## PREFACE

In the auricular structure introduced by this University for students of Post- Graduate degree programme, the opportunity to pursue Post-Graduate course in Subject introduced by this University is equally available to all learners. Instead of being guided by any presumption about ability level, it would perhaps stand to reason if receptivity of a learner is judged in the course of the learning process. That would be entirely in keeping with the objectives of open education which does not believe in artificial differentiation.

Keeping this in view, study materials of the Post-Graduate level in different subjects are being prepared on the basis of a well laid-out syllabus. The course structure combines the best elements in the approved syllabi of Central and State Universities in respective subjects. It has been so designed as to be upgradable with the addition of new information as well as results of fresh thinking and analysis.

The accepted methodology of distance education has been followed in the preparation of these study materials. Co-operation in every form of experienced scholars is indispensable for a work of this kind. We, therefore, owe an enormous debt of gratitude to everyone whose tireless efforts went into the writing, editing and devising of a proper lay-out of the materials. Practically speaking, their role amounts to an involvement in invisible teaching. For, whoever makes use of these study materials would virtually derive the benefit of learning under their collective care without each being seen by the other.

The more a learner would seriously pursue these study materials the easier it will be for him or her to reach out to larger horizons of a subject. Care has also been taken to make the language lucid and presentation attractive so mat they may be rated as quality selflearning materials. If anything remains still obscure or difficult to follow, arrangements are there to come to terms with them through the counselling sessions regularly available at the network of study centres set up by the University.

Needless to add, a great deal of these efforts is still experimental-in fact, pioneering in certain areas. Naturally, there is every possibility of some lapse or deficiency here and there. However, these do admit of rectification and further improvement in due course. On the whole, therefore, these study materials are expected to evoke wider appreciation the more they receive serious attention of all concerned.

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Sixth Reprint : December, 2017

Printed in accordance with the regulations of the Distance Education Bureau of the University Grants Commission.

## Subject : Mathematics

Post Graduate

## Paper : PG (MT) : IX A(I)

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## Notification

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## Unit 1 analytic Continuation

## Structure

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### 1.0 Objectives of this Chapter

In this chapter we shall introduce the idea of direct analytic continuation of an analytic function. The concepts of analytic continuation by means of power series, complete analytic function, natural boundary, analytic continuation along a curve will be explained with the help of examples. Homotopic curves, analytic continuation of multi-valued function and Monodromy theorem will also be discussed.

### 1.1 The idea of analytic continuation

The idea of analytic continuation rests on the notion of analytic function. A function $f(z)$ is analytic at $z=z_{0}$ if it is differentiable in some $\in$-neighbourhood of $\mathrm{z}_{0}$ or, equivalently if it can be expressed in the form of a Taylor series in a neighbourhood of that point. The domain of convergence of this power series will be the region of analyticity of the function $f(z)$.

Following Uniqueness Theorem : "If two functions $f(z)$ and $g(z)$, analytic on a region $D$, are such that $f(z)=g(z)$ on a set $A \subset D$ having a limit point in $D$, then $f(z)$ $=\mathrm{g}(\mathrm{z}) \forall \mathrm{z} \in \mathrm{D}$," we know that if two analytic functions agree in some small neighbourhood of a point situated in their common region of analyticity D , they
coincide everywhere in D . We first introduce the idea of analytic continuation by the following examples.

The geometric series

$$
1+z+z^{2}+\ldots
$$

converges for $|z|<1$ and its sum function $g(z)=\frac{1}{1-z}$ is an analytic function for $|z|<1$.

The geometric series diverges for $|z| \geq 1$.
However, the function

$$
h(z)=\frac{1}{1-z}
$$

is analytic for all z except $\mathrm{z}=1$. But we observe that

$$
\mathrm{h}(\mathrm{z})=\mathrm{g}(\mathrm{z}) \forall \mathrm{z} \in\{|\mathrm{z}|<1 \cap \mathbb{C} \backslash\{1\}\}
$$

Thus, we may regard $h(z)$ as determining an analytic continuation of $g(z)$ from the domain $|z|<1$ into the domain $\mathbb{C} \backslash\{1\}$.

Example 1.1 Consider the Laplace transform of 1 in the $z$-plane,

$$
F(z)=£\{1\}(z)=\int_{0}^{\infty} e^{-z t} d t=\frac{1}{z} \text { for Re } z>0
$$

We introduce a function

$$
\phi(\mathrm{z})=\frac{1}{\mathrm{z}}
$$

which is analytic in the complex plance $\mathbb{C}$ except the origin. Here

$$
\phi(\mathrm{z})=\mathrm{F}(\mathrm{z}) \forall \mathrm{z} \in \mathbb{C} /(0) \cap \operatorname{Re} \mathrm{z}>0
$$

and we consider $\phi(z)$ as analytic continuation of $F(z)$ from the domain $\operatorname{Re} z>0$ into the complex plane with the point $\mathrm{z}=0$ deleted.

We put these ideas more precisely in the following discussion.

### 1.2 Direct analytic continuation

Let (i) $f(z)$ and $g(z)$ be analytic functions on domains $D_{1}$ and $D_{2}$ respectively.
(ii) $\mathrm{D}_{1} \cap \mathrm{D}_{2} \neq \phi$
(iii) $f(z)=g(z)$ for all $z$ belonging to $D_{1} \cap D_{2}$

Then $g(z)$ is called a direct analytic continuation of $f(z)$ to $D_{2}$, and vice versa.

Theorem 1.1. A direct analytic continuation, if it exists, is unique.
Proof. Let $\mathrm{f}(\mathrm{z})$ be an analytic function with domain of definition $\mathrm{D}_{1}$ and let $\mathrm{g}(\mathrm{z})$, another analytic function with domain of definition $\mathrm{D}_{2}$, be its direct analytic continuation. We shall show that $\mathrm{g}(\mathrm{z})$ is unique. On the contrary suppose $\phi(\mathrm{z})$ be another analytic continuation of $f(z)$ into $D_{2}$. Then

$$
\mathrm{f}(\mathrm{z})=\mathrm{g}(\mathrm{z}) \text { for all } \mathrm{z} \in \mathrm{D}_{1} \cap \mathrm{D}_{2}
$$

Also, $f(z)=\phi(z)$ for all $z \in D_{1} \cap D_{2}$


Fig. 1
and so $\phi(z)$ coincides with $g(z)$ in $D_{1} \cap D_{2}$. Thus we have, by the Uniqueness theorem, $\phi(\mathrm{z})=\mathrm{g}(\mathrm{z})$ in $\mathrm{D}_{2}$.

### 1.3 Analytic continuation of elementary functions

The functions $\mathrm{e}^{\mathrm{z}}, \sin \mathrm{z}, \cos \mathrm{z}, \sinh \mathrm{z}$ etc are already known to us. These functions are regular in the entire complex plane. Let us assume, by definition, that

$$
\mathrm{e}^{\mathrm{z}}=\sum_{\mathrm{n}=0}^{\infty} \frac{\mathrm{z}^{\mathrm{n}}}{\mathrm{n}!}
$$

and observe that it coincides with $\mathrm{e}^{\mathrm{x}}$, known earlier, for real values of z . Thus we can take $\mathrm{e}^{\mathrm{z}}$ as the analytic continuation of $\mathrm{e}^{\mathrm{x}}$ from real axis into the entire complex plane. Likewise introducing $\sin \mathrm{z}, \cos \mathrm{z} \sinh \mathrm{z}, \cosh \mathrm{z}$ in the form of power series-

$$
\begin{aligned}
& \sin z=\sum_{n=0}^{\infty} \frac{(-1)^{n} z^{2 n+1}}{(2 n+1)!}, \cos z=\sum_{n=0}^{\infty} \frac{(-1)^{n} z^{2 n}}{(2 n)!} \\
& \sinh z=\sum_{n=0}^{\infty} \frac{z^{2 n+1}}{(2 n+1)!} \text { and } \cosh z=\sum_{n=0}^{\infty} \frac{z^{2 n}}{(2 n)!}
\end{aligned}
$$

We can treat them as the analytic continuation of the functions $\sin x, \cos x, \sinh x$ and cosh x respectively from the real axis into the entire complex plane.

