot 4 \vdots
$$

we firally obtain

On the other hand, within the adopted aceuracy,

Therefore, assuming that
we obtain

Adding this expression to equation (31) we ohtain $I$. In order to obtain expression IT, we put

$$
\begin{aligned}
& \therefore \quad a^{\prime}, \quad \because \quad i H^{\prime}
\end{aligned}
$$

This yiel.ds

It is sufficient leere to oniy keep the first of the two written tems since the term $\lambda+c$ in the argument of the :seond cosine differs significantiy from unity.

It remains for us to oltain the temm III and it is sifficient here to adopt that

$$
\begin{aligned}
& \therefore \prime^{\prime}=a^{\prime} \mid 1-k \cos (c u \quad-1 \mid
\end{aligned}
$$

We then Iinally obtain

The first term thequation (26) can be represented in the following form

The calculation of the expressions $I^{\prime}$ II' and III' is carried out in the same manner. For lack of space, we shall not givt this calculation here. After all the necessary substitutions, the final form of equarions (26) anỉ (27) will be

$$
\begin{aligned}
& \text { i } 1 \text { i } 1101: 1
\end{aligned}
$$

$$
\begin{aligned}
& \therefore \because \because \cdot 1!: \because-1 \quad 1 .
\end{aligned}
$$

In order to estimate correctly the urder of magnitude of the coefficients on the righi-hand side of trese equations, it is neeessary to take into account thit $\mu_{1}$ is a quantity of the first order. Trdeed, for the unperturbed motion

$$
\because \quad n \cdot a \cdot\left(\begin{array}{ccc}
i & \vdots \\
\vdots & \vdots & \\
\vdots &
\end{array}\right.
$$

and, $\mathrm{t}^{\prime}$ erefore, the quantity $\mu_{1}$ is of the order of $\frac{1}{13}$. In this manner, the coefficinnts of the crigonometric functions in equatinns (32) and (33) are given with an accuraty up to terns of the second order. The cocfficients of the terms with criticel arguments are found with an even
greater acruracy. Fourth-order quat: $\therefore$ is are kept in the constant term equation (32) cince we shall need : for future discussions.
'2.. The Second Approximation
Wo. shall now integrate tie : जrs that we have derived in the Provinu nection using the menc aderinite coefficients. We first constder eroution (33). Si*,
then, substituting $S: 0_{o} .6$ into this equation, we may take

$$
\text { rs.- } A \sin (111 \quad \text { re: }
$$

where $A$ is an arbitrary coefficient. Substituting tiais velue fnto the equation under investlyation and equating the coefficients of both of the cosines, we obtain two equations for the determination of the coefficients $g$ and $A$, from which it Follows that

In the latter equation, we may take

$$
\begin{array}{llll}
4 & -1 . & 1
\end{array}
$$

hence we will finally obcain as a result of the eecond approximation

$$
\begin{aligned}
& \therefore \quad \frac{3}{8} 19 \sim 110-1!1(1) \quad \text { (11) }
\end{aligned}
$$

Constitering equatior (32), the susatilution of the value for which we bave fiot found into this equation yields a tion-order term. We drop this term beciusn the coefficient of $v$ in fts argunent di-Etes
from unity by a large quanticy. Since,
we can put

In this vay, we obtain for the determination of the constants and $\subseteq$, Introduced above, and for the noefficients $B_{0}, B_{1}$ and $B_{2}$ the following equations:

$$
\begin{aligned}
& { }_{a}^{e}\left(1+\rho^{2}\right)\left(i \cdots, a^{\prime}\right)-3_{1_{1}^{2}}^{2} c\left(\underline{a_{1}}, e^{2}:\left(e^{\prime 2}\right)-11\right.
\end{aligned}
$$

$$
\begin{aligned}
& B_{1}\left|1-(2-2 \mu-c)^{2}\right|-\begin{array}{l}
3 \mu_{1} 5 \vdots 4 \mu_{1} \\
2 d_{1} 1-2
\end{array}
\end{aligned}
$$

We obtain fron the first equation

$$
\frac{1}{a}=\frac{1}{a_{1}} \cdots a_{1}^{3}\left(1+\frac{3}{2}\left(a_{1}\right) \cdot \ldots\right.
$$

from which it can be seen that the quantity a differs from $a_{j}$ by a secondorder quantity. The second equation enabies us to ottatn the value of $c$. It is easy to see that

$$
\begin{equation*}
\text { c } \quad 1 \ldots \frac{3}{4} \mu_{i}^{*}\left(1 \cdots \frac{1}{2}, \frac{1}{2}, a_{i}^{\prime}\right. \tag{30}
\end{equation*}
$$

The remalning squations yleld

$$
\begin{gathered}
b_{0}=\mu_{1}^{\mu_{1}^{2}}, \quad B_{1} \cdots \frac{15}{a_{1}} \mu_{1}^{2} e \\
B_{2}=\cdots a_{1} \\
I_{d}^{2}\left(g^{2}-1\right)+\frac{!}{4} r^{2}\left(\frac{1}{a}-\frac{1}{a_{1}}\right) .
\end{gathered}
$$

or, neglecting the third and higher-order quantities,

$$
B_{0}=\frac{\mu^{2}}{a}, \quad B_{1}=\frac{1 \Gamma!!e}{\sigma}, \quad i_{3} \quad 0 .
$$

Thus, as a result of the second apprerimation, we obtain, up to second-order terms, the ₹ollowing equations for the orbit


The quantity a involved here can te expressed in terms of the nean motion L of the moon by means of equations (36) ar:i (30).

## Annotarion:

For simplicity and clarity we have confined ourselves to the calculation of $u$ up to second order terms only. According the ergurents stated $\ln$ section 121 , we shaj. 1 in the future need sume of the third-crder tears, amely the arguments which have smail coefficients of $v$. The computation of these ierns is, in principle, not difjeult. Fience, we shall only give here the final rosilt which has the following fum
(3sbis)
224. The Relation Between the Longitude ard Tine in the Second Approximation

Let us consider the integration of equation (3), namely

$$
\frac{d t}{d u} \cdot h^{\prime} u:\left|:\left|: 2=\int_{0} u=0 U a \|^{\prime-\frac{1}{2}}\right|^{2}\right.
$$

We must use for $u$ expression ( 38 bis) in order to obtain the second appioximazion. Using the value of $h$, given by equation (30), we obtain

$$
\begin{align*}
& \left.n^{-1} u^{2}=\frac{a^{3}}{\sqrt{a_{1}}} \right\rvert\, 1-2 e \operatorname{cov}(c 11-\pi): \frac{3}{2} \cdot \operatorname{cosi}(2 c u-2 \pi)+ \\
& +\frac{1}{2} y^{2} \cos (2 g \cdot x-20)-2 \mu^{2} \cos 1 v-\frac{15}{1} x^{2} \cos (10-c v \quad \pi)- \\
& -\frac{15}{8}!^{2} \cos \left(10-2(11-1-2-1)-\frac{3}{8}!\gamma^{2} \cos (11-2 g 0+29)+\right. \\
& -1 \cdot 3_{4}=c^{\prime} \cos \left(c^{\prime} \mu t-\pi^{\prime}\right) ; \cdot . . \mid \tag{40}
\end{align*}
$$

We only keep the third-order terms, in whose argument the coefficient of $v .4$ : mint.

Ti: Is useful th note, that the unperioaic part of this expression must be equal Exactly to $a^{2} / \sqrt{a_{1}}$ because it should be reduced to $a^{3 / 2}=$ $a_{1}^{3 / 2}$ in the case of the unperturbed notion. On the other hond, in the derivation of equation (26), we nee' the following equality

Consequently,
because

$$
\begin{array}{ccc}
m a \cdot & n \dot{A}^{\prime} a & \\
a & a & \\
a
\end{array}
$$

However,

$$
\begin{aligned}
& \begin{array}{lll}
a_{1} & -1 & -c^{2} \\
h^{2} & -i & \ldots
\end{array}
\end{aligned}
$$

Noting that $a_{1}$ differs from a by a second-order quantity, and $\mu,{ }^{2}$ from $\mu^{2}$ by a fourth order quantity, and keeping among the third- and fourth-order only the terms which have very small crefficiente of $v$ in their arguments, fe finally obtain

Multiplying this series by the expressions we obtained previously for $i^{-1} i^{-2}$, we finally obtair the follcwing equation, which determines trime as a iunction of lnngitude:

$$
\begin{align*}
& \frac{1}{2} \because \operatorname{con}(2 \mu n-21)-\frac{11}{1} \because \operatorname{covar} \tag{4i}
\end{align*}
$$

It is of great importance to note that the unperiodic part on the rlaht-hand side of this equation is equal to
whore use is made of equation (36) ard where terms higher than the seconc power of $\mu$ arc neglected. Indeed, the multiplication of the terns in serles (40), the arguments of which do not involve $\boldsymbol{\lambda}$, by the terms in series (43'), cannot cesult in unperiodic expressions. The reason is that the argumerits of all the terms in serics (40') involve $\lambda \mathrm{V}$
which camnot disappear as a resrit of the multiplication of the series. On the other land, the terms in series (40), the argumente of whish involve have a multiplying factor $\mathscr{C}$, whereas all the terms in series (40') except the third and fourth terms are multiplied by $\mu^{2}$. We may tence witite

Let us first of all consider the unperiudic part of this expression and postpone the periodic term to section 126. Since the eccencricity e of the terrestrial orbit varies wi=h time, we denote by $e_{0}$ the value of $e^{\prime}$ at some given epoch $t=0$. Separating tine constant and the variable parts, we may write

$$
\left.\frac{d t}{d u}:-a_{1}=\left(1, \because \frac{3}{2} u_{0} e_{0} . .\right), \frac{\because}{2} a_{1} \because e^{\prime}: c_{1}\right) \text { tperiodic tums. }
$$

Integrating and retaining the fact that we have agreed to dencte the coefficients of $v$ by $n^{-1}$, we obtain
where

In this maner, confining ourselves to second powers relative to we obtain

$$
\begin{equation*}
" n_{0} t-1=" \frac{3}{2} n_{i}^{2} \int_{i}^{1}\left(n^{\prime 2}-e_{0}^{2}\right) n_{0} d t \text {-periodicterms } \tag{12}
\end{equation*}
$$

$T^{\prime}$ periocic: inequalities of $e^{\prime}$ only lead to periodic terms. Hence, we shall not consider these fnequalities here bat only confine ourselves to secular perturbations. As a consequence of these secular perturbations,
eccentricity $c$ ' is found to decrease at the present time. During the next few centrueis, the rule ty which e' decreases can be well approximated by the foliowing relation

$$
e^{\prime}=-f_{1}^{\prime}-x T-x^{\prime} T
$$

where

$$
e_{0}^{\prime}-0.0167 \% 101, x=0 .(1 ;) 01180, x^{\prime}=0.000100126,
$$

and $T$ is time measured in Julian centuries starting from the central border midday 0 January 1900. Substituting this value for $e^{\prime}$ into equation (42) and multipiying the terms by $\alpha^{2}$ and $\alpha^{\prime}$, we finally obtain the following expression for the longitude of the moon

$$
י=n, 1 ;=1 \rho T:+ \text { periodic terms. }
$$

where
' 'H1. : : : ! H1, !
and $n$ and $n^{\prime}$ are the annuai mean motion of the moon and the sun. Sice,
then

$$
=11 . ;
$$

Thus: the mean annual motion of the moon inoreases each century by $2 \sigma \approx$ 20". The coefficient $\sigma$ is called the secular acceleration of the mean motion: of the moon. At the presmul time, the most accurate valle of may ie considered to be the following
which has been given by Brown (Sec. 117). ORIGINAL $P_{A G E}$ IS
OF' ${ }^{\text {OOOR QUGLITY }}$ Annoticion

The second approximation developed here enables us to establish the preseuse of secular accelerations in the mocion of the perinelico and the node. As a metter of fact, equatior. (38) indfeate:s that the instantaneous longitude of the perihelion is equal to

$$
I=0-(c 11-\pi)-\pi+(1-a) 1,
$$

where $c$ is given by equation (37). Hence, the instantaneous speed of the motion of the perifelion is

$$
d \| \quad \frac{3}{1} n_{1}^{2}\left(1-\frac{1}{2} e^{2}+\frac{3}{2} r_{1}-\right)-i \frac{9}{3} n_{1}^{\prime}\left(r^{2}-r_{0}^{2}\right) .
$$

Integrating, we obta: i the longitude of the perihelion at any artitrary moment

In this way, we will obtain, after expressing the longitude in terms of time, a tern which is proportional to the square of ine A similar discussion can he applied to equations (39) and (35) to show that the longizude of the node is expressed by

According to Brown's calculations, the secular increments of the perihelion and the node's metions are respectively equal to

## 125. Secular acceleration ot the mean rotion of the mion

Let us discuss fn detail the question on the secular acceleration of the mean motion of the moon. The volution of this ecuation in one of the $m$ most interesting chanters in the history of eelestial mechanics.

The secular perturbation was discovered by Halley in 1963. Ee made an attempt to determine the mean motion of the moon, i.e. the quantity $n$, by combiring the observations of the darkenirgs, taken in different ages. He used the results of the observations made in Almageste and of those made by Arab astrononers as well as the results of more modern observations. Eaving determined the longltudes $v_{1}, v_{2}$ and $v_{3}$ of the moon in three different epoches $t_{1}, t_{2}$ and $t_{3}\left(t_{1}<t_{2}<t_{3}\right)$, Halley was able to write the following relations:
*here $\rho_{1}, \rho_{2}$ and $\rho_{3}$ are the sums of periodic terms. Solving these equations, we can obtain the two following vaiues:
which are expected to be equal within the adopted accuracy. However, Halley proved that the second value was definttely larger than the firsi one. This would suggest that the nean motion of the moon increased with tine.

Replacing equations (44) by

$$
\left.\because, \cdots t, \ldots,=\binom{t}{\hdashline M i} \text { i }, \quad \quad u=1,2,3\right)
$$

we are able co obtain from these equations, not only the values of $\in$ and $n$, but also the value: of the acceleration $\sigma$. The first reliable nui iral determination of $\sigma$ was performed considerably later, because of the difficulty in using the ancient obervations. In 1742, Dunthorne obtained $\sigma=10^{\prime \prime}$. Tobias Mayer adopted the val.ne $\sigma=6 . " 7$ in the first edition of his tables on lunar mution (1752). Fin the second edition (1770), he torix the value $\sigma=9.0$. Finally, Tandan (1757) gave the value $\sigma=9 " .886$.

Even at the present time, the accurate determination of the value of $\sigma$ derived from the results of these observations is silll regarded as a difficult problem. Hansen treated this problem several times and obtained the following values successively: $11^{\prime \prime} .93,11^{\prime \prime} .47,12^{\prime \prime} .18$ and $12^{\prime \prime} .56$. The values $\sigma=8^{\prime \prime}$ obtained by Newcomb in 1909 and $\sigma=10^{\prime \prime} .3$ found by Fotheringham in 1915 may be considered as the best.

The theoretical interpretation of the secular acceleration of the moon was considered as one of the most intellectual problems of the eighteenth century. Tremendeous cosmological investigations were devoted to this problem. The reason for this strong interest was the following: In the presence of an acceleration in the mean motion of the noon, the distance between earth and moon decreases. This means that whatever the decrease in rate (approximately 3 cm per year) the monn would eventually collapse on the earth.