### 1.4 Analytic continuation by power series

We now explain the concept of analytic continuation by means of power series.
Suppose the initial function $f_{1}(z)$ is analytic at a point $z_{1}$. Then $f_{1}(z)$ can be represented by its Taylor series about $\mathrm{z}_{1}$ as
$f_{1}(z)=\sum_{n=0}^{\infty} a_{n}\left(z-z_{1}\right)^{n} \ldots(1)$, where $a_{n}=\frac{f_{1}^{(n)}\left(z_{1}\right)}{n!}$
The circle of convergence $\gamma_{1}$ of the series (1) is given by
$\gamma_{1}:\left|z-z_{1}\right|=R_{1}$, where
$\frac{1}{R_{1}}=\lim _{n \rightarrow \infty} \sup \left|a_{n}\right|^{\frac{1}{n}}$

Let $D_{1}=\left\{\mathrm{z}:\left|\mathrm{z}-\mathrm{z}_{1}\right|<\mathrm{R}_{1}\right\}$. Then

$f_{1}(z)$ is analytic in $D_{1}$. We draw a curve $\gamma$ from $z_{1}$ and perform analytic continuation along $\gamma$ as follows :

We take a point $\mathrm{z}_{2}$ on $\gamma$ such that the $\operatorname{arc} \overparen{\mathrm{z}_{1}} \mathrm{z}_{2}$ lies inside $\gamma_{1}$.
We then compute the values $f_{1}\left(z_{2}\right), f_{1}{ }^{1}\left(z_{2}\right), \ldots, f_{1}{ }^{(n)}\left(\mathrm{z}_{2}\right)$ by successive term by term differentiation of the series (1) and write

$$
f_{2}(z)=\sum_{n=0}^{\infty} b_{n}\left(z-z_{2}\right)^{n} \ldots \text { (2) where } b_{n}=\frac{f_{1}^{(n)}\left(z_{2}\right)}{n!}
$$

The circle of convergence $\gamma_{2}$ of the series (2) is given by

$$
\gamma_{2}:\left|z-z_{2}\right|=R_{2} \text {, where } \frac{1}{R_{2}}=\lim _{n \rightarrow \infty} \sup \left|b_{n}\right|^{\frac{1}{n}}
$$

Let $D_{2}=\left\{z:\left|z-z_{2}\right|<R_{2}\right\}$. Then $f_{2}(z)$ is analytic in $D_{2}$. By uniqueness theorem, $f_{1}(z)=f_{2}(z)$ for all $z \in D_{1} \cap D_{2}$. If $\gamma_{2}$ extends beyond $\gamma_{1}$, then $f_{2}(z)$ gives an analytic continuation of $f_{1}(z)$ from $D_{1}$ to $D_{2}$. Similarly, considering a point $z_{3}$ on $\gamma$ such that
the $\operatorname{arc} \overparen{z_{2}} z_{3}$ lies inside $\gamma_{2}$, we get an analytic function $f_{3}(z)$ in a circular domain $D_{3}$ such that $f_{2}(z)=f_{3}(z)$ for all $z \in D_{2} \cap D_{3}$. If $D_{3}$ extends beyond $D_{2}$, then $f_{3}(z)$ gives an analytic continuation of $f_{2}(z)$ from $D_{2}$ to $D_{3}$. Repeating this process we get a number of different power series representing analytic functions $f_{i}(z)$ in their respective circular domains $D_{i}$ which form a chain of analytic continuations of the original function $f_{1}(z)$ such that $\left(f_{i}, D_{i}\right)$ is a direct analytic continuation of $\left(f_{i-1}, D_{i-1}\right)$.

Note : We may obtain the series (2) from the series (1) in the following way :
We rewrite the series (1) in the form : $\sum_{\mathrm{n}=0}^{\infty} \mathrm{a}_{\mathrm{n}}\left\{\left(\mathrm{z}-\mathrm{z}_{2}\right)+\left(\mathrm{z}_{2}-\mathrm{z}_{1}\right)\right\}^{\mathrm{n}}$
Using binomial theorem we then expand $\left\{\left(\mathrm{z}-\mathrm{z}_{2}\right)+\left(\mathrm{z}_{2}-\mathrm{z}_{1}\right)\right\}^{\mathrm{n}}$ and collect the terms in like powers of $\left(\mathrm{z}-\mathrm{z}_{2}\right)$ and obtain the series (2).

We give two examples.
Example 1.2 The function

$$
\mathrm{f}(\mathrm{z})=\frac{1}{1+\mathrm{z}^{2}}
$$

possesses two simple poles at $\mathrm{z}= \pm \mathrm{i}$; Otherwise it is regular throughout the whole complex plane. We first choose a point, say $z=0$ at which $f(z)$ is analytic and obtain its Taylor series expansion represented by $g(z)$ as

$$
\mathrm{g}(\mathrm{z})=1-\mathrm{z}^{2}+\mathrm{z}^{4}-\ldots,|\mathrm{z}|<1
$$

The series fails to converge on and beyond the unit circle, as is clear from the series for $\mathrm{z}=1$ even though the function $\mathrm{f}(\mathrm{z})$ is


Fig. 2 analytic at that point. We can in fact continue the expansion from one region to another. Let us consider a second expansion of $f(z)$, this time about a point $\mathrm{z}=\frac{3}{4}$ lying inside the unit circle (i.e. lying inside the region of convergence of the former series). We form the expansion as follows

$$
\frac{1}{1+z^{2}}=\frac{1}{(z+i)(z-i)}=\frac{1}{2 i}\left(\frac{1}{z-i}-\frac{1}{z+i}\right)
$$

$$
\begin{align*}
& =\frac{1}{2 \mathrm{i}}\left\{\frac{1}{\mathrm{z}-\frac{3}{4}+\frac{3}{4}-\mathrm{i}}-\frac{1}{\mathrm{z}-\frac{3}{4}+\frac{3}{4}+\mathrm{i}}\right\} \\
& =\frac{1}{2 \mathrm{i}}\left[\frac{1}{\frac{3}{4}-\mathrm{i}}\left(1+\frac{\mathrm{z}-3 / 4}{3 / 4-\mathrm{i}}\right)^{-1}-\frac{1}{\frac{3}{4}+\mathrm{i}}\left(1+\frac{\mathrm{z}-3 / 4}{3 / 4+\mathrm{i}}\right)^{-1}\right] \\
& =\frac{1}{2 \mathrm{i}}\left[(3 / 4-\mathrm{i})^{-1}\left\{1-(\mathrm{z}-3 / 4) /(3 / 4-\mathrm{i})+(\mathrm{z}-3 / 4)^{2} /(3 / 4-\mathrm{i})^{2}-\ldots\right\}\right. \\
& \left.-(3 / 4+\mathrm{i})^{-\mathrm{i}}\left\{1-(\mathrm{z}-3 / 4) /(3 / 4+\mathrm{i})+(\mathrm{z}-3 / 4)^{2} /(3 / 4+\mathrm{i})^{2}-\ldots\right\}\right],\left|\mathrm{z}-\frac{3}{4}\right|<\frac{5}{4} \\
& =\frac{16}{25}-\frac{3}{2}\left(\frac{16}{25}\right)^{2}\left(\mathrm{z}-\frac{3}{4}\right)+\frac{11}{16}\left(\frac{16}{25}\right)^{3}\left(\mathrm{z}-\frac{3}{4}\right)^{2}+\frac{21}{16}\left(\frac{16}{25}\right)^{4}\left(\mathrm{z}-\frac{3}{4}\right)^{4} \tag{2}
\end{align*}
$$

We denote this expansion by $\mathrm{h}(\mathrm{z})$, which converges in the right-hand circle $\left|\mathrm{z}-\frac{3}{4}\right|<\frac{5}{4}$ and coincides with $\mathrm{g}(\mathrm{z})$ in the shaded region. We see that $\mathrm{h}(\mathrm{z})$ is clearly a direct analytic continuation of $g(z)$.

Let us construct another analytic continuation of $g(z)$. Now we consider a neighbourhood of the point $\mathrm{z}=1$ (though it is a boundary point of the unit circle the function $\mathrm{f}(\mathrm{z})$ is analytic there) and obtain an


Fig. 3 expansion represented by

$$
\begin{align*}
& \phi(z)=\frac{1}{2}-\frac{1}{2}(z-1)+\frac{1}{4}(z-1)^{2}-\ldots \\
& \text { for }|z-1|<\sqrt{2} \ldots \tag{3}
\end{align*}
$$

In this way we can determine all possible direct analytic continuations of $g(z)$ and then continuations of these continuations and so on. A complete analytic function is defined as consisting of the original function and the collection of all the continuations so achieved.

Here the complete analytic function is $\frac{1}{1+z^{2}}$, defined in the whole complex plane barring the points $\mathrm{z}= \pm \mathrm{i}$.

Example 1.3 Consider the function

$$
\mathrm{f}(\mathrm{z})=\frac{1}{1+\mathrm{z}}
$$

Clearly this function is analytic everywhere except at $\mathrm{z}=-1$. We take a function

$$
\begin{equation*}
\phi(z)=1-z+z^{2} \tag{4}
\end{equation*}
$$

Then sum function $\phi(z)$ is $\frac{1}{1+z}$ in $|z|<1$. Take a point $z=-1 / 4$ inside the region


Fig. 4


Fig. 5 of convergence of $\phi(\mathrm{z})$ and in a neighbourhood of this point we determine

$$
\begin{align*}
& \Psi(\mathrm{z})=\frac{4}{3}\left\{1-\frac{4}{3}\left(\mathrm{z}+\frac{1}{4}\right)+\left(\frac{4}{3}\right)^{2}\left(\mathrm{z}+\frac{1}{4}\right)^{2}-\ldots\right\} \\
& \left|\mathrm{z}+\frac{1}{4}\right|<\frac{3}{4} \tag{5}
\end{align*}
$$

It can be checked easily that $\phi(\mathrm{z})$ and $\Psi(\mathrm{z})$ are direct analytic continuation of each other.

Again in the neighbourhood of $z=i / 2$ we obtain an expansion

$$
\mathrm{k}(\mathrm{z})=\frac{1}{1+\mathrm{i} / 2}\left[1-\left(\frac{\mathrm{z}-\mathrm{i} / 2}{1+\mathrm{i} / 2}\right)+\left(\frac{\mathrm{z}-\mathrm{i} / 2}{1+\mathrm{i} / 2}\right)^{2}-\ldots\right]
$$

$$
\begin{equation*}
\left|z-\frac{i}{2}\right|<\frac{\sqrt{5}}{2} \tag{6}
\end{equation*}
$$

In performing analytic continuations we notice that there are certain points which always lie on the boundary of domains in which expansions are not valid. These points are nothing but the singularities of the complete analytic function. In example 1.2 these are $\mathrm{z}= \pm \mathrm{i}$ whereas it is $\mathrm{z}=-1$ for example 1.3.

## Regular and Singular points

Let $f(z)$ be an analytic function defined in the domain D , bounded by a simple closed curve $\Gamma$. A point $\varsigma \in \Gamma$ is called a regular point of the function $f(z)$ if there exist a neighbourhood $|\mathrm{z}-\varsigma|<\epsilon$ of the point $\varsigma$ and an analytic function $\phi_{\varsigma}(\mathrm{z})$ such that $\phi_{\varsigma}(\mathrm{z})=\mathrm{f}(\mathrm{z}) \forall \mathrm{z} \in \mathrm{D} \cap|\mathrm{z}-\varsigma|<\epsilon$.

The boundary point $\zeta$ which is not a regular


Fig. 6
point is called a singular point of $f(z)$ i.e., in any neighbourhood of the point $\zeta$, there cannot be any analytic function coinciding with $f(z)$ in the part common to the neighbourhood of $\zeta$ and the domain $D$.

## Natural boundary

In examples 1.2 and 1.3 we have encountered with finite number of singular points situated on the boundary of the region of analyticity of the given function. It might happen that the boundary is dense with singular points. In this case analytic continuation across the boundary of the region is not possible. Such a boundary is called a natural boundary.

Example 1.4 Test whether analytic continuation of the function $f(z)=\sum_{n=0}^{\infty} z^{2^{n}}$ is possible outside its circle of convergence.

Solution : Applying the ratio test we find that the given series

$$
\begin{equation*}
f(z)=z+z^{2}+z^{4}+z^{8}+\ldots \tag{7}
\end{equation*}
$$

converges for $|z|<1$. The point $z=1$ is a singular point of $f(z)$ as it is seen for real z that the $\operatorname{sum} \sum_{\mathrm{n}=0}^{\infty} \mathrm{x}^{2^{n}}$ increases indefinitely as $\mathrm{x} \rightarrow 1$. Now to test whether the circle of convergence, the unit circle, is a natural boundary we examine the behaviour of the given function at the points.

$$
\mathrm{z}_{\mathrm{k}, \mathrm{~s}}=\mathrm{e}^{\frac{\mathrm{i} 2 \pi}{2^{\mathrm{k}} \mathrm{~s}}}, \mathrm{~s}=1,2,3, \ldots 2^{\mathrm{k}}
$$

( $k$ is any natural number). For this sake we consider the points $\tilde{z}_{k . s}=\mathrm{re}^{\frac{\mathrm{i} 2 \pi}{2^{\mathrm{k}} s}}$ $0<r<1$ and evaluate $f(z)$ at these points.

Then $f\left(\tilde{z}_{k, s}\right)=\sum_{n=0}^{k-1} r^{2^{n}} e^{\frac{i 2 \pi}{k^{k} s .2^{n}}}+\sum_{n=k}^{\infty} r^{2^{n}} e^{\frac{i 2 \pi}{2^{k^{s}} s 2^{n}}}$
and observe that the first term consists of a finite number of terms and hence bounded in absolute value, whereas the second term is absolute value reduces to $\sum_{n=k}^{\infty} r^{2^{n}}$. Clearly this sum increases indefinitely as $r \rightarrow 1$. This shows that the points $\mathrm{z}_{\mathrm{k}, \mathrm{s}}$ (as $\lim _{\mathrm{r} \rightarrow 1} \tilde{\mathrm{z}}_{\mathrm{k}, \mathrm{s}}=\mathrm{z}_{\mathrm{k}, \mathrm{s}}$ are singular points of the


Fig. 7
given function $f(z)$. Now as $k \rightarrow \infty$ these points form an everywhere dense set of points on the boundary of the unit circle. Thus analytic continuation outside the circle of convergence of the given function is not possible.

Example 1.5 Show that the function $f(z)=\sum_{n=1}^{\infty} z^{n!}$ has unit circle as its natural boundary.

Theorem 1.2 Every power series has at least one singular point on its circle of convergence.

Proof. Let $\mathrm{f}(\mathrm{z})=\mathrm{a}_{0}+\mathrm{a}_{1}\left(\mathrm{z}-\mathrm{z}_{0}\right)+\mathrm{a}_{2}\left(\mathrm{z}-\mathrm{z}_{0}\right)^{2}+\ldots$ be any power series with region of convergence $K:\left|z-z_{0}\right|<R$. We shall have to prove there lies at least one singular point on the circle of convergence $\Gamma:\left|z-z_{n}\right|=R$ of the function. Suppose, on the contrary, that every point on $\Gamma$ are regular points. Let $\varsigma_{1}, \varsigma_{2}, \ldots \varsigma_{i}, \ldots$ be certain number of regular points belonging to $\Gamma$ and $\mathrm{N}\left(\varsigma_{1}\right), \mathrm{N}\left(\varsigma_{2}\right), \ldots, \mathrm{N}\left(\varsigma_{\mathrm{i}}\right) \ldots$ be their neighbourhoods respectively. The points $\varsigma_{i}$ 's are chosen in such a way that $\mathrm{N}\left(\varsigma_{\mathrm{i}}\right)$ has non null intersection with $N\left(\varsigma_{i-1}\right)$ and $N\left(\varsigma_{i+1}\right)$ and the union of these neighbourhoods completely cover the boundary $\Gamma$. Let