After a series of untuccessful attempts to find the origin of the secular acceleration of the monn, Itagrange became persuaded by the dea that this acc:cleration was not real and that it probably appeared as a consequence of using whong infcrmation on darkenings occurying in ancient times. On the other hand, Laplace unsuccessfully tried to explain the secular acceleration by introducing a hypothesis on the finice velocity of propagation of gravity. The correct solution was found by Largrange in 1783. He was the first to raise the question on the influence of the secular inequalities of the eccentricity and slope of one planet on the longltude of another. Being convirced that this influence was negligible in the case of Jupiter and Saturn, he made a hasty conclusion that this should be the same for the ofler cases. Later on, whilst studying the theory of Jupiter'a sistwlliteß, Laplace discovered that the secular increments of the eccentricity of Jupiter's orbit producec accelerations
in the mean motion of these satellites. He hurriedly transferred this idea to the lunar theory and in thisway, he finally uncovered the secret of the secular acceleration of the moon's motion in 1787. At the same time, he discovered that the secular increment of theeccentricity of the terrestrial orbit also produced a secular perturbation in the motion of the perihelion and the node. By this way he gave a new and brilifant proof to the character of the unfrersal law of gravity. At the same time, he was able to give a new guarantee for the stability of the solar system. The fact remains that the theory of secular perturbations froves that the eccent-icity of terrestrial orbits varies perindicaily. This Induces the incrament of the mean motion of the noon to alsn vary periodically. At the present time, the eccentricity of the moor is decreasing and hence its mean mution is accelerated. This situation is expected to continue for abcut 24000 years, after which, the eccentricity will start to ircrease and consequently the mean motion of the noon hill start to decelerate.

Laplace pointed out in his nowfamous "iccount. on the system of the World" that it cocid be proved without any calculations, by using simple geometrical comsiderations, that the decrease in the eccentricity of the terrestrial orbit would produce an acceleracion in the lumar motion. He commented on this outcone by stating that one should wonder why this stuple interpretation always essaped geometers only "if it was not slear that the simplest ideas were almost alvays the last fo reach the heads of people".

For the investigation of the secular increments of the anoger ami the node, Laplace introduced further approximations which took into accuunt the second-order pertur?ations. He obtained results which were quite different from those obtained by using the first apprcximatorn.

The first approximation, however, led Laplace to the value $\sigma=10^{\prime \prime} .18$ for the secular acceleration of the mean motion of the moon. This valie was in good agreement with the findings of the observations. Laplace was decefved by the corroboration of results between theory and ohservation. He concluded without further investigation that the first approximation could lead to a resuit of sufficient accuracy. As it was shown later on, this c aclusfon was incorrect.

The first attempt towards improving the theoretical value of obtained by Laplace was almost simaltaneously made by Plana ard Damoiseau. They obtained the values $\sigma=i 0^{\prime \prime} .58$ and $\sigma=10^{\prime \prime} .72$ respectively which were essentially in agrement with Laplace's findings. Yowever, in 385, Adans showed that these authors had made a basic mistake. They fixed the value of $e^{\prime}$ in the integration of the differential equation and replaced the fixed value of $e^{\prime}$ in the resuiting solution by its expression as a function of time. This procedure could be only made in the first approximation. In order to integrate correctly the differential equations in the second approximation, Adms suggested that the coefficient $\frac{3}{2} \mu^{2}$ in the expression
of the acceleration, given by equation (42), should be replaced by

This result was supported by Delaunay ${ }^{(1)}$, who deduced a general expression for the secular acceleration. Delamay's method was essentially Enpromed by N-wcont and 3row, who found that the acceleration would bis equal to

[^2]This result may be considered as final．En error $\pm 0^{\prime \prime} .02$ is obtained by taking into account the ffect of the inaccuracies in the deffinition of all the quantities on which $\sigma$ depends．

Thus at the present time，there ls a discrepancy between the theoretical value 6＂． 0 for the acceleration and the value 8 ＂． 0 obtained froll the anaiysis of the observations．The origin of this diserepancy is not yet clear．All attemots to eliminate this difference by improving the theoretical value of have nor been successful．The most widely accepted suggestion is that the difference Eatwsen the theoretical and the observad vailuts is du：to tie fecete alion of thr earth，caused bv tidal friction．The shortening of the days required to eliminate this effect is too small to have an observable effect on the motion of etner luminaries at the present time．Jeffreys however，estimated theoretically the influence of the tidal friction and found that it could accomn for an acceleration of the order of $2^{\prime \prime}$ in the mean motion of the monn．

## 126．Periodic Inequalities of the Tongitude

Let us once more consider equation（4．1）ard this time concentrate our attention on the periodic terms．Taking into acrome equation（42），we obtain the fcllowing relation

$$
\begin{aligned}
& \frac{1}{f}: 1112(24-24) \frac{11}{5}-11111 \ldots \\
& 1.1
\end{aligned}
$$

We remind the reader that we have agreed in section 120 to replace tie quantity $\mu v+E \cdots \neq \ln$ ine arguments by $\mu$ e．Hence，by putting $\epsilon-\mu \epsilon=\beta$ and coosing the starting point for counting the time
so that $\mathcal{E}=0$, we finally obtain

- $\epsilon=0$ winally obraln
where
or

$$
\text { n/ } \quad n_{1, t} t-1-1:
$$

$$
a \quad \pi \quad ; 11^{\prime}: 1
$$

We solve this equation with respect to $v$. For thi: purpose, we construct the fonlowing successive approximations:

$$
\begin{aligned}
& \text { : :!' • ! !! い }
\end{aligned}
$$

$$
\begin{aligned}
& \text { - }
\end{aligned}
$$

from which, taking into consiberation that $\varphi_{\mathrm{r}}=\mathrm{n}^{\prime}$ and $\lambda_{\mathrm{n}}=2 \mathrm{n}-2 \mathrm{n}^{\prime}$, we obtain

Thits ts the final expression for the longitude of the moon up to secoveorder terms inclusively. We now constder this expression in detail. Let

$$
\begin{array}{cc:c} 
& 11 & =n t \\
11 & - & n-c) n t .
\end{array}
$$

$$
\text { - } 513 \text { - OFIGINAL PAGE I } \quad \text { OOOR QUALIT: }
$$

where $M$ is the mean anomaly of the moon, crunted from a prihelion of longitude $\Pi$. Tre first two periodic terms of expression (45) give us the leading terms of the series-cxpansion of the equation of the centre (section 82), while the expression
is nothing else but the longitude in the orbit. These two terms decine the elljptic inequalities of the motion of the morn. The next term

$$
-\frac{1}{T} r^{2} \sin (2 x n-20)-\frac{1}{1} \because \sin 2(t 6-2)
$$

where

$$
1+-4 \cdot(1-g) n
$$

gives the reduction to theecliptic (scetion 35) calculated to within the second powers of the smali quantities $X$ and $e$. The term

$$
\begin{array}{lll}
11 \\
x & \because & \| 11
\end{array} \because(n \quad r) \prime \quad \because \ddots
$$

has a period of $\frac{360}{\dot{E}\left(n-n^{1}\right)}$, which is equal co haif a synodic month (11.765 days) and gives the variation. The term
gives the erection. The period of this inequality equals to

$$
\begin{gathered}
., 11 \\
\because \because
\end{gathered}
$$

which is equal to one astral day (?7.3166 days), divided by

$$
\because \because!!\quad-1 \quad 1-\ddot{\square}
$$

1.e. approximately 32 days. Finally, the last 1 erm in equation (45) \& Lives the dunual inequality.

If we had kept une more term, ;amoly

$$
\pi a^{\prime \prime} \mid 3(1 \quad \text { is }) \cos \left(v \quad v^{\prime}\left|f \bar{i} \operatorname{con} 3\left(t-t^{\prime}\right)\right|\right.
$$

in the force Eunction t while expanding equation (7) fn section 322 , we would have obtained one more inequality in the longitude. The main part of this inoquality would have been

Such in equation is called a parallactic inequality As we heve already pointed out, this inequality enabies us to determine the ratio $-\frac{a}{q^{\prime}}$ Erom the observation.
127. Expression of the Radus Vector 3 nd the Latitude as Functions of

Time
In , reder to express the radjus vector by is function of +ime, i- is neeressry to substitute the expression of the bngitude, given hy squation (45), into eutation (38). Since
we then obtair, within the adopted accuracy,

$$
\begin{aligned}
& \text { alt } 1: \frac{1}{-i}: \cos (c m t \quad \pi) \mid a^{2} \cos 2(c n t-\pi)-
\end{aligned}
$$



```
Similarly, noting that
```



We obtair：from equition（3？）

$$
\begin{aligned}
& \text { : }
\end{aligned}
$$

In further approximations，the quant fite：r and g，which aly depend on will be given by ${ }^{(\mathrm{L})}$

In conclusion，we derive an expression for the equaturial horizontal parallax of the moon for the monenc t．Denoting this parallax by $p$ and the equatoriad radius of the earth by $A$ ，we finc

Then，vithin the adoptec accuraご，
where
ic A：：
is the parallax that cor，esponds to the mean distance be＇veen $t$ earth and the moon．

1之8．Further Developnent of Laplacn＇s Theory
In the previcus sections，we have compirted the ealcuiations for the second－order inecualities by using＿Laplace＇s method．By this same method
（1）The simplest way to calculate the coefficlents of these serfes is given by Hill＇s thecry（of：： rec tions $1.40,142$ ）．

Laplaw could obtain all the third~order inequalities. Tle most extencive application of this method was given ty Darciseau (1), who had the aim of finding the coefficients of the inequalities to within $0^{\prime \prime} . i$ by asing the method of indefinite coefficients.

Damofseau put

$$
\begin{aligned}
& \therefore \quad \text { Nifici i : } \quad \therefore \text { : }
\end{aligned}
$$

where

wherc the indices $\alpha, \alpha^{\prime}, \beta, \lambda$ and $\mathcal{V}$, on wisich the coefficients A, $B$ and C iopend, run over all positive and negative integral values. By $u_{o}$, the velut of $\because$ wich corresponcs to the Elliptic motion is denct hed. In contrast to what we have presented in section ?23, ie include in the


Damoigeau considered that it mas necessary to keep in the expressions given ibuve 85 coefficients of $A, 37$ coefficients of $P$ and 85 cocfficfonts o. ©. Expressing the soordinates $v$ " and ${ }^{\prime}$ ' of the sun in terms of the coeflicients 0 , he substituted :hese expressions intc the differential eqיarfons. He then obtained 209 equatinns for detining the coefficiuntis

(1) M.C.T. Namnisear, Memoire sur la theorie de la ture, Memolros pres par divers savarts, Faris, 3-e ser. $1,182 \%$, 315-598; Tables de la Lune, form par la semle theorie de l'attractinn et suivant la diverifon de $1 a$ ircomference en 360 degres, paris 1828.

The two equations that determire the coefficients $c$ and $g$ could be solved separatiely. Danoiseau solved the remaining equations by replacing primarily the symbois $c, 8, \mu, e, \epsilon^{\prime}, \dot{D}$ ard $\frac{a}{a^{\prime}}$ by their corresponding values. This sclution was oftained by means of tide method of successive approxinations fin a quite simple manrer inspite of the large number $\mathfrak{f}$ fouations. In conclusion, Damoiseau solvec: equation (48) relative tc $v$ and obtained the expression of the longitude in terms of an naplicit function of time.

We now compare the results obtained by Damoiseau with Hansen's theorctical predictions, the development of which requires a tremendous amotnt of voric. This comparison can be illustrated by weans af the following table which gives the number of coefficients in the expansion $c^{\prime \prime}$ the longitudus and the $l$ imits within which the differences betwern the values obtafned by Damoiseau and the corresponr.ing values obtained by ilansen vary:


The eight largest differences are $\left.3^{\prime \prime} .33,3^{\prime \prime} .15,1^{\prime \prime} .82,1^{\prime \prime} .64,\right]^{\prime \prime} .25$, 1".22, 1". 21. and $\mathbf{j " ~}^{\prime \prime} .20$. These differences niaturally depend on the use of different values for the constants. The posstbiltty of lhese being some computational errors is not exrludaz. Finally, we have to note, that Hansen's calc'ilations led to some vaiues for the enefficierts whinh involvad errors of́ about tens of seconds.

We have thus seen that Laplace's method enaties us to obtatn the numerical values for the perturbations of the lunar motion in a relatively simple ray. Howevcr, this method suffers from some essential drawbacks,
which practically prevent us fron applyjng this methed to the construction of a complete iunar motion which can satisfy the more resent requirements of accuracy. Among these drawbacks are the following:

1- In order to obtain the value of the longitude within a given accuracy, it is necessary to calculace u with a considerably high accuracy. For example, in order to obiain second-order terms in the expansion of $v$, we had to calculate a part of the third-order terns in the expansion of $u$ (sre section 123). In calculating higher crder larms of $v$, the situation will be much worse.

2- The solition of the system of equations, by which the unknown constants are obtained, rapidly becomes more difficult when the number of the unknowns increases.

3- The addition of new terms is quite afficult. At the same fime as Damoiseau, who developed a numerical improvement to Laplace's method Plana and Carlini improved this method by means of an analytical approach. Plane reported his results in three large viluaes winere he oblained all the coefficients in theform of a power series of $\mu, e, e^{\prime}, \gamma$ and $a / a^{\prime}$ up to the fifth order inclusively. In some particular cases, twe developed the expansion up to eight order terns. Inspite of this, he falled to obtain the accuraty ${ }^{2}$ tetained by Damoiseau in his numerical theory wnich required consilerably less efforl. j"ie reason for this sems to be the slon convergence of the series developed in powers of $\mu$. Laflace recognised this situation andfrom the very beginning preferred to construct the theory in a semf-numerical and a seni-algebraical manner. He imneafatily sulbsitituted $\mathcal{H}$ by its numerfcal value and at the same time: koft the other paramoters c. e', $\gamma$ and $a / a$ ' in their symolife forms.

We pointed out in section 1.19 that Laplace used as a first approximation, an orbit obtained from the elliptic orbit by replacing $v$ by cy in the
expression of the latitude. Instead of this arbitrary way of otaining the intermediary orbit of the moor, Guilden in 1885 suggested to keep, in the first approximation, some of the perturbing tems of the differential equations (1) and (2). In this manaer, ic was possible to obtain a more accurate intermediary orbit than that obtained by Laplace. This idea was further developed by Tisserand and Andoiyer ${ }^{(1)}$.

Hill ${ }^{(2)}$ suggested to separate the tern proportional to the square of the radius vector and that proportional to the square of the distance from the moon to the ecliptical plane in the expression of the pertirbation function which gave cine perturbation cause? by the sun. In other words, he suggested to put
where $R^{\prime}$ denotes the romaining pert of the perturbation Eunction. The first of the separated terms is proport ional ir $r$, and hence rasponsible for the rotation of the line of apses. The second tern produce:s the force, which is responsible for the mction ,f the node. In the first approximation, we neglect $R^{\prime}$ and express the other terms in terms of 1 and $s$, so that

$$
U^{\prime} u(1 \quad \vdots s) \quad:!\frac{1}{2}, a \quad{ }^{\prime} \because \frac{1}{!}(a \mid m s u:
$$

substituting this expression into equations (1) and (2), we obtain
(1.) F. Tisseranc, Traite de Me: nníque ceteste, 3, 1i8-140.
(?) G.W. Hill, On Internediary nrbits fin the Luner Theory, Astr. Journal, 18, 1897, 81-87 (incke, 色, 1907, 136-149).
where

The first of these equation defines the anknown intermediary orbit, it can be integrated in a closed form by means of tine elifpticai functions if $s=0$. Subsequently, the complete solution of the system can be easily purformed by means of successive approximations, due to the smallness of $s$. $^{(1)}$
original page . OF POOR QUAU
(1) The following works involve an application of the intermediary orhits in the theory of 1 unar motion:
A.M. Zdanov, The thecry of intermediary orbits and its development for the purpose of invertigalifig the lumar motion (Teorifa Promezutocnvh orbit u prilozente $\epsilon e \mathrm{k}$ jesledovaniju dvizenija Luny) 1882; A.V. Krasnov, Thecry of sclar inectalities in the lunar motion (Teorija solnecnyh neravenstv v dvizenii Luny) Kazan' 1894; A.W. Krassnow, Zur Theorie der intermediaren Bahren des Mondes, Astr. Machr. 146, 1898.