Fig. 8 D be the union of K and all these neighbourhoods $\mathrm{N}\left(\varsigma_{\mathrm{i}}\right)$. D is open since K and every $\mathrm{N}\left(\varsigma_{\mathrm{i}}\right)$ are open. D is also connected since.
(i) any two points lying in $\mathrm{K} \subset \mathrm{D}$ can be connected by a straight line segment lying in $K$, since K is connected.
(ii) one point $\mathrm{z}_{1} \in \mathrm{~N}\left(\varsigma_{1}\right)$ and the other $\mathrm{z}_{2} \in \mathrm{~K}$ can be connected by two straight line segments $\overline{z_{1} \zeta_{1}}$ and $\bar{\zeta}_{1} z_{2}$ lying within $N\left(\varsigma_{1}\right) U K \subset D$.
(iii) one point $\mathrm{z}_{\mathrm{m}} \in \mathrm{N}\left(\zeta_{\mathrm{m}}\right)$ and $\mathrm{z}_{\mathrm{n}} \in \mathrm{N}\left(\zeta_{\mathrm{n}}\right)$ can be connected by a curve consisting of $\overline{\mathrm{z}_{\mathrm{m}} \zeta_{\mathrm{m}}}+\zeta_{\mathrm{m}} \zeta_{\mathrm{n}}+\overline{\zeta_{\mathrm{n}} \mathrm{z}_{\mathrm{n}}} \subset \mathrm{D} \quad$ since $\quad \overline{\mathrm{z}_{\mathrm{m}} \zeta_{\mathrm{m}}} \subset \mathrm{N}\left(\zeta_{\mathrm{m}}\right) \subset \mathrm{D}, \zeta_{\mathrm{m}} \overline{\zeta_{\mathrm{n}}} \subset \Gamma \subset \mathrm{D} \quad$ and $\overline{\zeta_{\mathrm{n}} \mathrm{Z}_{\mathrm{n}}} \subset \mathrm{N}\left(\zeta_{\mathrm{n}}\right) \subset \mathrm{D}$.
and finally if two points lie in the same neighbourhood $\mathrm{N}\left(\zeta_{\mathrm{i}}\right)$ it is always connected by a curve $\gamma \subset \mathrm{N}\left(\zeta_{\mathrm{i}}\right) \subset \mathrm{D}$. Now we introduce an analytic function $\psi(\mathrm{z})$ on the open connected set D which satisfies

$$
\begin{gathered}
\psi(\mathrm{z})=\phi_{\varsigma_{\mathrm{i}}}(\mathrm{z}), \mathrm{z} \varepsilon \mathrm{~N}\left(\zeta_{\mathrm{i}}\right) \\
\mathrm{f}(\mathrm{z}), \mathrm{z} \varepsilon \mathrm{~K}
\end{gathered}
$$

where $\phi_{\zeta_{i}}(z)$ is a direct analytic continuation of $f(z)$ in the neighbourhood $N\left(\zeta_{\mathrm{j}}\right)$ of the regular point $\zeta_{\mathrm{i}}$.

We now prove that $\psi(\mathrm{z})$ is well-defined on D . Let $\alpha, \beta$ be any two points on $\Gamma$ such that $H=N(\alpha) \cap N(\beta) \neq \phi$ and since $\alpha, \beta$ are regular points there exist functions $\phi_{\alpha}(z)$ and $\phi_{\beta}(z)$ as direct analytic continuations of $f(z)$ in $N(\alpha)$ and $N(\beta)$ respectively i.e.

$$
\begin{aligned}
& \phi_{\alpha}(\mathrm{z})=\mathrm{f}(\mathrm{z}) \forall \mathrm{z} \varepsilon \mathrm{~N}(\alpha) \cap \mathrm{K} \\
& \phi_{\beta}(\mathrm{z})=\mathrm{f}(\mathrm{z}) \forall \mathrm{z} \varepsilon \mathrm{~N}(\beta) \cap \mathrm{K}
\end{aligned}
$$

so that $\phi_{\alpha}(z)=\phi_{\beta}(z)=f(z) \forall z \varepsilon G=(N(\alpha) \cap K) \cap(N(\beta) \cap K) \subset H$. Now since $\phi_{\alpha}(\mathrm{z}), \phi_{\beta}(\mathrm{z})$ are analytic in H and G is a part of H , by the uniqueness theorem $\phi_{\alpha}(\mathrm{z}) \equiv \phi_{\beta}(\mathrm{z})$ $\forall \mathrm{z} \in \mathrm{H}$. As $\alpha$ and $\beta$ are arbitrary points of $\Gamma$ we conclude that $\psi(\mathrm{z})$ is a well-defined analytic function on $D$. Let $C$ be the boundary of $D$ and let $\rho=\overline{z_{0} \zeta}, \zeta \varepsilon C$ be the minimum distance from $z_{0}$ to the boundary $C$ of $D$. Then clearly $\rho>R$ as $\varsigma$ lies outside the circle $\Gamma$. Thus we observe that $\psi(z)$ coincides with $f(z)$ on the disc $\left|z-z_{0}\right|<R$. Then it is obvious to conclude that the radius of convergence of the given power series $\sum_{n=0}^{\infty} a_{n}\left(z-z_{0}\right)^{n}$ is $\rho$, not $R$, which is a contradiction. Hence every point on $\Gamma$ cannot be regular points, i.e., there must be at least one singular point on $\Gamma$.

### 1.5 Analytic continuation along a curve

Earlier, analytic continuation by power series method, we have extended $f(z)$ to a


Fig. 9


Fig. 10 $D_{1}$, which is equal to $f(z)$ on $D \cap D_{1}$.
larger domain considering its power series expansion about a point a from its original circle of convergence with centre at $\mathrm{z}_{0}\left(-\mathrm{a} \neq \mathrm{z}_{0}\right)$ and radius r . We know, this power series converges in the disc $D_{1}:|z-a|<R$, where $\mathrm{R} \geq \mathrm{r}-\left|\mathrm{z}_{0}-\mathrm{a}\right|$ [(see Fig. 9), for example 1.2]. Then it converges to an analytic function $g(z)$ defined on

Analytic continuation along a curve is an extension of this idea to the situation where a curve is covered by an overlapping sequence of discs and an analytic function defined on the first disc, can be extended succesively to each disc in the sequence (see figure 10). We will make this idea more precise after introducing the definition of function element.

Definition 1. An ordered pair ( $f, D$ ), where $D$ is a region and $f$ is an analytic function on $D$ is called a function element. We say that it is a function element at $\mathrm{z}_{0}$ if $\mathrm{z}_{0}$ belongs to D . Two function elements $(\phi, \mathrm{G})$ and $(\psi, \mathrm{H})$ are equal if and only if $\phi(\mathrm{z}) \equiv \psi(\mathrm{z}), \mathrm{G}=\mathrm{H}$.

Clearly a function element $\left(f_{1}, D_{1}\right)$ is a direct analytic continuation of another function element ( $f_{2}, D_{2}$ ) when $D_{1} \cap D_{2} \neq \phi$ and $f_{1}=f_{2}$ in $D_{1} \cap D_{2}$. In this case the two function elements ( $\mathrm{f}_{1}, \mathrm{D}_{1}$ ) and ( $\mathrm{f}_{2}, \mathrm{D}_{2}$ ) are said to be equivalent.