## CHAPTER XVIII

## THEORY OF LLLNAR MOTTON BASIS OF HILL'S METEOD

## 129. Introduction

Hi11 and Adans developed a method for obtaining the main inequalities of the lunar motion with arbitrarily high accuracy in a simple maner. As we have already pointed out in section 117 , this method is mainly based on the old ideas cieveloped by Euler. It enabled Brown to construct one of the most complete theories of lunar motion. The characteristic features of Hjll's work is that he makes use of the rectangular rather than the polar coordinates. Hill pointed out that when rectanguliar coordinates are used, the differcntial equations of motion involve pure alyebrafe functions, while if the longitudes and latitudes are used, trisonometric functions will appear in these differential equatione. In addition, in the case of unperturbed elliptical motion, the rectangtiar coordinates can easily be expressed in terms of explicit functions of time (see section 85), while the corresponding expressions of the true ancmaly, given in section 82 , and consequently those of the iongitudes, are incomparably more conplicated. One has the right to believe that also in the case of perturbed motion, the explicit expressions of the rectangular coordinates will be much simpler than the corresponding expressions of the polar coordinates.

Comparing the integration method of the differentisl equations of motion in rectangular coordinates with the method of variation of elements and with the methor suggested by Dalaunay, Hill agai: showes the advantage of the former method. In order to see this, 1 t.t us assume that we wish to calculate the verturbations to a very high degre of accuracy. Then, we have to use the method of indefinite coefficients friondac

$$
\begin{aligned}
& \text { ORIGINAL' PAGA } \\
& \text { OF FOOR QUAS }
\end{aligned}
$$

to obtain the scheme of the successive approxinations. This method car. equally be applied to differential equations of any order, However, when the number of whnowne and number of equations increase, the volume of the required work will ise considerably increased.

On the basis of all the above arguments, Hill was convinced that it was more advantageous to integrate the differential equations of lunar motion by using rectangular coordinates. Once the rectangulat cootinates are obtained, the calculation of the corresponding polar coordinates becomes quite simple.

Euler, in the second memoir, had already applied an elliptueal rectangular coordinate system, rotaling with a velocity equal to the mean velocity of the mocn. Hill made use of a similar system, though rotating with a velocity equal to the mean velocity of the sun. Adams and subsequently Hill, systamatically coveioped Euler's idea on the separatc determination of irequalities of different powers relative to the parameters. For example, Hill rirst calculated the part of the motion of the perihelion whicin dicl not depend on the eccentricity of t.A solar orbit. Then, he calculated the part which was proportional to the Eirsi prater of the eccentricity, and so on. Hill applied this idea ever to the first approximation. Instead of start g with the elliptical orbit which results from the assumption that the mass $m^{\prime}$ of the sun is erqua: z. fero, he assumed that, in the first apaproxination, the parallax of the sun can be set equal to zero. This assimption leads to an original version of the three-budy problen, in which one of the three bodies goes to finfinity and at the same tine continues to infisence the motion of the other two. By this way, one obtains a varlatinnal orbit which incluass all the inequalities that depend on the angular distances of the moon and the sun. This oriit can be used 7 a a trial intermediary orbit.

We should pay attention to one more feature of H111's work. While all the previous authors used the ratio of the mean notions of the sun ant ihe moon, i.e.

$$
\mu=\frac{n^{\prime}}{n}
$$

as one of the parameters by which the expansions of the pelturbations are developed, Hill preffrred to expand the perturbations in powers of the ratio

$$
m=\frac{n}{n-n} .
$$

sin:e, in this case, the series converge more rapidiy.
130. Equations of Motion

We take a rectangular reliocentric coordinate system, in which the xy-nlane coincides with the plane of the ecliptic and which rotates about the z-axis with a constant angular velocity $n$ '. The equations of motion of the moon in this coordinate system ave given by (cf. section 38)

$$
\begin{align*}
\frac{d^{2} x}{d t^{2}}-2 n^{\prime} \frac{d y}{d t}-n^{2} x & =\frac{\partial V}{\partial x} \\
\frac{d^{2} y}{d t^{2}}+2 n^{\prime} \frac{d x}{d t}-n^{\prime 2} y & =\frac{\partial V}{\partial y}  \tag{1}\\
\frac{d^{2} z}{d t^{2}} & =\frac{d V}{3 z}
\end{align*}
$$

where $V$ is the force fructinn. According to the resultr obtafaed in sectior 4 , the function $V$ is given by the following expression

$$
V=R^{2} \frac{T+1}{r}+R^{2} \frac{r^{\prime}+L m^{\prime}}{L}+R^{2} T+L \frac{m^{\prime}}{\Delta}
$$

where $T, L$ and $m$ denote as previcusly indicated, the " sises of the earth, moon and sun, $r$ and $v^{\prime}$ are the heliocentric distances of titw noon ant the sun, whereas $\boldsymbol{\Delta}$ is the distance between them. As we have already see: in section 120 (annotation Jl), the eeries-expansion of the last tro torms yields.
where $r_{1}$ is the distance from the sun to the centre of gravity of the earth and the moon, and $H_{1}$ is the angle between the radius rectors $r$ and $\mathrm{r}_{1}$.

Using the arguments given in annotation I of section 120, we assume that the motion of the sun relative to the centre of gravity of the earthmoon system is strictly elliptical. Hence, denoting the semi-major axis of the sun's orbit by $a^{\prime}$, we obtain the following relation

$$
n^{\prime}: a^{\prime 3}=R^{3} m^{\prime}
$$

where $T+L+n^{\prime}$ has been replaced by $n^{\circ}$. Consequently, noting that $x_{2}=1$,

The terms which we did not write are multiplied by

$$
\frac{h^{\prime}: m^{\prime}}{r_{1}^{4}}=n^{\prime} \because\left(\frac{a}{r_{1}}\right)^{\cdot} \frac{1}{r_{1}}
$$

They tend to zero, when a and $m^{\prime}$ increase to infinity in such a way that the ratio

$$
\begin{gathered}
k^{\prime} m^{\prime} \\
a^{\prime},
\end{gathered}=n^{\prime \prime}
$$

remains finite. It is thus clear that, in order to obtain the inequalities of the mon n that do net depend on the parallax of the sim, wis must replace: $V$ in equation (1) by the function

Hill calculated the inequalities that depen 9 RIGINAL PAGE IS
OF POOt Quthetrara11ax: of the sun, nor on the eccentricity of the solar orbit. If $e^{\prime}=0$, then the sun moves uniformly with a velocity $n^{\prime}$ and we car. therefore take the $x$-axis in such a way that it always passes by the sun.. In this case, we obtain

$$
x^{\prime}==a^{\prime} \quad v^{\prime}=0, \quad z^{\prime}=0,
$$

and, therofore,

$$
r_{1}=a^{\prime}, \quad r \cos H_{1}=\frac{x x^{\prime}+y y^{\prime}+z z z^{\prime}}{r_{1}}=x .
$$

[n this case, the force function may ie taken as

$$
V_{2} \cdots k^{2} \frac{T-L}{r}+\frac{1}{2} n^{\prime 3}\left(3 x^{2}-r^{2}\right) .
$$

We finally introduce the parameter $m$ by means, of the following relation

$$
m=\frac{n^{\prime}}{n-n^{\prime}}
$$

where $n$ ' is the mean sidereal motion of the moon. We then write the finction $V_{2}$ in the following final form

$$
\begin{equation*}
v_{:}=k^{2} \frac{7+L}{r}-\frac{1}{2} n^{\prime}\left(x^{3}+v^{2}\right)+\frac{1}{2} m^{2}\left(n-n^{\prime}\right)^{2}\left(3 x^{2}-z^{2}\right) \tag{t}
\end{equation*}
$$

Furthermure, we put

$$
V_{1} \cdot V_{3}: \because
$$

where, as we can ecsily see,

Erpation (1) for the force function $V=V_{1}$ can be written as

$$
\begin{aligned}
& \frac{d^{2} x}{d t}-2 n^{\prime} \frac{d y}{d t}=\left(n-n^{\prime}\right)^{\prime}\left|-\frac{x x}{r^{3}}-1 \cdot 3 n^{2} x \cdot+\frac{d \varphi}{d x}\right|
\end{aligned}
$$

$$
\begin{aligned}
& \begin{array}{c}
d^{2} z
\end{array} \quad-\left(n-n^{\prime}\right)^{2}\left|-\frac{a z}{r^{3}}-m^{2} z+\begin{array}{l}
d U \\
d z
\end{array}\right| .
\end{aligned}
$$

where

$$
x=\frac{k^{\prime}(T \mid I)}{\left(n-n^{\prime}\right)^{2}}
$$

Putting $\mathcal{C}=\left(n-n^{\prime}\right) t$, we finaly obtain the following equations of motion

$$
\begin{align*}
& \frac{d^{2} x}{d=2}-2 m \frac{d y}{d z}+\frac{x x}{r^{2}}-3 m^{2} x=\frac{\partial L}{\partial x} \\
& d^{2} y+2 m \begin{array}{l}
d x \\
d=
\end{array}+-\frac{x y}{r^{3}} \quad==\begin{array}{l}
\partial Q \\
d y
\end{array}  \tag{6}\\
& \begin{array}{ll}
d: z & -\frac{x z}{r^{3}}+m^{2} z=\frac{d u}{d z}, \\
d: 2
\end{array}
\end{align*}
$$

which correspond to the case in which the farallax of the an is taken to be equal to zero. If the eccentricity of the solar oibit is also zere, then $\Omega=0$, Since

$$
n=\underset{27.32166}{\frac{2 \pi}{\pi}} . \quad . \quad \imath^{\prime}=\frac{2 \pi}{365} \cdot \frac{2 \pi}{2 \cdot 12} \frac{20}{20} .
$$

where $t$ is ex esscd in seconds while $\mathcal{Z}$ is expressed in unjts, ir whicil the period of the moon's siciereal rotation is equal to 2 .

The moon is at syzygies when $\mathcal{Z}=0,77,27, \ldots$ and at quadratures when $\tau=\frac{\pi}{2}=\frac{3 \pi}{2}, \ldots$

## 131. The Hill Transformailinn

In order to simplify the application of the mothod of indefinite coefficients to equation (6), H111 introduced new variables. Putting

$$
u=-. \mid-y i, \quad s \cdots x-y i
$$

ther the firgt two of equations (5) may be replaced by

Let us introduce the independent variable $\zeta$ by means of the following relation
since

$$
\begin{array}{lll}
i & d: d & \vdots \\
\therefore & i: a & \ddots i
\end{array}
$$

then, using the operator

$$
\square \quad \begin{aligned}
& d \\
& \because
\end{aligned}
$$

we reduce the rquations of motion into the finiumiay form
whare $r^{2}:=u s+z^{2}$.
In the following we need, apart from the previous equations, another $r \in l a t i o n$ analogous to the kinetic-encrgy integral. In order to dedure this relat ion, we multiply equations ( 6 ) bv $2 \frac{d x}{d t}, 2-\frac{d y}{d t}$ and $2 \cdot \frac{d z}{d i}$ respectively, add and integrate the resultigg equations. This procec'ure yields
..here, we have put

On the left-hand side of this equation is the square of the relativ. velocity of the moon, which will be denoted by $V$. Ir serting here our new variajles, we obtain

$$
\begin{equation*}
D u \cdot s) \left.s+(D z)^{2}+\frac{2 a}{r}+m^{3}(u ; s)^{2} \quad m^{2} z^{2} \right\rvert\, 2 \int_{i}^{0} d \leq . \tag{}
\end{equation*}
$$

When $\Omega=0$, this relation gives first integral, known as the Jacubi integral (cF. section 38 ).

Equations (8) are not convenierr fer the application of the successive approximation method due to the prsence of terms involving $r^{-3}$. In order to obbtain a more converient set of equations, we multiply equations (8) by $s,: 1$ and $2 z$ and add. We then obtain

Bing equation (9), we elimirate t'u cerm that involves $r^{-1}$ and obtain tie first of the following two equations


The second equarion has bee. obtalned ty multiplying equations (8) rewpetively by $-3,+s$ and 0 and adding.

Corsidering now the simple case in which the motion of the moon is assmed to froceed in an eclipticne plane and the eccentricity of tho srilat orbit to be eçual to zero; in thls case, $z=0$ an $\Omega=0$, whict refuces equations (1.0) and (9) to tice foliowing:

$$
\left.\begin{array}{l}
: \\
0
\end{array}\right\} \quad 111
$$

$$
\begin{equation*}
D u \cdot u s+\frac{3}{4} n^{2}\left(u+s r^{2}+2 \cdot(u s)^{-\frac{1}{2}}=C\right. \tag{12}
\end{equation*}
$$

It necessary to point out that equations (11) do not completely replace the initial equations

Which: cause obtained from equations (8) by putting $z=0$ and $\Omega=0$. When the solution of equation: (l) is already obtained, it is still necessary to substitute this solution into either equation (12) or equation (13) and find the relation between the constants $x$ and $C$,

## 132. The variational Curve

The general solution of equations (11) or the equivalent equations.
which are obtained from equatime (6) $y$ putting $z=0$ and $\Omega=0$, includes four arbitrary constants. We shall try to find these Four constants in such a way, that the corresponding trajectory will lp symmetric relative to the two coordinate axes. Since the trajercory is assumed to be free from singular points, we then reçire that it fruitersects each of the coordinate axis at right angles. Denoting by f te absiciste oi the point of intersection with the maxis, we then ovtain

$$
\begin{aligned}
& D\left(u D_{s}-s(i n)-\cdots 2 m D(u s)+\frac{3}{2} m^{2}\left(u^{\prime}-3^{2}\right)\right.
\end{aligned}
$$

where $\beta$ is an arbitrary constant. Combining these with the foiiowing equations

$$
1=0=0, \quad \frac{!y}{d!} \cdots 1
$$

which empress the fact that the curve also intcrsects the y-axis at right angles (at moment ' $C$, which can be excluded from these equations). We ottain four condilions which should, generally speaking, be satisfied by the solutson of system (14). This solution will depend on the two arbitrary constants A and $\beta$. M. can essily see that this solntion is symmetric relativ- io che $x$-axis. In fact, ecuation (14) will not be alterec if ve replace $\beta+\tau$ by $\beta-\mathcal{C}$, and at the same tire rcpla=e y by - y, keeping the value of x maltered. Similarly, we can verlfy the symetry of the solution relative to the $y$-axis.