Definition 2. Let $\gamma:[0,1] \rightarrow \mathbb{C}$ be a curve and $\left(f_{0}, D_{0}\right)$ be a function element at $\mathrm{z}_{0}=\gamma(0)$. Suppose there exists
(i) a partition $0=\mathrm{t}_{0}<\mathrm{t}_{1}<\ldots<\mathrm{t}_{\mathrm{n}}=1$ of $[0,1]$ and
(ii) a finite sequence of function elements

$$
\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right),\left(\mathrm{f}_{1}, \mathrm{D}_{1}\right), \ldots,\left(\mathrm{f}_{\mathrm{n}}, \mathrm{D}_{\mathrm{n}}\right)
$$

with $\gamma\left(\left[\mathrm{t}_{\mathrm{j}}, \mathrm{t}_{\mathrm{j}+1}\right]\right) \subset \mathrm{D}_{\mathrm{j}}$ and (iii) $\mathrm{f}_{\mathrm{j}}(\mathrm{z})=\mathrm{f}_{\mathrm{j}+1}(\mathrm{z})$ on $\mathrm{D}_{\mathrm{j}} \cap \mathrm{D}_{\mathrm{j}+1}$ for $\mathrm{j}=0,1, \ldots \mathrm{n}-1$.
Then $\left(f_{n}, D_{n}\right)$ is called an analytic continuation of $\left(f_{0}, D_{0}\right)$ along $\gamma$. Apparently, it seems that the function element $\left(f_{n}, D_{n}\right)$ of the above definition, depends on the choice of partition $0=t_{0}<t_{1}<\ldots<t_{n}=1$ of $[0,1]$ and the finite sequence $\left(f_{0}, D_{0}\right)$, $\left(f_{1}, D_{1}\right), \ldots,\left(f_{n}, D_{n}\right)$ of function elements. It turns out that up to equivalence, it is actually independent of these choices.

Theorem 1.3 Given a curve $\gamma:[0,1] \rightarrow \mathbb{C}$ beginning at $\mathrm{z}_{0}$ and ending at $\mathrm{z}_{\mathrm{n}}$ and a function element $\left(f_{0}, D_{0}\right)$ at $z_{0}$, any two analytic continuations of $\left(f_{0}, D_{0}\right)$ along $\gamma$ give rise to two function elements at $\mathrm{z}_{\mathrm{n}}$ that are direct analytic continuations of each other. [Though the theorem can be proved by taking different partitions of [ 0,1 ] for two different analytic continuations of $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ along $\gamma$, here we prove the theorem taking the same partition of $[0,1]$ for two analytic continuations along $\gamma]$.

Proof. Let $\left(f_{0}, F_{0}\right),\left(f_{1}, F_{1}\right), \ldots\left(f_{n}, F_{n}\right)$ and $\left(g_{0}, G_{0}\right),\left(g_{1}, G_{1}\right), \ldots,\left(g_{n}, G_{n}\right)$ be two analytic continuations of $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ along $\gamma$, using the same partition,

$$
0=\mathrm{t}_{0}<\mathrm{t}_{1}<\ldots<\mathrm{t}_{\mathrm{n}}=1
$$

where $\gamma\left(\mathrm{t}_{\mathrm{j}}\right)=\mathrm{z}_{\mathrm{j}}$ and $\gamma\left(\left[\mathrm{t}_{\mathrm{j}}, \mathrm{t}_{\mathrm{j}+1}\right]\right) \subset \mathrm{F}_{\mathrm{j}}$ and $\gamma\left(\left[\mathrm{t}_{\mathrm{j}}, \mathrm{t}_{\mathrm{j}+1}\right]\right) \subset \mathrm{G}_{\mathrm{j}}$ for $\mathrm{j}=0,1, \ldots$, n .
By given hypothesis, $\left(f_{0}, D_{0}\right)=\left(f_{0}, F_{0}\right)=\left(g_{0}, G_{0}\right)$. Now we set $E_{j}=F_{j} \cap G_{j}$ for $j=1,2, \ldots n$, and $E_{0}=F_{0}=G_{0}$. Then each $E_{j}$ is a connected open set containing $\gamma\left(\mathrm{t}_{\mathrm{j}}\right)$ and $\gamma\left(\mathrm{t}_{\mathrm{j}+1}\right)$. To prove the theorem we show, by induction, that $\mathrm{f}_{\mathrm{n}}=\mathrm{g}_{\mathrm{n}}$ on $\mathrm{E}_{\mathrm{n}}$.

We have $f_{0}=g_{0}$ on $E_{0}=F_{0}=G_{0}$ by definition. Suppose $j<n$ and $f_{j}=g_{j}$ on $E_{j}$. But we have
and

$$
\begin{array}{ll}
\mathrm{f}_{\mathrm{j}}=\mathrm{f}_{\mathrm{j}+1} & \text { on } \mathrm{f}_{\mathrm{j}} \cap \mathrm{~F}_{\mathrm{j}+1} \\
\mathrm{~g}_{\mathrm{j}}=\mathrm{g}_{\mathrm{j}+1} & \text { on } \mathrm{G}_{\mathrm{j}} \cap \mathrm{G}_{\mathrm{j}+1}
\end{array}
$$

and $\gamma\left(\mathrm{t}_{\mathrm{j}+1}\right)$ is common to both the open sets $\mathrm{F}_{\mathrm{j}} \cap \mathrm{F}_{\mathrm{j}+1}$ and $\mathrm{G}_{\mathrm{j}} \cap \mathrm{G}_{\mathrm{j}+1}$. So it follows that

$$
\mathrm{f}_{\mathrm{j}+1}=\mathrm{g}_{\mathrm{j}+1}
$$

in a neighbourhood of $\gamma\left(\mathrm{t}_{\mathrm{j}+1}\right)$ and hence on $\mathrm{E}_{\mathrm{j}+1}$ by the uniqueness theorem. By induction the proof is therefore complete.

Homotopic curves. Two arcs $\gamma_{1}$ and $\gamma_{2}$, with common end points, contained in a region R are said to be homotopic if one can be obtained from the other by continuous deformation where the process of continuous deformation must be confined in R.


In the given figure $\left\{\gamma_{1}, \gamma_{2}\right.$ and $\left.\gamma_{3}\right\}$ is one set of homotopic curves while $\left\{\gamma_{4}, \gamma_{5}\right\}$ is the other set. Here no curve of the first set is homotopic to any curve of the second set. These are geometrical interpretations. We now explain such a deformation in an analytical manner.

Let us suppose $\gamma_{0}: \mathrm{z}=\sigma_{0}(\mathrm{t}), 0 \leq \mathrm{t} \leq 1$ and $\gamma_{1}: \mathrm{z}=\sigma_{1}(\mathrm{t}), 0 \leq \mathrm{t} \leq 1$ be two curves, lying in a region $R$, having common end points $a$ and $b$ i.e., $a=\sigma_{0}(0)=\sigma_{1}(0)$ and $\mathrm{b}=\sigma_{0}(1)=\sigma_{1}(1)$ hold. We say that the curve $\gamma_{0}$ can be continuously deformed into the curve $\gamma_{1}$ keeping the process confined to $R$, if there exists a function $\sigma(t, s)$ which is continuous in the unit square $\mathrm{I}^{2}=\mathrm{I} \times \mathrm{I}, \mathrm{I}=[0,1]$ and satisfies the following conditions :
(i) for each fixed $\mathrm{s} \varepsilon[0,1]$ the curve $\gamma_{\mathrm{s}}: \mathrm{z}=\sigma(\mathrm{t}, \mathrm{s}), 0 \leq \mathrm{t} \leq 1$ lies in R .
(ii) $\sigma(\mathrm{t}, 0)=\sigma_{0}(\mathrm{t})$ and $\sigma(\mathrm{t}, 1) \equiv \sigma_{1}(\mathrm{t}), 0 \leq \mathrm{t} \leq 1$
(iii) $\sigma(0, \mathrm{~s}) \equiv \mathrm{a}$ and $\sigma(1, \mathrm{~s}) \equiv \mathrm{b}, 0 \leq \mathrm{s} \leq 1$.