In crier to reduce the voiume of the deduction, we fut $\beta=0$, wish is equiralent to the time count from the moment of intersention of the z-axis. Te choose the moment of antersection of the y-axis as the initial ine and oqull vo $\frac{\pi}{2}$. This shoice is definei hy the rotation period under consideration. We search for the particuise solution of equations (14) in the rome

$$
\begin{align*}
& x=A_{1} \cos =-A_{2} \cos 3=: A_{1} \cos \pi  \tag{15}\\
& y-\cdots A_{1}^{\prime} \sin =-A_{1} \sin 3-+-A_{1} \sin \pi \\
& i
\end{align*}
$$

or, by transforming to variabales $u, s$ and $\zeta$,

Putting,

$$
A_{\therefore+1}=a\left(a_{i k} \mid a_{0}, a^{i} \quad A_{:+1} \quad a\left(a_{k}-a, 1\right)\right.
$$

we obt:in

The problem is thus reduced to a search for coefficients $a_{c}, a_{2}, a_{i .2}, a_{4}$, ${ }^{1}-4$, .... Since we have separaced the comon facter a, we can for example ©e ${ }^{-} a_{0}=1$, after which the value of a can be defined. In order to simplify the substitution of expressioen (16) into equations (11), we initially evaluate the expressions involved in thesc equations. It is easy to see that

$$
\begin{aligned}
& =\mathrm{a}: \searrow_{i} a_{-k} a_{-k} .
\end{aligned}
$$

where $i=k+h+1$ runs ail the values between $-\infty$ to $+\infty$. Similarly,

Since,
then,

$$
\begin{aligned}
& \text { Lu } 1 \mathrm{~s}_{\mathrm{s}}-\mathbf{a}_{1}^{2}(2 k+1)\left(21-2 \mu-11 u_{-k} u_{k} \quad a_{a}\right.
\end{aligned}
$$

Fir.aily,

$$
D^{\prime \prime}(11 s)=a=\sum_{1} \sum_{k} l_{1}: a_{i k} a_{n k-a}:
$$

Suhs'. ituting all of these expressions into equations (13) and Equating t', corfficients of $\zeta^{\dot{i}}$, we ortain the following relations:

$$
\begin{aligned}
& \sum_{0} \left\lvert\, 4 i^{2}+(2 k+1)(2 k-2 i+1)+4(2 k--i+1) m+\left\{\left.\frac{9}{2} m^{2} \right\rvert\, a_{2 k} a_{2 k} z_{2}+t\right.\right.
\end{aligned}
$$

$$
\begin{align*}
& 4 i \sum_{k}(2 k-i+1+m) a_{2 k} a_{3 k}=
\end{align*}
$$

For $i=0$, the second of these ecuations hecomes an identity, wile the first equaticr conid be replaced by

In the next section, we shall see that the coefficients $a_{2 k}$, which satisfy equazion (17), will not only exist, but will also be relatively easy to olvair, for at least mall values of the parameter $m$.

Since the corresponding series (16) or (17) are convergent, the existance of a particular sclution fov equation (14) kaving the required form, is thas proved ty the abose argumentis.

The curve which is cefined by equations (14) is usualiy calied the variational curve. We shall now undertake to prove that this curve is inderd related to the inequality of the moon's motion, which we have called a vai fation. Ne first of all observe that also in this case expressions (15) satisfy equatioas (14), fa which $\mathcal{Z}$ is replaced by $\mathcal{Z}-B$ where $\beta$ is an arbitrary constant. Fi $d=n o t e$ the erue longitude of the moon at moment $L$ by $v$. We represent the nean Longizude of the mona at the same moneric in the form of nt $+\mathcal{E}$. We consider the following -xpression:

$$
r \cos (l--n t-\cdot) \text { nd } / \sin (t-n t-\varepsilon) .
$$

We have
where $n$ 't $+\mathcal{E}^{\prime}$ is the nean longitude of the sur:, and

However, the x-axis is thosen to pass via the sun, then
and, therefore,

Feplacing $\mathcal{\tau}$ in equation (15) by $\mathcal{\mathcal { C }}-\boldsymbol{\beta}$, we finally obtain


These two equations define the motion of tie moon that corresponds to the particular solution of system (14) unoer consideration. In order to obtain the iongitude of the movi from these equations, we En make use uf the following equatios

$$
\therefore \quad:!\quad 1: 1-: 1 \cdot 1
$$

 relative to m . In particulat,
'ihus, Confinting eurselves to rhtra order terms, we ottajn'

Comparing this expansion with formulae (45) of section 727 , we note that the terms which have been kept in the rorce function, reproduce the varjation. This justifies the name whick we have given io carve (15). 133. Galculation of the Cofefficients

Let us now consider the solution of equations (17). First cf all, we note that these eqiations can be rewritten in a much simpler manner by multipiylng them by 2 and 3 respectively, adding, then constructing the sum and the differsnce of the resulting equations. This proceudre yields

In order to obtain the terns that inrolve products $a_{0} a_{2 i}$ and $a_{0}{ }^{a}{ }_{-2 i}$, it is recessary to take in the first sum of each of these equations, the torns that correspond to the values $k=0$ and $k=f$. These ter are

Recause of thir, we mulitply the firet equatinn by

$$
\begin{aligned}
& \rightarrow 9 m^{2} \text { V. } 1 . \quad . \quad-11
\end{aligned}
$$

and the second by
and add the resulting equations. We obtain anaedation which dnes not cuntain a term involving the product $\mathrm{a}_{0}{ }^{\text {a }}-2 \mathrm{i}$. Putting

we finally ohtain


It is asy to see that this single equation completely replaces system (17). In fact, equations (17) are equivalent ic equationa (19), and each of the latter equations is a consequence of the cther. For axample, if we replace $k$ and $i$ in the first of equations (19) by $k-i$ and -1 , we obtain the second.

Equations (21) have the most convenient form for the determination of ${ }^{a_{21}}$ ince

$$
|\therefore, 川 \quad:| 1, \| \quad \cdots i
$$

while the quantities $[1]$ and (i) are second-order quantities rejative to the paraneter $m$, which we have agreed to consider as a small quantity. Let us now ass..ne that the quantity $a_{2 k}$ is of the $|2 k|=$ nrder relative to $m$. In thits ceise, the sum

$$
\sum^{\prime \prime \prime}
$$

$$
(. \therefore \cdot)
$$

where $1>0$, will consist of terms having orders at least four unirs larger than the order of equivalent terms in equation (21). The sum
in whicin $i<0$ will have exactly the same property. We are not going to consider the yalue $i=0$ etther here or in equations (17).

Let us now consider the calculation of the coefficients $a_{2 k}$ in the first approximation. For this purpose, we vrite equations (21) for different yalues of i keeping each time only the terms having the least order. Taking into constideration the properties of the sums(2.2) and (23) which we have just mentioned, we obtain

$$
\begin{aligned}
& u_{i}=\| 11 a_{1} a_{.} \\
& \Delta:(-1) a_{0} d \text {. } \\
& \left.a_{1} \quad|\therefore|(a) a: \mid-a, a_{u}\right) \cdots|\therefore 1| a_{3} a \\
& \text { a. },(-\because)(a, a,-a, a): \mid-\therefore,-1) a, a
\end{aligned}
$$

where we keep in these equations the coefficient $a_{0}=i$ in order to cleariy show the srheme construction of the successtve terms. The solution of this syeter: of recurring equations is not difficult. It yields for the coefficient $a_{2 k}$, a quentity having the order $\mid 2 \mathrm{~F}$. .

In order to obtain more accurate values, we have to repe tit
calculatior keeping not only the terms that have the least creer but also the terms which have next to lenst order. T.: che first arroyfnetiou,
we obtain $a_{2 k}$ with an error of the $(|2 k|+4)$. order, and in the second approximation an error of the $|2 k|+8$ order, and so on. Thls illustrates thet suo: ossive approximations converge sufficiently rapidly for small values of the parameter $m$.

In order to show how simple these calculations cre, we write the equation which can be used to calculate the coeffizient $a_{2}$ within an error of the 14 th order. This equation is

Sefore calculating the quantities $[i, k],[i]$ and (i), we can simplify equations (20). Indeed, it is not difficult to see that

Similarly,

$$
\begin{align*}
& |i| f(-1)=-\begin{array}{cc}
3 & 3 i: 11 \cdot 2 m \\
2 i:\left(1 i^{2} \quad 1\right)-4 m ; m:
\end{array} \tag{2.4}
\end{align*}
$$

These formulae are nore convenfent than equations (20).
Tet us consider the case when it is required to calculace the variational curvo for only one given yalue of $m$. In this case, it if easier to immeciately calculate the numerical values of the cuefficients. Hill adopted that

$$
\vdots, 1 . i \operatorname{lin} \quad: \quad . \quad \text {.. i: ir }
$$

for which case be obtained

At the start, he calculated all the recessary valses for the quantittes
$[i, k],[1]$ and (i), by means of the above equ: tions for ihis value of $m$. He then proceeded to compute the valves of the coefficients $a_{2}, a_{-2}, \ldots$ by means of successive approximations. The computation of the first two


Hili's final results were given in thefollowing form:

Giving this result in a supplenent to the translation of Euler's book
"New Theory of Lunar Mo n:" cited above in section 117, Academfian A.N. Xiylov observed that a $\approx 10^{-14}$ produ es a correction to the distance between the centres of gravity of the earrh and moon, approximately equal to 4 microns. This urusual and practicaliy useirss accuracy, shows the tremendo:ls power of hill's mether, which enables us to obtain this accuracy througit a relatively small amount of work.

## 134. General Fxpressions for the Coefficients The equations, deduced in the previous section, can be net only

 used for obtaining the numerical values of the coefficfents $a_{2 i}$, but also for deriving genera? expressions for these coefficients as functions of $m$. Formulae (24) and (24') show that the factors $[1, k],[1]$ and (i) are rational functions of with denominators of the form$$
\because(: 1: 11-1 \Gamma): \because \because
$$

It is thus clear thac each of the unknown coefficients can easily be rerresented in the result of the successire approximations by a double series in the form
where each of the quanities $M_{0}, M_{y}, \ldots, N_{1}, \ldots$ is a double term in the form
raving rational onefficients. This merfes only converges for the values of m which are smaller than the least, by a modulus of the roots of the denorimatur (2i), i.e. for $m<\sqrt{6}$.

Expanding sach term of series (26) in powers of $n$, Hili obtained

$$
\begin{aligned}
& \because \text { 肘い! } \because 1
\end{aligned}
$$

The dffficult problem of the convergence of the power－series obtained here was studfer by A．M．Ljapliov in his excellent book＂9n the series suggested by Hill forrepresenting the lunar moifon＂ （O rjadan predilozennyh Hillom dila predstavlentja dvizenija Luny），in which he proved that these serles converge for m $<\frac{\square}{7}$ ．Since，for the moon，$m=0.0808 \ldots \approx \frac{1}{12}$ ，we can then consicler that tin application of H111＇s method to this case is fustiffed．The exact Imits of convergence of Hill＇s series are still unkorn．

Iت we confine ourselves to second－crder terms relative to m ， thent
and, hence, equations (15) yield


Taking tiato account that

we can easily reduce the equation of the variat fomil eurve, withia tre accuracy required, into the form


In this maner, the varjational curve has the furm of a circle for $m=0$. When $n$ is incraaed, this curve will have a form similat to an ellipse with a centre at the origin of the coordinates and a semimajor axis equal to $\left(1-\mathrm{m}^{2}\right) /\left(1+m^{2}\right)$. We note that it rollows from the last tro equations that

$$
y .111 .60 .
$$

We finally consider the calculation of the cominn factor a giver by equation (15). For this purpose we take gny one of tite nonhoncegeneous equations relative to $u$ and $s$. We tare the first of equations (13), which may be written as
or
because
У:
when $\zeta=1$, this equation becomes

Since
then

Denoting by a the semi-major axis ccresponding to the means motion n of the unperturbed motion, the third law of Kepler gives

$$
1 . u^{\prime}: 11: l .!
$$

Comparing this equation with the previous ones and using the values of the coefficients $a_{2 k}$ obtained above, we easily chtain

We thus conclude that the form of c ne variational curve is completely ceflnet by tine value of $m$. In order to find the dimensions of this curve,
characterizer by the quantity $a$, we have also to hnov the vaius of $n$.

La the following, we sholl also need the relation (*) between the quabities $x$ and a. This raiarion may be given, to withtr terms oit tae order of magnitude of $m^{2}$ in luefve-y, by the rotioning equation

$$
\text { -1 } \quad 1 \cdot \because \cdots: \quad: \quad:: \text {. . . }
$$

135. Orbits Infinfte'y ilose to the Variational Curve

In the previous section, we stidied in detail the variational curves which appear as a particular solution of equetions (14) having the turm (15). It is interestirg to know to what extent this solution fs applicable from the point of yiew ef closeness to che actual lunar orbtt. If $m=0$, equations (if) are reduced to the well fnown equations of the two-body problen. They describe ar elliptical motion. On the otne: hand, equations (15) are recuced $\pm 0$
i.e., represent a circuiar motion. Hence, for small values of me we can regard the moition described hy the variational curve as a motion along a cirnalar orbit, deformed by the gttraction of the sui: The actual orbit of the moon lonis mire like an ellipee than a circle, hence. we cannst: confine surselves to the study of soluticr (l.) for equations (14). We have to consider more general solutions far these equations.
lie cail the general solution or squations (14) which invulves four arbitrary constants, a varfational orbit. Fe ihall consider the difficult problen of deterninime the variational orbit and start in seudving the particular case of calculating orbits, Enfinftely close to a varational orbjt, i.e. on'its that corresponc to the Eifiptical onits cif the twobody problem, the eccentricizies of which are sc smati thot their aquares may le regleated. de wrlte equations (i4) in the fotioving manir
where

$$
r \quad, \quad 1: m=1
$$

We denote by $x$ and $y$ the cocrdinates of an arbitrary point on the variational curve (15), and by $x+5 x$ and $y+\delta y$ the coordiates of a corresponding point $P$ ' cri a clese curve. By "corresponding point", we nean a puint related to the same moment $\mathcal{C}$. Considering that the inciements $\delta x$ and $\delta y$ are infinitesimai quantities, their squares may be negiected.