Let $\alpha$ and $\varsigma$ be two points lying in a domain D and suppose that $\gamma_{0}$ and $\gamma_{1}$ are two curves connecting $\alpha$ to $\varsigma$. Let there exist, as in definition 2, two finite sequences of function elements $\left(f_{0}, G_{0}\right),\left(f_{1}, G_{1}\right) \ldots,\left(f_{n}, G_{n}\right)$ and $\left(g_{0}, H_{0}\right),\left(g_{1}, H_{1}\right), \ldots,\left(g_{m}, H_{m}\right)$ along the curves $\gamma_{0}$ and $\gamma_{1}$ respectively. We also suppose that the function elements ( $\mathrm{f}_{0}, \mathrm{G}_{0}$ ) and $\left(\mathrm{g}_{0}, \mathrm{H}_{0}\right)$ at the point $\alpha$ are equivalent. Then a question arises whether the function elements ( $f_{n}, G_{n}$ ) and $\left(g_{m}, H_{m}\right)$ at the point $\varsigma$ are also equivalent? If $\gamma_{0}$ and $\gamma_{1}$ are the same curve the Th. 1.3 confirms the answer for equivalence. However, if $\gamma_{0}$ and $\gamma_{1}$ are distinct there is no definite answer. The reason behind this is the fact that the regions enclosed by the curves $\gamma_{0}$ and $\gamma_{1}$ may contain points at which we can not find any function element that can be included in the sequence of function elements from the point $\alpha$ to $\varsigma$ along any curve passing through these points. Here we discuss a few problems highlighting these facts :

Example 1.6 Let $\mathrm{Q}_{1}=\{\mathrm{z} \varepsilon \mathbb{C} \mid \operatorname{Re}$ $z>0, \operatorname{Im} z>0\}$ denote the first quadrant and set $\mathrm{f}(\mathrm{z})=\log \mathrm{z}$ for all $\mathrm{z} \& \mathrm{Q}_{1}$

Show that, if $g_{1}$ is the analytic continuation to $\mathbb{C} \backslash(-\infty, 0]$ of $\mathrm{f} \quad$ and $\mathrm{g}_{2}$ is the analytic continuation to $\mathbb{C} \backslash[0, \infty)$ of f , then $\mathrm{g}_{1} \neq \mathrm{g}_{2}$ throughout the third quadrant, $\mathrm{Q}_{3}=\{\mathrm{z} \varepsilon \mathbb{C} \mid \operatorname{Re} \mathrm{z}<0$, $\operatorname{Imz}<0\}$.

Proof. Clearly, $\mathrm{g}_{1}$ is the principal branch of $\log \mathrm{z}$ throughout $\mathbb{C} \backslash(-\infty, 0]$


Fig. 10
by the uniqueness theorem. That is

$$
\mathrm{g}_{1}(\mathrm{z})=\int_{[1, z]} \frac{\mathrm{d} \varsigma}{\varsigma}
$$

for all z barring the negative real axis including origin. We define
(i) $\mathrm{g}_{2}(\mathrm{z})=\int_{[-1,2]} \frac{\mathrm{d} \varsigma}{\varsigma}+\mathrm{i} \pi$ for all $\mathrm{z} \varepsilon \mathbb{C} \backslash[0, \infty]$
and show that
(ii) $\mathrm{g}_{2}(\mathrm{z})=\mathrm{g}_{1}(\mathrm{z})+2 \pi \mathrm{i}$ for all $\mathrm{z} \varepsilon \mathrm{Q}_{3}$.

Let $\gamma$ be the closed curve (see figure) consisting of the line segments $[1, \mathrm{z}]$, $[\mathrm{z},-1]$ and a semi-circular path $\Gamma$ with centre at the origin and radius 1 , where z is any point in $\mathrm{Q}_{1}$.

Now we wish to calculate

$$
\int_{\gamma} \frac{d \rho}{\varsigma}
$$

By Cauchy's Residue Theorem, it is equal to $2 \pi i$ origin is the only pole inside $\gamma$ ). So breaking up the contour $\gamma$, we can equate

$$
\begin{aligned}
& \qquad \begin{aligned}
2 \pi i= & \int_{[1, z]} \frac{d \rho}{\varsigma}+\int_{[z-1]} \frac{d \zeta}{\varsigma}+\int_{\Gamma} \frac{d \varsigma}{\varsigma} \\
= & g_{1}(z)-\int_{[-1, z]} \frac{d \zeta}{\varsigma}+i \pi
\end{aligned} \\
& \text { i.e., } \quad g_{1}(z)-\int_{[-1, z]} \frac{d \varsigma}{\varsigma}+i \pi=g_{2}(z)
\end{aligned}
$$

Hence $\mathrm{g}_{2}(\mathrm{z})=\mathrm{g}_{1}(\mathrm{z})=\log \mathrm{z}$ for all $\mathrm{z} \varepsilon \mathrm{Q}_{1}$,


Fig. 11 that is, the mapping $\mathrm{g}_{2}$ defined in (i) is the unique analytic continuation of $f$ to $\mathbb{C} \backslash[0, \infty)$.

To prove (ii) Let $\mathrm{z} \varepsilon \mathrm{Q}_{3}$ and $\gamma$ be the curve joining the line segments $[-1, z],[z,+1]$ and a unit semi-circular path $\Gamma$ in the upper half plane. Thus

$$
\begin{aligned}
2 \pi i & =\int_{\gamma} \frac{d \zeta}{\zeta}=\int_{\Gamma} \frac{d \varsigma}{\zeta}+\int_{[-1, z]} \frac{d \varsigma}{\zeta}+\int_{[z-1]} \frac{d \rho}{\varsigma} \\
& =\pi i+\int_{[-1, z]} \frac{d \zeta}{\varsigma}-g_{1}(z)
\end{aligned}
$$

i.e., $g_{2}(z)=g_{1}(z)+2 \pi i$ for all $z \varepsilon Q_{3}$.

Remark : The preceding example presents the following observation: If $\gamma_{1}$ and $\gamma_{2}$ be the two curves joining $\mathrm{z}_{0}$ and $\varsigma$, $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ be a function element at $\mathrm{z}_{0}$, then the resulting function elements of $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ along the curves $\gamma_{1}$ and $\gamma_{2}$ at $\varsigma$ may not be direct analytic continuations of each other. We shall now discuss for what reasons such type of situation occurs.

### 1.6 Multi-valued Functions and Analytic continuation

When we define both real and complex functions we always keep in mind that for each value of the independent variables the value of the function must be unique. For example, even Cauchy's theorem is based on the assumption that a function can be defined uniquely in the region under consideration. All the same, multivaluedness often arises out of necessity in the actual construction of functions, the simplest example is perhaps the logarithm :

In section 5.2 [14] we showed that if $z$ is a non zero complex number, then the equation $z=e^{\omega}$ has infinitely many solutions. Since the function $f(w)=e^{\omega}$ is a many-to-one mapping, its inverse (the logarithm) is multi-valued.

Definition 3 : [Multi-valued logarithm] : For $z \neq 0$, we define the function $\log \mathrm{z}$ as the inverse of the exponential function; that is,

$$
\begin{equation*}
\log z=\omega \text { if and only if } z=e^{\omega} \tag{8}
\end{equation*}
$$

If we go through the same steps as we did to obtain (5.5) [14], we find that, for any complex number $\mathrm{z} \neq 0$, the solutions $\omega$ to equation (8) take the form

$$
\begin{equation*}
\omega=\log \mathrm{z}=\log |\mathrm{z}|+\mathrm{i} \theta, \text { for } \mathrm{z} \neq 0 \tag{9}
\end{equation*}
$$

where $\theta \varepsilon \arg z$ and $\log |z|$ denotes the natural logarithm of the positive number $|z|$. Because $\arg \mathrm{z}$ is the set $\arg \mathrm{z}=\operatorname{Arg} \mathrm{z}+2 \mathrm{n} \pi$, where n is an integer, we can express the set of values comprising $\log \mathrm{z}$ as

$$
\begin{align*}
& \log \mathrm{z}=\log |\mathrm{z}|+\mathrm{i}(\operatorname{Arg} \mathrm{z}+2 \mathrm{n} \pi), \text { where } \mathrm{n}=\text { integer }  \tag{10}\\
& \text { or } \quad \log \mathrm{z}=\log |\mathrm{z}|+\mathrm{i} \arg \mathrm{z} \text { for } \mathrm{z} \neq 0,
\end{align*}
$$

where it is understood that the identity (11) refers to the same set of numbers given in identity (10).