Substituting the cocrdinates of points $P$ and $P^{\prime}$ into equation: (27) and substracting the resuiting equations term by team, we ottain for the determination oi the increnterts $6:$ and $\delta$ y the following equations:

$$
.1
$$

$$
\therefore 1
$$

where

$$
\therefore \quad 4 \quad \therefore \quad 4.4
$$



Fig. 15
In cur cast, $z=0$ and $\Omega=0$ and hence ccuation (9) ieads tc the tolluwing form of the lecobi integral


We confiae ourselves to the consideration of only those adjac?nt orbits, fur whter, the conseant $i$ has the same vains as the initial variaticrad curve. We use t'ie lacobi fucsera? In the same marner ae we inve just used ecluation (2T). We thus uhtain

$$
\begin{array}{llll}
d & 4 & : \therefore \\
4 & d & d & d
\end{array}
$$

(:i)

Denoting by $\mathcal{E I}$ and $S N$, the tangential anc normal displacement of point $F^{\prime}$ relative to point $P$, and by $\Psi$, the angle formed by the tangent to the variational curve with the $x$-axis (see figure 15), we obtain

$$
\therefore \quad \therefore \quad 1,1 \times 1,0,
$$

which enabies us to obtain Erum equations (28) and (29) the equations required for the decermination of $\delta N$ and $\delta T$. We start by transfurming equation (29), First of all, we have
anċ since
we rep..ane equation (30) by

On the other hand, we oidain from equation (33)
then, because of equation (2T),

$$
\begin{array}{llllll}
1 & \ldots & \therefore & \therefore 1 & ! & . .1 \\
\text {.. } & & .1 & \therefore & \cdots & .
\end{array}
$$

or, using asion equation (32) and noting that

we finally obtain

Once again we consider the Jacobi integral (29), whtch yfelds

$$
\begin{array}{cc}
1: 3 \\
. & \therefore \\
6
\end{array}
$$

Huwever,
from which it foricus wat

$$
\begin{array}{ll}
\therefore & \therefore \\
\therefore \vdots & \ddots 1 s
\end{array}
$$

We wite equation (34) in the following manner

Eliminating the partial derivatives of the function $F$ by means of equations (36) and (38), we obtail.

$$
\begin{array}{cccccc}
\therefore & \because & 1 & \cdots & \cdots & n \\
\therefore & \cdots & 1 & & . &
\end{array}
$$

$$
1 .!\cdot
$$

liaving transformed the jacobi integral snd derived the duxiliary relations, we consifer equations (28). Multiplying these equations by $-\sin \psi \quad n 2+\cos \psi$ respectively and addirg, we obtein

Using equations ( $\because 心$; , we transform the quatities founa fnsife the rquere brackets into

Similarly,

$$
-\frac{d \hat{d} x}{d:} \sin \psi+\frac{\sin y}{d t} e \cos \dot{\alpha}=\begin{gathered}
d i v \\
d=
\end{gathered} d ;
$$

Differentiating this equition, and using again equations (32), we obtain

This enables us to represert equation (40) in the following maner

TE now have to transform the last term. First of all, wo have

$$
\begin{equation*}
\frac{d_{i} F}{\partial N}=\hat{i} \frac{\partial F}{\partial N}=\frac{\partial^{2} F}{\partial N} \partial \hat{\partial N}+\frac{\partial^{2} F}{\partial N \partial T} \partial T \tag{42}
\end{equation*}
$$

Differentiatiag by $\mathcal{C}$ equation

$$
\frac{\partial F}{\partial N}=\cos \psi \frac{\partial F}{\partial y}-\sin \psi \frac{\partial F}{\partial x}
$$

we obtain

$$
\begin{gathered}
d-\frac{\partial F}{d \bar{N}}=-\frac{d \psi \partial F}{d \tau}+\frac{\partial F}{d F} \cos ^{d} \frac{d}{d \tau} \frac{\partial F}{d y}-\sin \psi \frac{d}{d \tau} \frac{\partial F}{d x}= \\
\left.=-\frac{d \psi}{d \tau} d F+V \left\lvert\,\left(-\frac{d^{2} F}{\partial x^{2}+}+\frac{\partial F}{\partial y^{2}}\right) \cos \psi \sin \psi+\cdot \frac{1 F}{\partial x \partial y}\left(\cos ^{2} \psi-\sin ^{2} \psi\right)\right.\right]
\end{gathered}
$$

or, noting that it follows from equation (35) that

$$
\frac{d: \%}{d N d \%}=\left(-\frac{d^{2} F}{d x^{2}}: \frac{\partial F}{d y^{3}}\right) \cos \psi \sin \psi+\frac{d^{2} F}{\partial x d y}\left(\cos ^{2} \psi-\sin ^{2} \psi\right)
$$

we Inally obiain

$$
\frac{d}{d:} \frac{d F}{d N}=-\frac{d \psi}{d:} \frac{d F}{\partial T}+V \frac{d^{2} T}{d i V d T^{\cdot}}
$$

Using this relation to exclude the second derivative of $F$ from equation (42), we obtain
where $\frac{\partial F}{\partial N}$ is replaced by expressicn (36). Finally, using equation (38), we obtain

Having on hand this flnal expression for the last term of equation (41) we can rewrice this equation in the following manner
or, using equation (29),

$$
\begin{equation*}
\frac{d=i N}{d i^{2}} \quad H_{i} N=0 \tag{43}
\end{equation*}
$$

where

$$
\begin{equation*}
H=3\left(\frac{d^{\prime}}{d=}+m\right)^{3}+m^{2}-\frac{d^{2} F}{d N^{2}} . \tag{-44}
\end{equation*}
$$

Ir order to calculate the function $(\mathcal{Q}$, ve male use of the followirg formulae

$$
\begin{aligned}
& d N^{2}=\sin ^{2}: \frac{d: F}{d x^{2}} \quad 2 \sin \dot{4} \cos \dot{\psi} \frac{d^{2} F}{d x d y}+\cos ^{2} \psi d y^{2} \\
& \frac{d^{2}}{d}=V^{2}
\end{aligned}
$$

in comhination with equatior (33).
After fiuding the incremeat $\delta v$ fron equaticn (43), wo can obtain increment 8 T ky using equation (39).

This method of obtaining infinitely close soluticns can be applied to all equations having the forri (27), in which the force function $F$ does not explicitiy depend on time. As we have already seen, the problem will be reduced to che solution of the principa] equation (43).

In our particular case, the force function has the following form:

$$
I=x\left(x^{n}+y^{\prime}\right)^{3} \pi^{2} x^{n},
$$

where $x$ and $y$ are defined by the series (15). Since the coordinates $x$ and $y$ either do not change or only change their signs when the variaiole $\tau$ is changed into $-\mathcal{V}$ or $\tau \therefore 7$, then the force function $F$ is an even periodic function of $\tau$, having a peried equal to $\pi$. It: is easy tc see that the second partial derivative $\frac{\partial^{2} F}{\partial N^{2}}$ as well as the derivative $\frac{d \psi}{d q}$ will also have this property. Accordingly, the function $(\mathcal{S}$ can le expanded into a series of the type

$$
H-4-2-2 s_{1} \cos 2--2 \varphi_{1} \text { cos } t-+\ldots .
$$

The cortesponding equation (43) is cailed Hill's equations.
It can be proved that the function ( 4 can be expanded in a series, deveioped by positive integral powers of the quantities $m, m^{2} \zeta$. and $m^{2} \zeta^{-2}$, from whith it follows that the expansion of the coefficfent $\eta_{k}$ in powers of $m$ starts from a tern having the order $2 k$. We shall n.t give hare the proof of this property of the cocfficionts $q_{k}$. lhe expansion of these coefficients is conveniently carried out if the coordinates $x$ and $y$ are rcplared ty tie variabies $s$ and $u$. The resuit of the expansion is tioe foilowing ${ }^{(1)}$ :
(1) U.W. Hill, Literal Expansiun for the Kotion of the Moon's Perigete, Amais os liathematics, 9, 1894, 31-41 (Works, iV, 4i-50).

$$
4^{2}=1+2 m-\frac{1}{2} m^{2}+\frac{255}{35} m^{4}+19 m^{5}+\frac{89}{3} m^{5}+
$$

$$
\begin{aligned}
& -33 \\
& 2.3^{2}
\end{aligned} m^{-}+\begin{gathered}
11230225 \\
2^{1} .3^{3}
\end{gathered} m^{\times}{ }^{1} 1576037 m^{3} .3^{4}+
$$

$$
\begin{gathered}
49359583 \\
2: .3^{5}
\end{gathered} m^{10}!\begin{gathered}
720508007 \\
2^{4} .3^{2} .5
\end{gathered} n^{11}-\ldots .
$$

$$
q_{1}=-\frac{15}{2} m^{:}-\frac{57}{4} m^{3}-11 m^{4}-\frac{23}{2.3} m^{i}-\frac{68803}{2^{4} \cdot 3^{2}} m^{6}-
$$

$$
-\quad \begin{gathered}
2641291011773 \\
21^{19} .3^{6} .5^{3}
\end{gathered} m^{10}+\ldots .
$$

$$
q_{2}=+\frac{111}{16} m^{4}+\frac{1397}{2^{6}} m^{5}+\frac{8807}{24.3 .5} m^{6}+\frac{319003}{2^{5} \cdot 3^{2} \cdot 5^{\frac{3}{3}}} m^{7}+
$$

$$
+\begin{gathered}
252382507 \\
2^{10} .3^{3} .5^{3}
\end{gathered} m^{n}+\ldots .
$$

$$
q_{8}==-{ }_{2}^{11669} m^{6} \cdot \vdots . .
$$

In the following, we give the numerical yalues of these coefficients for $\mathrm{m}=0.080848033908212$ used by Hill:

However, tie rumerical method enables us to obtain the coefficients of this series with the same accuracy but more simply than the aigetrate method based on the above-mentioned expansions in powers of $m$.

In conclusion, we note that equation ( $j$ ( $)$ devoted to the delermination of $\delta$ ? can be rewritcen to within the second powers of $m$ ir the foliowing ma"ner.

$$
\begin{aligned}
& \theta==\quad 1.158843939596583 \\
& -0.1140880374!13807 \cos 2 \div \\
& -0.000766 .175995109 \cos 45 \\
& -0.000018316577790 \text { cos } 6 \mathrm{t} \\
& -0.000000108895009 \cos 8: \\
& \text { - 0.0000' (N020 98071 cos } 10 \text { r } \\
& \text { - } 0.00000 \text { (x)000 12103 cos 12: } \\
& -0.100000001000211 \cos 14 .
\end{aligned}
$$


136. Some roperties of Hili's Equation

In the prevjous section, we have reduced the problen: of finding orbits infinitely close to the variational orbit, to the solution of Hill's equation

$$
\|^{\prime} \mid H_{1}=11
$$

in which the coefficient

$$
H \ldots q^{\prime}-2 q_{1} \cos 2=-24, \cos 1=1 \because q_{9} \cos (i=: .
$$

is a periodic function of period $\pi$. We shall first of ail co.sider some properties of this equaionn, which are particulor cases of the properties of ali the linear differential equations with pericdic co:fficients.

Let as denote by $E(\tau)$ and $\varphi(\tau)$ two of the particular solurions of equations (46), which satisfy the following initial condtions

$$
f(0)=1, \quad f^{\prime}(0)=0 ; \quad \because(0)-0 . \quad \because^{\prime}(0)=1
$$

Since these solutions from a fundanatal system, then any arjitrary solution, $F(\tau)$, of equation (46) may be represented by

$$
F(\xi)=A f(\tau)+B ;(:),
$$

where $A$ and $P$ are con rents. Equations (46) does not change when $\tau$ is replacel by $\tau+\pi$. Hence, the functions $f(\tau+\pi)$ and $\varphi(\tau+\pi)$ are also soiutions of this equation. Consequently there exisis such constant numbers as $\alpha, \beta, \gamma$ and $\delta_{1}$, thus

$$
\begin{equation*}
f(\tau+\pi)=a_{f}^{\prime}(\tau)+\beta f(\tau) \tag{47}
\end{equation*}
$$

We prove that equation (46) has a solution which satisfies the following condition

$$
\begin{equation*}
F(s-1-\pi)=v f(t), \tag{48}
\end{equation*}
$$

Where $V$ is a constant. This condition gives

$$
A(a f+\beta q)+B\left(\gamma f+i_{q}\right)=v(A / ; B q)
$$

where the argument $\tau$ is dropped. The functions $f$ and $\varphi$ form a fundamental system and hence it follows from the previous condition that

$$
A(\alpha-v)+B ;=0, \quad A_{i}-B\left(i_{1}-v\right)=0 .
$$

Sirce $A$ and $E$ cannet simultaneously be equal to zero, then

$$
\left|\begin{array}{cc}
a-v & 7 \\
i & i-v
\end{array}\right|==0,
$$

cr

$$
\begin{equation*}
v^{2}-(x+i) v \mid(x i \quad, i)=0 . \tag{49}
\end{equation*}
$$

Each of the roots of this equation gives a solution $F(\tau)$ which satisfyes relation (48). In this way, the determinetion of che factor Which will have a [undamental value in our tuture disutssions, is reduced to the search for the substltution (47) thai the fundamental systen $\dot{F}, 母$ is subject to when the argument $\tau$ is increased by a per:od of $\not T$. If the functions $f$ and $\varphi$ satisfy the above-mentiones initial co ditions, ther equation (49) may be stap? ified. Indeed equation

$$
\left.\begin{array}{ll}
d \prime \\
d
\end{array},+1 f=0, \quad d:+H:-1\right)
$$

gives

$$
\int_{d \tau^{2}}^{d^{2} \hat{5}} \quad 4^{d^{2} f}=0
$$

Integrating and making use of the initiai conditions, we obtain

$$
f(-) \xi^{\prime}(:)-r(-) f^{\prime}(:)=1 .
$$

Putting $\mathcal{C}=\pi$ in this equation and noting that when $\mathcal{C}=0$, equations (47) Bive

$$
f(\pi)=\alpha, \quad \gamma^{\prime}(\pi)==\gamma, \quad f^{\prime}(\pi)=\psi_{i}, \quad \varphi^{\prime}(\pi)=\bar{i},
$$

ve obtain

$$
2^{i j-1}-1 .
$$

In the case under sonsideration, equation (49) will then have the Following for 1

$$
v:-(x \vdots i) v:-1=0,
$$

so that jits roots nay be denoted by $\nu$ and $1 / v$. Hence

$$
v+\frac{1}{v}=a+i
$$

When $\tau=0$ and $\tau=-\pi$, equation (4४) yieids

$$
r(\pi)=: v F(0), \quad F(-\pi)=\frac{1}{r} F(0)
$$

from which it follows chat

$$
v+\frac{1}{v}=\underset{F(0)}{F(\pi)+F(=\pi)}
$$

On che other hand, it is easy to see that $f(\mathcal{Z})$ is an even function of $\tau$, while $\varphi(\tau)$ is an odd function. Therefore,

$$
\begin{array}{r}
F(\pi)=A f(\pi)+B \varphi(\pi) \\
F(-\pi)=A f(\pi)-B \varphi(\pi) .
\end{array}
$$

Moreover, since $¥(0)=A$, then we fina $1^{1} . y$ obtain

$$
v+\frac{1}{v}=2 f(\pi)
$$

This form of equation (49) shows that, for small values of the paraneter n, the roots $\nu$ and $?^{-1}$ are complex conjugate numbers having a modulus equal to unity. Indeed, the lazter equation yieids

$$
v=f(\pi) \pm V|f(\pi)|^{2}-1
$$

Sn the other hand, using the approximete values of $r^{2}, q_{1}, q_{2}, \ldots$, given at the end of the previous seccion and neglecting terms of the order of $\mathrm{m}^{\text {? }}$, we obtain

$$
\frac{d^{2} x}{d x^{2}}+(1+2 m) x=0
$$

from whicn it follows that, in the first approsimarion

$$
f(:)-\cos (1+m) \cdot
$$

and inence $|\mathrm{f}(\pi)|<1$. Thus, if we $\mathrm{P}: \mathrm{C}$

$$
v=\operatorname{cop}(i c \pi)
$$

then $r$ would be a real number differing slightiy fron unity, a: lesec for small values cf m.

Considering the function erp (ic $\tau$ ). Fvident1y,

$$
\exp |i c(s+\pi)|=v \exp (i c s) .
$$

i.e. thitis function satisfies the same relation (48), satisfed by $\mathrm{F}(\tau)$. Hence, it follows that the ratio

$$
\frac{F(=)}{\exp (i c:)}=\omega(=)
$$

does :lot change when $\Pi_{i s}$ added to the argument.
Finally, noting that equation (46) does not change when $+\tau$ is replaced by $-\tau$, we may concluade that equation ( 4 (6) has two solutions of the form

$$
\text { W( } \tau) \cdot \exp (i c:) W(-\tau) \exp (-i c \tau),
$$

where $\Phi(\tau)$ is a function of the period $\Pi$. As it can be easily seen, these solutions form a fundamental system.