We call any one of the values given in identities (10) or (11) a logarithm of z . Notice that the different values of $\log \mathrm{z}$ all have the same real part and that their imaginary parts differ by the amount $2 n \pi$, where $n$ is an integer. Regarding analytic continuation, we treat $\log \mathrm{z}$ for complex valued z as the extension of $\log \mathrm{x}$ from positive real domain to complex domain. Consider the Taylor series expansion of $\log \mathrm{x}$ :

$$
\begin{equation*}
\log x=\log \{1+(x-1)\}=\sum_{n=1}^{\infty} \frac{(-1)^{n-1}}{n}(x-1)^{n}, 0<x<2 \tag{12}
\end{equation*}
$$

We take this series for complex valued $z$ and write

$$
\begin{equation*}
\mathrm{f}_{0}(\mathrm{z})=\sum_{\mathrm{n}=1}^{\infty} \frac{(-1)^{\mathrm{n}-1}}{\mathrm{n}}(\mathrm{z}-1)^{\mathrm{n}} \tag{13}
\end{equation*}
$$

which converges in the disc $\mathrm{K}_{0}:|\mathrm{z}-1|<1$ so that $\mathrm{f}_{0}(\mathrm{x})=\log \mathrm{x}$ for $0<\mathrm{x}<2$. Thus $f_{0}(z)$ and $\log x$ are direct analytic continuations of each other.

Our object is to specify the curves along which the analytic continuation of the function element $\left(f_{0}, K_{0}\right)$ is possible. For this purpose it is advantageous to apply the integral representation.

$$
\begin{equation*}
\log x=\int_{1}^{x} \frac{d s}{s}, 0<x<\infty \tag{14}
\end{equation*}
$$

Lemma 1.1. The following formula

$$
\begin{equation*}
\mathrm{f}_{0}(\mathrm{z})=\int_{1}^{\mathrm{z}} \frac{\mathrm{~d} \varsigma}{\varsigma} \tag{15}
\end{equation*}
$$

holds for $\mathrm{z} \varepsilon \mathrm{K}_{0}$ where the integral is taken along any path lying completely within $\mathrm{K}_{0}$.

Proof. The function $f_{0}(z)$ given by (13) is regular in $K_{0}$ and following Theoren $3.2[14]$ the integral on the r.h.s of (15) is also regular in $K_{0}$. But we see that this integral coincides with $\log x$ in (14) for $0<x<2$. By the uniqueness theorem.

$$
\mathrm{f}_{0}(\mathrm{z})=\sum_{\mathrm{n}=1}^{\infty} \frac{(-1)^{\mathrm{n}-1}}{\mathrm{n}}(\mathrm{z}-1)^{\mathrm{n}}=\int_{1}^{\mathrm{z}} \frac{\mathrm{~d} \varsigma}{\varsigma}, \mathrm{z} \varepsilon \mathrm{~K}_{0}
$$

In continuing $f_{0}(z)$ analytically to an arbitrary point $\omega$ we isolate a single-valued piece of $\log \mathrm{z}$, as we shall do later for other multivalued functions, called a branch of the function. The standard way to isolate single-valued branches is by the use of branch cuts to different branches. For $\log \mathrm{z}$ the question of multivaluedness arises when z goes around the origin, as a result argument changes by $2 \pi$. Such a point is called a branch point. If we do not allow the paths to travel around a branch point of a multi-valued function then certainly we would not face varied values at a point lying in the domain of definition of the function.


Fig. 12


Fig. 13

Let C be any simple curve from 0 to $\infty$, so that z cannot go around the origin crossing C.

The above consideration shows that if analytic continuation along a given curve $\Gamma$ is possible, then one can get from a function element at the initial point of the curve another function element at the terminal point of the curve by a finite number of applications of direct analytic continuation. If there is no function element at the initial point of $\Gamma$ that can be continued along $\Gamma$, then there exists a definite point on the curve $\Gamma$ which is a singular point at which the process of analytic continuation must stop.

The following question immediately arises : if $\omega$ is some non-singular point outside the disc $\mathrm{D}_{0}$, then there may two or more chains of function elements which eventually continue analytically $\mathrm{f}_{0}(\mathrm{z})$ onto a disc D containing $\omega$. For example, let $\left(f_{j}, D_{j}\right)$ be the function element of one chain and $\left(f_{k}, D_{k}\right)$ be the function element of a different chain and that $\omega \varepsilon \mathrm{D}_{\mathrm{j}} \cap \mathrm{D}_{\mathrm{k}}$; will then $\mathrm{f}_{\mathrm{j}}(\mathrm{z})=\mathrm{f}_{\mathrm{k}}(\mathrm{z}) \forall \mathrm{z} \varepsilon \mathrm{D}$ ?

## The Monodromy Theorem

The above question is answered by the Monodromy theorem, which, simply stated, is : if there are no singular points in between the two paths of analytic continuation, then the result of analytic continuation is the same for each path. Another way of stating the theorem is :

Theorem 1.4 [Monodromy Theorem] Let $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ be a function element at $\mathrm{z}_{0}$ and $R$ be a simply connected region containing $D_{0}, \varsigma$ be any point lying in $R$. We suppose
(i) ( $f_{0}, D_{0}$ ) can be analytically continued along every curve in $R$.
(ii) $\gamma_{0}$ and $\gamma_{1}$ are homotopic curves from $\mathrm{z}_{0}$ to $\varsigma$.

Then the continuations of the function element $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ along $\gamma_{0}$ and $\gamma_{1}$ at $\varsigma$ are equivalent.

Proof. A homotopy from $\gamma_{0}$ to $\gamma_{1}$ determines a continuous one parameter family of curves $\left\{\gamma_{\mathrm{s}}\right\}, 0 \leq \mathrm{s} \leq 1$ from $\mathrm{z}_{0}$ to $\varsigma$ given by the equations $\mathrm{z}=\sigma_{\mathrm{s}}(\mathrm{t}), 0 \leq \mathrm{t} \leq 1$.

By hypothesis, the function element $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ has an analytic continuation along each of the curves, $\gamma_{\mathrm{s}}$. Denote the terminal function element at $\varsigma$ for the continuation along $\gamma_{\mathrm{s}}$ by $\phi_{\mathrm{s}}$. We claim that, for each $\mathrm{k} \varepsilon[0,1]$, there is a $\delta>0$ such that $\phi_{\mathrm{s}}$ is equivalent to $\phi_{\mathrm{k}}$ whenever $|\mathrm{s}-\mathrm{k}|<\delta$.

Let $0=\mathrm{t}_{0}<\mathrm{t}_{1}<\ldots<\mathrm{t}_{\mathrm{n}}=1$ be a partition and $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right),\left(\mathrm{f}_{1}, \mathrm{D}_{1}\right), \ldots,\left(\mathrm{f}_{\mathrm{n}}, \mathrm{D}_{\mathrm{n}}\right)$ be a finite sequence of function elements defining $\phi_{k}=\left(f_{n}, D_{n}\right)$ as the terminal function element at $\varsigma$ for the analytic continuation of $\left(f_{0}, D_{0}\right)$ along $\gamma_{k}$. Then

$$
\mathrm{E}_{\mathrm{j}}=\sigma_{\mathrm{k}}\left(\left[\mathrm{t}_{\mathrm{j}}, \mathrm{t}_{\mathrm{j}+1}\right]\right) \subset \mathrm{D}_{\mathrm{j}} \text { for } \mathrm{j}=0,1, \ldots, \mathrm{n}-1
$$

For each $\mathrm{j}=0,1, \ldots \mathrm{n}-1$, let $\varepsilon_{\mathrm{j}}$ be the minimum distance from the compact set $\mathrm{E}_{\mathrm{j}}$ to the boundary of the $\mathrm{D}_{\mathrm{j}}$. If $\left|\sigma_{\mathrm{s}}(\mathrm{t})-\sigma_{\mathrm{k}}(\mathrm{t})\right|<\varepsilon_{\mathrm{j}}, \mathrm{t} \varepsilon[0,1]$, then it will also be true that $\sigma_{\mathrm{s}}\left(\left[\mathrm{t}_{\mathrm{j}}, \mathrm{t}_{\mathrm{j}+1}\right]\right) \subset \mathrm{D}_{\mathrm{j}}$. Thus, if $\varepsilon=\min \left\{\varepsilon_{0}, \varepsilon_{1}, \ldots . \varepsilon_{\mathrm{n}-1}\right\}$ and we choose $\delta>0$ such that $\left|\sigma_{\mathrm{s}}(\mathrm{t})-\sigma_{\mathrm{k}}(\mathrm{t})\right|<\varepsilon$ whenever $|\mathrm{s}-\mathrm{k}|<\delta$, then for each s with $|\mathrm{s}-\mathrm{k}|<\delta$, the partition $0=t_{0}<t_{1}<\ldots<t_{n}=1$ and sequence of function elements $\left(f_{0}, D_{0}\right),\left(f_{1}, D_{1}\right), \ldots$, $\left(f_{n}, D_{n}\right)$ also defines $\left(f_{n}, D_{n}\right)$ as the terminal function element at $\varsigma$ for the analytic continuation of ( $\mathrm{f}_{0}, \mathrm{D}_{0}$ ) along $\gamma_{\mathrm{s}}$. Since, by the previous theorem 1.3, any other continuation of ( $f_{0}, D_{0}$ ) along $\gamma_{s}$ results function element equivalent to this one, we conclude that $\phi_{\mathrm{k}}$ is equivalent to $\phi_{\mathrm{s}}$. This proves that $\phi_{\mathrm{s}}$ is equivalent to $\phi_{\mathrm{k}}$ whenever $|\mathrm{s}-\mathrm{k}|<\delta$.