Introducing, as in the previous section, the folloring independent variable

$$
:=-\exp (i=) .
$$

and putting $q_{0}=q$ and $q_{-k}=q_{k}$ we can write the function

$$
N=q^{2}-2 q_{1} \cos 2 x+\ldots=\sum^{2} q_{k} \cos 2 k:
$$

in the following manner

$$
H \quad \vdots \quad q_{k}=k .
$$

In the following, we assume that the cocfficientis $q_{k}$ are such that the series $\sum\left|q_{k}\right|$ converges. Since the function $\oint(\tau)$ has the same perind $\pi$, it can thcu be expanded in a series of a similar form. Hence, puting

$$
\|(-) \quad \sum A_{k} \cdots
$$

we obtain

Thus, Hill's equations have a genera? :clution of the form (51). If this solution is found, for which it is necessary to compute the coefficfents $=$ and $b_{p}$, the gencral sointion may be given in the form

$$
C_{1} x(5)+C_{2} x(--=)
$$

where $C_{1}$ and $C_{2}$ are arbitrary constants.
137. App tcation of the method of indefinite coefficients

In order to find the constan': $c$ and $b_{1}$, we substitute eypression (51) into equation (46), and, takinf, into account equation (50), we obtain

Equac:... ine coefficient $\tau^{2 k+c}$ to zero, we obtain the following system of equations:

$$
\begin{equation*}
140-\left(2 k+c!3!b_{k}+\sum a_{k}, b_{1}=0 . \quad(i+k)\right. \tag{5,2}
\end{equation*}
$$

This is an infinite system of linear squations with infinite number of urinuown $b_{k}$. This did not prevent Hill in applying theorems oniy proved ior the sase of finite systems of linear equatiots to tifs sytem. The cesults he corained werc strictly justified by Poincar?, who dieveloped for thi: purposu, a therory of infinite determinants.

Let us consider the simple iase in which $q_{1}=q_{2}=\ldots=0$. Since $q_{0}=q^{2}$ and $q_{-k}=q_{k}$, ihen erdation (52) will in this case have the following in.m:

$$
\left|\varphi^{2}-(x k+c)^{2}\right| b_{A}=0 .
$$

We are only interestec ir the solutions of the =ype (ri) for shiria all the coefficients $!_{k}$ are equal to zeri. ilence, re will have

$$
c=-2 n: q, \quad b_{4}-0 \text { м.ля } \beta: n \text {; }
$$

and conseçuertiy obtain tro such solutions

$$
\lambda-b_{n} \Gamma_{V} \quad . \quad l=L_{n}:-4
$$

of equation (46) which in tl:e prosent care reads

$$
\frac{d: x}{d:^{2}} ; q^{2} x=0
$$

We shatll now consider thegeneral case jr. tr-ch the co-feici=eaj $4_{1}, 7_{2}, .$. are not equal o zero. We shaji enaidn. ourgeives to the
 walues of wiil rot be equa? io - 2n $\neq$, whore $n$ is an integer, flu
 that neither of the expressions

$$
q^{2}--(2 t+c)^{2}
$$



$$
\begin{equation*}
b_{k}-1 \sum_{q^{2}-\left(2 k^{\prime}:()^{2} b=-U\right.}^{4_{n}} \tag{i+k}
\end{equation*}
$$

cr, in the miolred fomin:

$$
1-1,1
$$

[^3]the :lass of nomal infinite determinants. Those are the tetermirants of the type
for which the doubie series $\sum_{i, j} f_{i j} \mid$ is comvergent. In fact, in the case uncer censideration,
and hence
where bott: of the sexius sisujing ar the right-inad sides are ecnvergent.
In the folinwing, we shall maly consider the bounce syster: for the ad dind di equations (53). This is the system or soiutionr which satisfies the followirg condition
$$
\mid b_{2}!\quad A_{1}
$$

Where A is some constant. It is wejl know thet in the case of bounded rysiens of soiutions, an intintte system: of linear equation for whit the leterminant composed by coefficionts fanornal, wil? have the ame properties as that of finite systoms. In paricular, one may conclude that when the determinant courosed by the cuefficients vanisles, the systan (53) will haye a soiution only in the case in which all of the coeffic; nts are nua! to zero. We denote this determinart by $\Delta(c)$.

Thus, the problen of índing solutions of the type (53) of Hirl's cequation ean be divided into two parls. Firstev, is is requirod to find the roots of the fcllowing equation:

$$
\begin{equation*}
د(c)=10 \tag{171}
\end{equation*}
$$

and, secndyy, to sclve equaticrs (53) for the resuleing vaires of $c$, de start by the first part and try to find tine sojution of equation (54). The existance of the roots of this equation fellows from tie arguments given in the preceding section. Ve inftably ronsider the function
$\Delta(z)$ of the complex variable $z$. Ry this function we mean the valce of the determinant, the $k \frac{t h}{}$ row of which

$$
\cdots u_{k}=a_{k, k-1}, 1+a_{k, k}, a_{k, k+1} \ldots \ldots
$$

consisis of terns respectively edual to
where $k=\ldots,-2,-1,0,+1,+2, \ldots$. It follows fron the p:exious alguments that sisch a determinanl is norm? for all values of $z$, except

$$
=14 \therefore . \quad 1 \text { in) }
$$

Hence, the functica $\Delta(z)$ is holomorphic for all the points $z$, except the points deffned by equation (56). It is easy to see tat fheste iatter points are first-order poies of this function. In fact, ench of peirts (56) is a poit of the first order for all the points of the li-h rov, except for the single terna that is ecual o unity, and is a :egnlar point for all of the other terms. When we vork out the nernal geceminabic, ve ohtain a convergent series, ir which eact. tenu har une of the k.th fow terms as a multiplying factor, ft is : fus elear lat the points (55)
can only be pales of an order not higher than the first for the function $\Delta$ ( 2 ). Therefore, $\Delta(z$; is a meromorphic furction. It is easy to see that this function is ar even function such, that

$$
د(\cdots z) \quad د(z)
$$

(.)7)

Indeed, if $z$ is reflaced by $-z$ and if the columns are at the same time replaced by the rows, then the deterainant $\Delta(z)$ will not te charged. Sindilarly,

$$
\Delta(z:-\because)=-\lambda(z) .
$$

since if $z$ is replaced by $z+2$, each column ard each row rwíl change place with the zext coilumi and the next row. Accordingly, the function $\Delta(z)$ is a periodic function witit $\equiv$ puriod equal to 2.

It follows from equations (54), (57) and (58) that all furits $z=$ $\pm$ C - 2l, were $k$ is Er: antitrary integer, are nodes of the function $\Delta(z)$.

We point out another property of the functicn $\Delta$, $\because$, rrirg $=-2 y i$ and letting $y$ tend to $\pm \infty$, all terms (55) of the kin row tend to zero, except the term $1+a_{k l}=1$.

Cons guentiy,

$$
\lim _{\rightarrow \pm \infty} د\left(\begin{array}{ll}
x & \because
\end{array}\right)=1 .
$$

We shilil now prove that this latter property completely defines the ianction $\Delta(z)$. We consider the foilowing meromorpilic funetion

```
Con:- cosic
cos:=-cos-.l
```

Which has first-order nodes at points (59) and first-order poles at pointe (55). This is an sven Eunction having a period fqual te $\therefore$. If wo put $:=x+y i$ and let $y$ tend to $\pm \infty$, then this function tonds

ORIG:.
1.7.!...
$-561$.
ORIGRAL, YAGE Y
to unity in analogy with function $\Delta(z)$. Henceitwerenclide that the ratio of the functions uncer corsicieration, i.e.,

$$
f(z)-د(z) \begin{aligned}
& \cos : \pi-\cos \pi y \\
& \cos -z-\cos -c
\end{aligned}
$$

is a zegulaz funciion of period 2 , which tends to unity then y tends to $\pm \infty$. However, this function mist be equal to a sonsiant because it renairs finite for all the points of the somplex 2 -pane. Nat fing $y \rightarrow \pm \infty$, we obtain that $F(z):=1$, and henca we finally obtafo

$$
د(z)-\begin{aligned}
& \cos \therefore=\cos =c \\
& \cos -i-c o s=4
\end{aligned} .
$$

When $z=0$, tinis equation yields

$$
\because: 1^{\circ}!=11,11-\vdots
$$

In tris maruer, we h-y: : eciursd tion probler of solving equation is io the penjlem ó fincing the determinatit $\Delta$ (o), whict is a reiat ivery 2a;" $\quad$ : cobler.

Futing $z=1$ or $z=1, n r z=?$, we obiain different forms for che equation that determines :. Ho:sever these forms are not as convenient as equation (50).
138. Calculetion of tiouserninant $\triangle(0)$

Introducing the followirg rotation

$$
\therefore \quad \frac{1}{4}+15^{-}
$$

we write the determinant $\Delta(0)$ as

At the end of the previous sectior, we fointeg cut that the coeificient $q_{j}$ is a quantity of the $2 j$ order relative to $m$. This situaticn enabies us to arrik out cie deceminant $\triangle(C)$ inte a rapidy convorgent series. We earsy out this expansion on tle basis of the following property of the determiriant (61):

If $A q_{\alpha} q_{\beta} \cdots q_{\gamma}$ is one of the terns of the expensior zif the determinant (61), then the suig f findices

$$
2 \text { - } 3+\cdots
$$

is always an even number.
Ir orier to rove ibis, ne =epiace the quantity $q_{i}$ in all the terms of expressiun ( 6,1 ; by the quantity $\mathrm{q}_{\mathrm{j}} \mathrm{z}^{j}$ and shew that the deterinant $\Delta$ ( 0,2 ) obtained thisway is en even function of $z$. indeed, the deterainant: $\Delta(J,-2) \therefore s$ otrained from the determinant $\Delta(0, z)$ if the sirm of aii the terms of the rovs and of the columns are alternately changed. Since the number of rows is equal to the numter of colurrs in ail the riait: Eeterminants, the limit of which is (G1). then this shange in signs of the terms wili nct cianaze tie feceminant. Lt therefore frmediately iollust that the expansior of the deterrinan föz will =onsist of terns, each



$$
د(0)=1+A \varphi_{1}+B \varphi_{1}^{1}+C v_{1}^{2} Q_{1}+\Delta \varphi_{:}^{2} .
$$

where the sum of indices are only $\cap, 2$ and $\dot{t}$. It is easy to see tha:

$$
\begin{aligned}
& A \varphi_{i}=\sum\left\{\begin{array}{lll}
0 & i_{1}, \theta_{1} \\
B_{1} \sigma_{1}, & 0
\end{array}=-4_{i} \sum_{1}, \theta_{1}\right.
\end{aligned}
$$

In the latter eritation, $k$ cannot be equal to $i$, $i-1$ and $i$, 1 . Hence

The coefficients $A, B, C, E, \ldots$ are easily expressed in terms of $q$.
For example

$$
\begin{aligned}
& =-\operatorname{tq}(1-4)^{\pi}{ }^{c \mid g} \begin{array}{c}
\pi 4 \\
2
\end{array} .
\end{aligned}
$$

Hill calculated all the terns of the expansion of $\Delta(0)$ having an crier re? aifue to less than in. The results which he cbtadred are

$$
\begin{aligned}
& 1-\operatorname{ctg} \begin{array}{c}
\pi \\
i
\end{array} \\
& { }^{+} \operatorname{sq(1-q^{2})(1-q^{\prime })} q_{i} q_{i}!
\end{aligned}
$$

$$
\begin{aligned}
& 3-\operatorname{ctg}^{-4} 2^{-}
\end{aligned}
$$

$$
\begin{aligned}
& \therefore=0 b^{-4} \\
& \text { ll,411-4:11 } 0 \cdot 0\left(\begin{array}{ll}
4 & 4^{2}
\end{array}\right)^{4 \cdot \theta_{0}}
\end{aligned}
$$

Substituting into these equations the valres: of $q, q_{1}, q_{2}, \ldots$
that correspond to the adopted value of $m$ (section 135), Hill Jbtained

The zero-order term 1.0000000030000000
The 4-order term 0.0018046110934227
The sum of 8-order terms 0.0008001808631059
The sumi of 12 -crier terms $\quad 0.900800900064478 \dot{0}$
1.0018047920240112

Judging by the law of decrease of terms of the different orders, we may conclude that the first thinteen decimals are sorrent.

Considerirg agai: squatior, ( 69 ), we find

$$
c=10715 \cdot . .3771100 .
$$

where we expect that the error in $\Delta(0)$ is transferrec to re Increasing it 2.8 times. The fact that we have found a real value for $c$ i.v $f$ great japortance. This finding suggests that the variational curve is a stable solution of equations (14). Irdeed, by constiting sections i35 and i36, we Find that the deviations $\delta N$ and $\delta_{T}$ from the motion, represented by the variational curve, will be such tiat tit the initial deviacion from the nuction along the variational curve is small, the deviai ion will rerain sral. during eny further notion only is the value of $c$ is real.

If we replace $:^{2}, q_{1}$, ... by their expressions in terms of the paraneter $m$ (section 135), wis then obtain $\triangle(0)$ as wel: as c in terms of explicit functions of $m$. In the book, quoted at the end of section 135 , Hill obtained by means of the successive approximations method

$$
\begin{aligned}
& c-1 ; m-\frac{3}{1} m=-\frac{201}{2} m^{2}-\frac{2417}{2:} m^{4}-\underset{24}{111719} m \text {. }
\end{aligned}
$$

## 139. Calculation of the Coefficients

Once the value of the fundanental constant $c$ is ohtained, the solution of equations (52) relative to the coefficients $b_{k}$ is quita simple. In oroer to ot tain the numerical values of these coefficients, it is recomended to replace the infinite system (5) , written in the form
by a finite system, obtafned by negioching ai! the negligibly small coefficients. We are ascuning thet the unknown $b_{k}$ are bounded, the magnitude of each term of this equation will then depend on the absolute value $\left|q^{2}-(2 k+i)^{2}\right|$ of the denominator and on the absolute value $|k-1|$ of the fnder of $g_{k . .}$, and will rapidly decrease when these quantities increase. Conequentiy, this proves that the unknown coefficients wili rapidly zend to zeru when the quantity $|\mathrm{k}|$ is increased. Tre easiesi manner to whain the general expression of $b_{k}$ is the following: we repince the zero row
of the determinant $\Delta(z)$, composed by the tertis (35), by the roliowing indefinite quantitics

$$
\cdot A_{2}, \lambda_{1} \lambda_{1} \lambda_{1}, \lambda_{2} .
$$

The determinant $\mathrm{D}(\mathrm{z})$ obtained by this repiacement will be convergent if the condition $\left|x_{i}\right|<A$, where $A$ is independent of $i$, is satiafipf, Fxpanding $D(z)$ by the elements of the sezo row, we find
where $B_{k}(z)$ is a meromorphic function of $z$, laving the same poles (56) that the function $f(z)$ has, with theexception of points $z= \pm q$. It is easy to see that

$$
b_{A}=\cdot T_{A}(c)
$$

Indeed, if we replace the quantity $x_{j}$ in the determinant D(c) by

$$
\begin{gathered}
4_{b}, \\
4^{:}-(2 k ; c)
\end{gathered}
$$

where $i \neq k_{1}$ and the quantity $x_{k}$ by unity, we then obtain $D(c)=0$ since the determinant will buve two equal rows in the case of $k \neq 0$, and will tend to $\Delta(c)$ in the case $k-0$ and hense $\bar{D}(c)$ will disappear. However, this replacement makes $\rho\left(c^{\prime}\right)$ ident leal to the left-hand side of the $k$ th row of equations (63). This proves our conclusion.

It is possible to show that the ccefficient $b_{k}$ will haye a multiplying factor of $m^{|2 k|-1}$. However, we shall not consjder this here.
140. The most important inequalities of the luner notion

In order to elucidate taf relation betwoon Hill's and Laplace's theories, we calculaif the loding tarms in the expansions of the quantities under consideration in powers of m. Confining vurselves to an order of terms not higher thar the thind in equ tions (i2), and taking into acccunt the values of $q^{2}, 4_{1}, q_{2}, \ldots$ obtained in section 135 , we find

$$
\begin{aligned}
& 1 \varphi=-r \mid t, i \varphi_{1} b_{1}, \psi_{1} b, \quad(1,
\end{aligned}
$$

$$
\begin{aligned}
& q^{2}-(c+2)^{2}-x-1 m^{2} 3 m^{2}+\cdots . . q^{2}-(c-2)^{2}=3 m-3 m^{2}-\ldots \\
& q^{3}-(c+f):-21-8 m+\ldots \quad . . q-(c-4)^{2}-5-5: 8 m-6 m^{2}-\ldots
\end{aligned}
$$

Neglecting the tems having an or der rif magnitude of $\mathrm{m}^{3}$, we obtain from these equations that

The corresponding Formald (51) gives the gereral solution of the equatior that defines $\delta:$ in the following forn
where $C_{1}$ and $C_{2}$ are arbitrary constants. Futing

$$
C_{1}: C_{2} \quad d \cos w, \quad i\left(C_{1}-C_{2}\right)--A \sin w_{1}
$$

where $A$ and $\omega$ are new arbitrery constarts, we obtair

$$
\begin{aligned}
& A^{-1} i N-15 m \cos [(c+2)++\cdots \mid \cdot+\cos (c \pi+\omega)-i- \\
& +\left(\begin{array}{cc}
15 & 159 \\
8 & m i \\
8 & 32
\end{array}\right) \cos |(c-2)=-i-\omega| .
\end{aligned}
$$

I: order to sinplify the following deductions, we also neglect the sincendorder terals. Then, equation (45) becomes

$$
\frac{d}{d} i r-2(1 \mid m) i N
$$

The previous expression for $\delta$ in gives

$$
i N-A \cos (c:-j-\omega): \int_{8}^{15} A m \cos ((c-2) r+f-\omega) .
$$

Hence,

$$
\text { ir: } \quad: A \operatorname{un}\left(c r \cdot u_{1}\right)-\int_{4}^{15} A m \sin ((c-2)=+\cdots l+B .
$$

where 3 is a now constant, since $C=1+m$ and $C-\%=-(1-m)$ within the degre of accuracy desired. On theother hand, we have seen in sectio: 134 that the equation of the variat ional curve is given by

$$
x=-a \cos :\left|\begin{array}{lll}
1 & m^{2} & -3 \\
4 & m^{2}, m^{2}-
\end{array}\right|, \quad y \quad a, m:\left|1: m^{2}+\frac{3}{4} m^{2} \cos ^{2}=\right| .
$$

to within terms of the crder of $\mathrm{mi}^{2}$. Hence equation (33), that defines' the angle between che rengent 10 the variational curve and the $x$-axis, yielos

$$
\dot{\psi}-\frac{\pi}{2}+\tau
$$

within an error of the order of $\mathrm{m}^{2}$. Consequently,

$$
i x=-i T \tan \div--i N \cos :, \quad l y=i T \cos :-i N \sin \tau .
$$

Bubstituting here the values of $\delta \mathrm{N}$ and $\delta$ t just foand, and errating the resulting $\delta x$ and $\delta y$ to the coordiates of the points of the variatiunai curve, we obtain

The coefficients $A$ and $B$ should be regarded as infinitesimal quantities of the first order. Hence, it is possible to put

$$
a \cos =-\beta \sin =a \cos \left(-+-\alpha r_{0}\right), \quad a \sin -+\beta \cos =a \sin \left(-f-i_{0}\right),
$$

where $\delta \tau_{0}$ is an infinitesimal constant equivalent to s . Since the origin of counting $\mathcal{C}$ is unknom, we can drop $\delta \tau^{\prime}$. This will change the previous values of $x$ and $;$ by only second order quantities relative to $A$ and $S$ To.
we shall first consider the case of $m=0$. Tn this case the rowturbation caused fy the sun is absent. The differential equations (14, defining the variational orbit are reduced to the equal ions of the twobody problem. Hence, when $\pi_{1}=0$, equation (64) must describe an eiliniticai motion in which only first powers of the eccentricity are included, However, when $\eta=0$, these equations yield

Rotating the coordinate axis by an angle of w , the nev coordinates

$$
x^{\prime} \quad \therefore \cos \omega-y \sin \omega, \quad y^{\prime} \quad \therefore \sin (1), y \cos \omega
$$

will be equal tc
comparing these expressions with the formulae of section 3 ) that define the coordinates in the elliptical motion, names y
and ronfining ourselves to first pouens of the eccentricibs, whain

$$
\text { is id, is iac, }=: \quad \text { ur } \quad \text { " }
$$

Let us now consider again the motion of the moon in the gencral cas: when $n \neq 0$. Danoting as previcusly the redius rector and the lone, itude of the moon $b y r$ and $r$, and the longitade $f$ the sun $t y v^{\prime}=r ' t+\epsilon^{\prime}$ and by putting $B=0$ and $A=$ ae, we obtain

$$
\begin{align*}
& -\frac{15}{8} \text { me } \cos (c-2)=f w \mid \cos r-c(\cos : \mid w) \cos t-t \\
& -\frac{15}{4} m e \sin |(c-2):+-\omega| \sin \tau-2 c \sin (c-\{-\omega) \sin \tau \\
& { }_{a} \sin \left(u-v^{\prime}\right)=\sin =-1-a m^{\prime}\left(1+\frac{3}{4} \cos ^{2}-\right) \sin 5 \cdots  \tag{6}\\
& -\frac{15}{8} m e \cos [(c-2)=+\omega] \sin \tau-e \cos (c \tau+\omega) \sin \tau- \\
& \left.-\frac{15}{4} m e \sin (1, c-2)=+\omega\right\} \cos t+2 e \sin (c t-1 \omega \mid \cos t .
\end{align*}
$$

Squaring these equatiors ard adoing, we obtair after evidnne nanipulations

$$
r^{2}=a^{2}\left\{\left.1-2 \cos (c,+-\omega)-2 m \cos 2 \tau-\frac{15}{4} m e \cos \right\rvert\,(c \cdot i t \cdot-(0) \mid\right.
$$

from which, aking into account that (Ser. i34)

$$
a=a\left(1-\frac{1}{i} m^{2}+\ldots\right)
$$

we cttain

$$
r \ldots a\left\{:-\frac{1}{6} m^{2}-e \cos (r-+\omega) \cdot \frac{15}{8} m e \cos |(c-?):+u|-m^{2} \cos 2 \tau\right\} .(i b)
$$

Dividing each of equations (bj) term by tein $t$, ihe cttairad wlue of $r$, we oh: aill

$$
\begin{aligned}
& \cos \left(v-\theta^{\prime}\right)=\cos \left[1-\ln _{4}^{11} \min ^{2} \sin ^{2}\right]+
\end{aligned}
$$

$$
\begin{aligned}
& \sin \left(\theta-0^{\prime}\right)=\sin 8\left[1+{ }_{4}^{11} m^{2} \cos 2 \mathrm{z}\right]- \\
& -\cos \tau\left\{\begin{array}{l}
15 \\
4 \\
4
\end{array} m e \sin [(c-9)=+\omega]-2 e \sin (c s+\infty)\right\} \text {. }
\end{aligned}
$$

from whee .s: it follows that

$$
\left.\sin \left(v-v^{\prime}-r\right)=\frac{11}{8} m^{2} \sin 2 \tau-\int_{4}^{15} m c \sin \right\rvert\,(c-2) \tau+c j+2 c \sin (c \tau+c)
$$

or, within the same accuracy, ice., to within small quantities of the second order relative to $m$ and of first order relative to $e$,

Evident.'y, the periodic terms in the expansions (66) and (67) represent the elliptical inequalities of the moon. Hence, the argument of these terms is nothing else but the mean anomaly. Keeping the notations of section 126 , the mean anomaly is equal to nt $-\sqrt{7}$. Consequently,

$$
\therefore \cdot f \omega=n t-I I .
$$

Differentiating this equation gives the motion of the perihelion:

$$
d \|=, n-c\left(n-i^{\prime}\right)=n\left(1-c \begin{array}{c}
c \\
1-1-m
\end{array}\right) .
$$

Substituting here the value of $c$ obtained in section iso, we obtain for that part of the perinelions motion, that does not depend on the eccentricities of the moon and the sun, the following equation

This series converges so slowly, that it is reconended to use the numerical method given in section 138 for the actual calculation oi the perihelion's motion. By this method, Hill found

$$
1 d \mathrm{dl}=0.008572573004864
$$

On comparing the results obtained here with those obtained by the methods developed by Laplace, de Pontecoulant and others, it is interesting to note that

$$
a_{0}=\frac{n^{\prime}}{n-n^{\prime}}=1^{\mu}-\beta^{\prime} \quad c=\frac{n}{n-n^{\prime}}, c=(1+n) c_{0}
$$

Expressing the motion of the perihelion in terms of $\mu$, we obtain

$$
\begin{aligned}
& 1 d 1 \\
& n d t=1-c={ }_{4}^{3} \mu^{2}+{ }_{32}^{225} \mu^{3}+{ }_{28}^{4071} \mu^{0}+\begin{array}{c}
265493 \\
2 i 1
\end{array} \mu^{3}+\begin{array}{c}
12822631 \\
2^{12} .3
\end{array} \mu^{6}+\ldots . .
\end{aligned}
$$

The coefficients of these series are considerably larger than the corresponding coefficients of the series developed by powers of $m$. Accordingly, it is more useful to use parameter on in the lunar thecry than parameter $\mu$. However, in the case when a particularly high degrae of accuracy is desired, it is also not recommended to use the expansion in powers of m. It is simpler and more direct to apply the numerical method as we have already pointed out in section 138.

Let us again coasider equations (66) and (6T). Since the mean longitudes of the moon and the sun are $n t+\epsilon$ and $n^{\prime} \dot{c}+\epsilon^{\prime}$ respectively, then

$$
u^{\prime}+f:=n t+8, \quad \nabla==\left(n-n^{\prime}\right) t-\beta,
$$

where $\beta=\epsilon^{\bullet}-\epsilon$ and $\tau$ is the angular distance between the nean position of the moon and the sun. Putting
where

$$
\text { r. } \quad \frac{c n_{1}^{4}}{n-n^{i}}-\infty .
$$

we represent equations (67) and (66) in the following way

$$
\begin{aligned}
& -\frac{15}{8} m e \cos [(2 n-2 n-c n) t-2 \beta \div \cdot \pi]!.
\end{aligned}
$$

We compare these expressions with those obtained in Laplace's theory particularly with formula (45) obtained in section 127. We find that Mill's theory leads to the principal terms of the equation of the centre, the variation and the evection. The advantage of this theory over Laplace's consists, first of all, in that Hill's theory allows us to calculate these inequalaitits as well. as the motion of the perihelion in a reiatively simple way and with the desired a:bitrarily high degree of accuracy for the parameter $\mu$.
141. Inequalities depending on the eccentricity of the Iunar orbit

We have studied in decail the method developed by Hill to calculate the inequalities that depend on the first power of the eccentricity of the lunar orbit. We shali now consider the inequalities that have higher powers of eccentricity as multiplying factors. This problem is equivalent to calculating the general solution of equations (14) or (13). We first consider the solution, orbits infinltely close to the variational curve, which has been found in sections 135-139. It follows from equations (31) that

$$
\begin{aligned}
& 3 N=-i x \sin 4+i y c o n+\frac{1}{2} \quad i\left(3 s \cdot e^{i f} \ldots i, \mu e^{-i t}\right) \\
& 8 T=i x \cos ?+8 y \sin \div=\frac{1}{!}\left(\text { is } \cdot e^{i t}-i-i d \cdot e^{-6}\right) \text {. }
\end{aligned}
$$

if we again put $u=x+y i$ and $=x-y i$. Ca the other hand, equations (33) yield

$$
V_{f \cdot \prime}:=i D u, \quad V_{e}^{-d t}=-i D_{s},
$$

from which it follows that

$$
\begin{equation*}
i_{\mu}=V_{V}^{D u}(i T-d M)_{0} \quad \delta_{3}={ }_{V}^{D s}\left(i ; r_{-}+i N\right) . \tag{68}
\end{equation*}
$$

In order to simplify the forthcoming deductions, we rewrite the generai expression of $\delta_{\mathrm{N}}$, given by equation (51), in the following manr.er

$$
i, V=\because!\text { S } b, x^{24} .
$$

where

$$
\therefore \quad \text { exici( }=- \text { ris }
$$

and $r_{l}$ is an arbitrary constant.
Equation (30), which can given the form

$$
v_{d:}^{d}\binom{i v}{v}=2\left(\begin{array}{l}
d_{r} \\
d r
\end{array}+m\right) B N
$$

or,

$$
\left.i V_{d=}^{d}\binom{\vdots 7}{V} \cdots\left(\begin{array}{cc}
D: u & D: s \\
D u & \cdots \\
D s
\end{array}\right] \cdot m\right) i N .
$$

enables us to concluai that $\delta \mathrm{T}$ has the same form as $\delta \mathrm{N}$ although it is necessary to take into consideration that $V$ is not an even function of Considering equation (68), we see that $\delta_{u}$ and $\delta_{s}$ are even functions of $\zeta$, expandable in the following series
that have constant coefficients. Indeed, the $\operatorname{sim} \delta_{u}+\delta_{s}=2 \delta_{x}$ should have a real quantity, while the difference $\delta_{u}-\delta_{s}$ should be an imaginary one. In this way, taking equation (16) into consideration, we can represent the solution of equations (13), that differs slightly from the variational curve, in the following manner
where $k$ runs over all the integral values from $-\infty$ to $+\infty$, while $p$ takes only the three values $-1,0$ and +1 . In particular, if $p=0$ then $A_{2 k}=a_{2 k}$.

Since the general solution differs from the variational curve by finite quantities, we can then consider that the solution just given is obtained by the leading terms of a more general expansion, representing the general solution of equatiors (13). Following Brown, we search for the general solution in the form of the same series (69) but under the condition that $p$ runs over all the integral values from $-\infty$ to $+\infty$. In doing this, we assume that the coefficients $A_{2 k+p c}$ are quantities
 esting to note, that Brown obtained expressions (69) for $u$ and $s$ by using de Pontecoulant's theory.

In order to show that there exists a solution of the type (69) for the system (13), it is necessary to be convinced in the possibility of finding such coefficients $A_{2 k+p c}$, for which the series (69) formally satisfles condittons (13). We suhstitute series (69) into equations (11), which follow from equations (13). Since, on one hand,
and on the other hand,

$$
\left.f_{1}:=\times r\right)-(: t ; 1: p c)_{0}^{-t} 1 \cdot r
$$

then the equations resulting from the substitution and the equation of coefficients of equal powers of $T_{2}$, do not change if we put $\zeta_{1}=\zeta$ ia the series (69). It is only necessary to remember that, in the final result, the quantities $\zeta^{2 k+l+p c}$ are to be replaced by

In this way, instead of substituting in equation (11) the series given by equations (69), we substitute series

We shall not repeat here the calculations carried out in section 137 but directly give the result of this substitution, which leads us to the following equating.

These equations correspond to equations (21). Starting with the above values of $c$ and $a_{2 k}$, we can find the solution of these equations by means of the successive approximations method. In this way, the part of the motion of the perihelion, that depends on the eccentricity of the lunar orbit, can be obtained. However, we are not going to go through the details of these calculations since they principally involve nothing new.

## 142. Inequalities depending on the slope of the lizar orbit

We have always assumed that the moon moves in an eelfptical plane and for this reason we have substituted $z=0$ into the equations of motion, derived in section i30. We shall now investigate the variations introduced to the lunar motion on discarding this assujmption and taking into account the slope of the lunar orbit.

Neglecting as before the eccentricity of the solar orbit and consequently putting $\Omega=0$, we obtoin from equations (8) and (10) the following equations:

$$
\begin{align*}
& D:(u s+=z)-D u \cdot D s-(1)_{-}^{2}-2 m(u D) s \quad s(J u) ; \\
& +{ }_{4}^{9} m \pi^{2}(u+s)=-3 \pi:^{2}-0  \tag{71}\\
& D\left(u D_{j}-5 D u\right)-2 m(1(: i)+\underbrace{3}_{2} m \cdot(u \quad s-1 \quad 1) \\
& D^{2} z-m^{2} z \cdots x_{i}^{3}=0 . \tag{72}
\end{align*}
$$

One of the arbitrary constants involved in the general solution of equation (72) should be chosen such, that when this constant disappears the coordinate $z$ also disappears. Denoting this constant by $X$, we assume that 2. has $\gamma$ as a multiplying factor, where $\gamma$ is a small quintity of the order of the slope of the lunar orbit.

If we neglect quantities of the order of magnitude of $\gamma^{2}$, then equations (71) will not depend on $z$ and will give the solutions that we, have already studied in detail in the previous section. Substitutirg the values of $u$ and $s$, obtained this way, into equation (72), we are able to find $z$ with an error having an order of magnitude of $\gamma^{2}$. Substituting this value of $z$ into equation (71) we obtain the terms in the expansions of $s$ and $u$ that have the same order of magnitude as $\gamma^{2}$, and so on. By means of this alternate application of equations (71) and (72), we can obtain terms having an arbitrarily high order of magnitude relative to $\gamma$

In the expressions of $m i .1$ three coordinates $u$, $s$ and $z$. Evidently, $v$ and $s$ will involve even powers of $\gamma$ while $z$ will involve odd powers.

We shall now consider in greater detail the calculation of the firgtorder terms of the coordinate $z$, assuming that the eccentricity of the lunar orbit can be set equal to zero. The corresponding solution of equations (71) is the variational curve

We substitute these values of $u$ and $s$ into equation (72). Since

$$
m^{2}+x r^{3} \cdot m^{n} \cdot \therefore \alpha\left(u^{2}+s^{2}\right)^{3}
$$

is an even function of $\mathcal{Z}$, and does not change when $\mathcal{C}$ is replaced' by $\tau^{-1}$, then

$$
\begin{equation*}
m^{2}-1-x r^{-3}-2 \sum_{k} M_{k}{ }^{*}: k \tag{73}
\end{equation*}
$$

where $M_{-k}=M_{k}$. It is easy to see that $M_{k}$ is a quantity of the same order of magnitude as $m|2 k|$, from which it follnws that equation (72) can be reduced to the form

$$
\begin{equation*}
D^{\prime \prime} z-z \because \cup M_{i} \because *=0 \tag{74}
\end{equation*}
$$

or

$$
\frac{s^{2} z}{d r^{2}}+1 \cdot\left(2 A_{0}+f 4,1 f_{1} \cos 2: \mid 1, H_{2} \cos 15+1 . .\right) z=0
$$

i.e. becomes lientical to Hill 's equation, studied in detail in sections 136-139.

Thus, the general solution of equation (74) is
where $C_{1}$ and $C_{2}$ are arbitrary constants while the characteristic exporents $g$ and $-g$ are the roots of the equation
in which the determinant $\Delta(0)$ has been obtained from the determinant $\Delta(0)$ by means of replacing $q^{2}$ by $2 M_{0}$ and $q_{k}$ by $2 M_{k}$. Finally, the coefficients $\beta_{k}$ are defined by the equations

$$
\begin{equation*}
i_{i}(g \times 2 k)=-2 \Xi, H_{k-i} \dot{\vec{r}_{i}} \cdots U_{1} \tag{77}
\end{equation*}
$$

Which correspona to equations (52).
At the end of section 134, we have seen that the following relations hold

$$
\begin{aligned}
& 2 a^{-3} \cdots 1+2 m \cdot f \cdot{ }_{2}^{3} m^{2}+\ldots . \\
& r=a\left(1-m^{2} \cos 9: 1 . . . .\right) .
\end{aligned}
$$

to Withirg terms of the order of magnitude of $m^{3}$. Consequently,

$$
m^{2}+x r^{-3} \neq 1 \pm 2: n-j m^{\pi} \mid 3 m^{2} \cos 2 x+\ldots
$$

Therefore, within the degree of accuracy desired,

$$
2 M_{0}==1+2 m \cdot f \cdot \begin{aligned}
& 5 \\
& 2
\end{aligned} m^{3}, \quad 2 M_{1}=2 M_{1}=3 m^{2}
$$

while all the other coefficients $M_{k}$ are equal to zaro. Since the coefficients $\beta \pm 2, \beta \pm 3, \ldots$ are not less than a third odder, then equations (77) yield

$$
\begin{align*}
& \left|1+2 m+\frac{5}{2} m^{2}-g^{2}\right| \beta_{0}++_{2}^{3} m: 3,{ }_{2}^{3} m^{2} \beta_{1}-0 \\
& \left\{1+2 m+\frac{5}{2} m^{2}-(g-2)^{2} \left\lvert\, \beta_{-1}+\frac{3}{2} m^{2} \beta_{11}=0\right.\right.  \tag{78}\\
& \left|1+2 m+\frac{5}{2} m^{2}-(g+\mid 2)^{2}\right| \beta_{1}+\frac{3}{2} m^{2} \beta_{0}=0
\end{align*}
$$

We can approximately obtain the value of $g$ frow the equations. Indeed the first equation shows that the foliowing relation holds

$$
g^{2}=1!a m+{ }_{2}^{5} m^{2 \cdot 1} \ldots \ldots
$$

to an error of the order of not 1 ess than $m^{3}$, from which it follows that

$$
g=1+m+{ }_{4}^{3} m^{2}+\ldots \ldots
$$

Substituting this value for $g$ into the second of equations (78), we see that the coefficient $\beta_{-1}$ oniy involves a first order terms, so that

$$
\beta_{.1}=-\frac{3}{8} m_{i_{0}}
$$

Therefore, it is of no consequence to keep terms of the second order in the expression of $\beta_{1}$. Hence we can take $\beta_{1}=0$. Thus, to an error of the order of magnttude of $\mathrm{m}^{3}$, equation (75) yields

$$
\begin{equation*}
=\quad i_{0}\left\{\cos \left(g-\left\{-z_{1}\right)-\frac{3}{8} m \cos \| g-2 x_{2}\left|-z_{1}\right|\right\}\right. \tag{79}
\end{equation*}
$$

and making use of the arbitrary nature of $\beta_{0}$, we can put

$$
C_{1}=\frac{1}{2} \operatorname{vxp}\left(i i_{1}\right), \quad c_{2}=\frac{1}{2} \operatorname{rxp}\left(-i i_{1}\right)
$$

The firet term in formula (\%) evidently cotresponds to the unperturbed motion. The second term is called the evection of the latitutde.

Denoting by $i$ the slope of the lunar orbit, and by $\Omega$ the longitude of the ascending rode, and using the arguments in section 119 , we write

$$
z=x r \sin (v-4)
$$

where $\gamma=t g 1$. Sirce,
then, neglecting terms of order of magnitude of $\mathbf{w}^{2}$, we obtain

Let us take the derivatives with respect to $t$ on both sides of the last equation. Then

$$
K\left(\begin{array}{ll}
1 & n
\end{array}\right) \cdot n \quad d!
$$

from which it follows that

$$
d u=n\left(1-1 \begin{array}{l}
g \\
d t
\end{array}\right)
$$

Since, in Laplace's theory, we have put

$$
\frac{n 1}{d 1}=n(1 \cdots n) .
$$

then

$$
\stackrel{\&}{\stackrel{g}{\perp} m}
$$

Equation (76) enables us to obtain thevalue of $g$ and consequently that of $g$, with an arbitrary degree of accuracy, The leading terns in the expansion of $g$ in powers of mare

If a very high degree of accuracy is required, then it is recommended not to expand the coefficients $g$ and $\beta_{K}$ in powers of $m$, but to directly calculate their values numerically. In this case, it is better to obtain the expansion of the function (73) by calculating separate values for $r$. Using the aboverinentioned expansion of the function $r / a \cos v$ and $r / a \sin v$,

Rill obtained the following specific values for the function


Using the conventional formulae of gamic analysis, we obtain from the previous value

Adam ${ }^{(1)}$ was the first to obtain the value of the quantity $g$, that characterises the translational motion of the node of the lunar orbit but means of this method.
(1) J.C. Adams, On the Motion of the Moon's node in the cense when the Orbits of the Sur: and Mon are supposed to have no eccentricities etc., Monthly Notices R.A.S., 38, i877, 43-49. Coil. Works, 181-188.

Table. I
The coefficients $C_{k}^{n, m}$ of the expansion in powers of the accentricity (sec. $8^{n}$ )


Tab1e I - (continued)


Table I - (continued)


Table I- (continued)


Table II
OREANAL PAGE
OR FOOR QUALIIX
The coefficients $\mathrm{s}_{k}^{\mathrm{n}, \mathrm{m}}$ of the expansion in powers of the eccentricity (sec. 82)


Table II- (continued)


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Table II- (continued)


Table LI- (continued)


## Tab1e III

The Enke function and its logarithm（section 96）

| 4. | 4－0 | $0 \cdot 11$ | $9>0$ | $4 \cdot 6$ | 11 |
| :---: | :---: | :---: | :---: | :---: | :---: |
| 061） <br> 0（n）？ <br> 11（x13 <br> l）．（1）：7 |  |  |  |  |  |
|  | $3 \ln$（n） | 3 （0）0， | 0.17712 | $0.17712 \mathrm{l}(4)$ | O（x） |
|  | ：2xis．${ }^{\text {a }}$ | 30.57 | 11 istur 10x | $\begin{array}{lll}0.4 & (821 & 1(4) \\ 10.4\end{array}$ | Uivi |
|  | $2450!$ if |  | 析 | $1.8741 .60104$ | 0 nic |
|  | ：97i7 7 | $30^{\prime 2} 77^{7}$ | $0.47^{10.7}{ }^{1 / 8}$ |  | 0 （0）S |
|  |  | $3^{1013} 70$ | 04：2＊）${ }^{107}$ | $114814{ }^{114}$ | 0.101 |
|  | 71 | 76 | （1b） | 110 |  |
| 11.1405 | 2！ $0_{0} 0^{4} 7$ | 30i79 7 | $017172$ | 0.96 .54 | 0．00） |
| 000 ； | ？9．5xi 7 | $001 I_{i} i$ | ＂1iocs | 0．43．608 $8^{110}$ | a， 0 ar |
| $1 \cdot 1007$ | $2: 31+37$ | $3 . は 23$ | 11.101 .8 | $0.18 .178^{110}$ | 11.007 |
| （1）n9 | 291117 | $\begin{aligned} & 30.11 \\ & 3.00 \mathrm{j}=17 \end{aligned}$ | $\begin{aligned} & 0.11851 \\ & 0117711 \end{aligned}$ |  | 11108 |
| lolues | 2 以 ${ }^{3}$ |  |  |  | 0 0，9 |
|  |  | 79 | 106 | 111 |  |
| 0010 |  |  |  | 11． $18 \times 10$ | 1010 |
| 0111］ | $\because 11!k i$ | $30 \times 41$ | $0.1 \text { csis }$ | U－4bi21 111 | 01111 |
| （1112 | $241: \therefore$ | $3092{ }^{\circ} 79$ | 11．41 4：3 | U．49！5： | 0012 |
| $\begin{aligned} & 11,11: 3 \\ & 0021 \end{aligned}$ | $\begin{aligned} & 24 . i 1 \\ & \therefore x^{\prime} 14: \end{aligned}$ | $3 \text { lun; }$ | $0.41: 1.20$ | 0.491811．4325， 112 | 0101.1 |
|  |  |  | $\text { 4. } 41,: 15$ |  | 00114 |
|  | 10 | 31 | 11.4 | 117 |  |
| 1015 |  |  | 0.4ilive | 0．41．ids 112 | 10.015 |
| 0.116 | $234 s$ | $31217$ | $0.1(\alpha N 1$ | U．13180 | 0010 |
|  |  |  |  |  | O） |
| 19117 | $2 \times 171$ | $3.3 i^{8 t}$ |  | 0．19．0． 413 | 0017 |
| 108 ．10；y | $\begin{aligned} & \because 4 i{ }^{\prime} \\ & : \operatorname{Guj} \end{aligned}$ | $3 \mid 100$ | $01.18!6$ | 0497w， | ． 1018 |
|  |  |  | （1）M品 <br> lo＇s | $0.141 / 1!115$ | （1．11） |
|  | （i） | 42 |  | 11.5 |  |
| 00.0 | 2 ainit ix | 1157．：$\times 3$ | 01，M6 lu＇d | 0．194．32 114 | 11.020 |
| （10．！！ | Exfmis |  | 14，14．9 10． | 13．druls ${ }^{114}$ | coll |
| （1）．0：$:$ | $2+131^{1.7}$ | 3.17390 | 0．1737，101 | Usunis 1lis | （） 12 |
|  |  | ．18．188 | 0.4 .275104 | （1）Sunt 115 |  |
| OHIS | $\because$－16 |  | 0．4．27！ 10.3 | 1）$)^{(271}$ 114 | 111．3 |
| 0102 | 2HE＊＊ | 311807 | $0.1317{ }^{10.7}$ | 1． 51.598 | U $1: 1$ |
|  | 67 | 81 | 10.5 | 115 |  |
| 10：3： | ． 1.241 | 3．lirsi | 0．4irwnit | 0． $0^{\text {cupdo }}$ | 0．125 |
| U．1\％${ }^{\text {a }}$ | $: 8162^{17}$ | 3.2 Nit ${ }^{\text {RS }}$ | $0.11 \times i^{10.3}$ | O5 $0^{1 / 115}$ | U．1231） |
| 0127 | $2.40 \cdot 137$ | ．2161 45 | 01186310.3 | $0_{1.517 .143} 115$ | 110！ |
|  | － C6 $^{\text {c }}$ | $30.14 \times 5$ | 1903 10． | uresis 11：4 | U，19 |
| 10.8 | －$\lambda(1.9$ ， | 32.46 | 0．4．4． $10 \%$ |  | 011.5 |
| 110．： | $27^{4} \times 1$. | 32 LT | $0.11 w^{\prime}$ | USALIM， 1 | 11，049 |
|  | 18. | $\kappa 7$ | 102 | 116 |  |
| 1103 | 2 in 0 | 3219 | $0+1557$ | S310，0 | ＂UW） |

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Table IV
Different constants used in computing the perturbations
Value logarithm
The radius of the arc in degrees The radius of the arc in minutes The radius of the arc in seconds

| The gaussian constant | ( $\mathbf{k}$ |
| :---: | :---: |
|  | ( |
|  | ( k' |
|  | ( 2 |
|  | ( $\mathrm{k}^{2}$ |



35.\& 2i.us א:4
5.114 +2:13.3
$82355514411 \quad 10$
3550 a $4 \times 1$
6.471 16.5-10



[^0]:    (1) Banachiewioz indicated th existance of these equations (T. Banachiewicz, Sur quelques points fundamontaux de la theorie des orbites, Acta astronomica, Ser., a, 3, 1933).

[^1]:    (1) In this chapter, we shalil keep as much as possible Leverifer's notations. We only note that he denotes the quantities J, , , and by , , and respectively.

[^2]:    (1) Comptes rendus de 1 'Academie des Science de Paris, 58 (1369);

    72 (187i).

[^3]:    