Fig. 14
This means that for every s $\varepsilon \mathrm{I}=[0,1]$ there is a positive $\delta(\mathrm{s})$ such that if s lies in the interval $\mathrm{I}_{\mathrm{s}}=(\mathrm{s}-\delta(\mathrm{s}), \mathrm{s}+\delta(\mathrm{s}))$, then the analytic continuation of $\mathrm{f}_{0}(\mathrm{z})$ along all such curves $\gamma_{\mathrm{s}}$, result equivalent function elements at the point $\varsigma$. Now by the Heine-Borel theorem, we can always choose a finite number of intervals $I_{\mathrm{sj}}, 0=\mathrm{s}_{0}$ $<\mathrm{s}_{1}<\ldots .<\mathrm{s}_{\mathrm{n}}=1$ that cover the segment I and are such that the intervals $\mathrm{I}_{\mathrm{sj}}$ and
$\mathrm{I}_{\mathrm{sj}+1}, 0 \leq \mathrm{j} \leq \mathrm{n}-1$ have a non-empty intersection. Then, if $\mathrm{s} \varepsilon \mathrm{I}_{\mathrm{s} 0} \cap \mathrm{I}_{\mathrm{s} 1}$, the analytic continuation of $\mathrm{f}_{0}(\mathrm{z})$ result equivalent function elements at the point $\mathrm{\varsigma}$. The same is true for $\mathrm{s} \varepsilon \mathrm{I}_{\mathrm{s}_{1}} \cap \mathrm{I}_{\mathrm{s} 2}$ and so on. Continuing in this way we observe that the analytic continuation of the function element ( $\mathrm{f}_{0}, \mathrm{D}_{0}$ ) along all the curves $\gamma_{\mathrm{s}}, 0 \leq \mathrm{s} \leq 1$ produce equivalent function elements at the point $\varsigma$. This completes the proof of the theorem.

The above theorem leads us to the following most important corollary.
Corollary. Let R be a simply connected region and
(i) $\left(\mathrm{f}_{0}, \mathrm{D}_{0}\right)$ be a function element at $\mathrm{z}_{0}$ belonging to R
(ii) ( $\mathrm{f}_{0}, \mathrm{D}_{0}$ ) admit analytic continuation along every curve in R .

Then there is a function $F$ which is analytic on $R$ and coincides with $f_{0}$ on $D_{0}$.
Proof. Let $\mathrm{z}_{1}$ be a point in R . Then, since R is simply connected any two curves from $\mathrm{z}_{0}$, to $\mathrm{z}_{1}$ are homotopic in R . The Monodromy theorem implies that any two terminal function elements of analytic continuations of ( $f_{0}, D_{0}$ ) along curves from $z_{0}$ to $\mathrm{z}_{1}$ in R will be equivalent and hence, will determine a function $\mathrm{F}_{1}$ analytic in some neighbourhood of $z_{1}$, say $Q_{1}$.

Clearly, $F_{1}(z)=f_{0}(z)$ on $D_{0}, F_{1}(z)=f_{1}(z)$ on $D_{1}, \ldots$, etc for the continuation along the curve $\gamma_{1}$ from $z_{0}$ to $z_{1}$.

Again let $z_{2}$ be a point in $R$, and $\gamma_{2}$ be a curve in $R$ joining $z_{0}$ to $z_{2}$ and let ( $g_{n}$, $E_{n}$ ) be the function element at $z_{2}$ continuing along the curve $\gamma_{2}$ with $f_{0}=g_{0}$ on $D_{0}=$ $\mathrm{E}_{0}$. We simply join $\mathrm{z}_{2}$ to $\mathrm{Z}_{1}$ by a curve $\gamma$ and claim that continuation of $\left(\mathrm{F}_{1}, \mathrm{Q}_{1}\right)$, along the curve $\gamma$ to $\mathrm{z}_{2}$, will be equivalent to $\left(\mathrm{g}_{\mathrm{n}}, \mathrm{E}_{\mathrm{n}}\right)$ (since the curves $\gamma_{1} \cup \gamma$ and $\gamma_{2}$ are homotopic), which gives rise to the fact that there is a function $\mathrm{F}_{2}$ analytic in some neighbourhood of $z_{2}$, say $Q_{2}$, which coincides with $F_{1}$ On $Q_{1}$.

Clearly, $\mathrm{F}_{2}(\mathrm{z})$ possesses larger domain of analyticity than $\mathrm{F}_{1}(\mathrm{z})$. Proceeding in this way finite number of times we can achieve a function F analytic throughout the region R.

## Unit 2 - Harmonic Functions

## Structure

### 2.0 Objectives

### 2.1 Harmonic Function

### 2.2 Gauss' Mean Value Theorem for harmonic

### 2.3 Inverse point of a given point with respect to a circle

### 2.4 The Dirichlet Problem

### 2.5 Subharmonic \& Superharmonic Functions

### 2.0 Objectives

In this chapter we shall mainly study harmonic functions and their basic properties. Gauss' mean value theorem, Poisson's integral formula, Dirichlet's problem for a disc and Harnack inequality for harmonic functions will be discussed. Subharmonic and super harmonic functions will be explained through examples.

### 2.1 Harmonic Function

A function $\mathrm{u}(\mathrm{x}, \mathrm{y})$ of two real variables x and y defined in an open set D is said to be harmonic in D if it has continuous derivatives of the second order and satisfies the equation

$$
\begin{equation*}
\frac{\partial^{2} u}{\partial x^{2}}+\frac{\partial^{2} u}{\partial y^{2}}=0 \tag{16}
\end{equation*}
$$

known as Laplace's equation.
The differential operator $\frac{\partial^{2}}{\partial x^{2}}+\frac{\partial^{2}}{\partial y^{2}}$ is called the Laplacian and is denoted by $\nabla^{2}$.
We introduce the differential operators

$$
\begin{equation*}
\frac{\partial}{\partial \mathrm{z}}=\frac{1}{2}\left(\frac{\partial}{\partial \mathrm{x}}-\mathrm{i} \frac{\partial}{\partial \mathrm{y}}\right) \text { and } \frac{\partial}{\partial \mathrm{z}}=\frac{1}{2}\left(\frac{\partial}{\partial \mathrm{x}}+\mathrm{i} \frac{\partial}{\partial \mathrm{y}}\right) \tag{17}
\end{equation*}
$$

in order to achieve a condition equivalent to (16) for $f(z)$. If we write

$$
\begin{equation*}
x=\frac{1}{2}(z+\bar{z}) \text { and } y=\frac{1}{2 i}(z-\bar{z}) \tag{18}
\end{equation*}
$$

$$
\begin{gather*}
\frac{\partial f}{\partial z}=\frac{\partial f}{\partial x} \cdot \frac{\partial x}{\partial z}+\frac{\partial f}{\partial y} \cdot \frac{\partial y}{\partial z}=\frac{1}{2} \frac{\partial f}{\partial x}+\frac{1}{2 i} \frac{\partial f}{\partial y} \\
\frac{\partial f}{\partial z}=\frac{\partial f}{\partial x} \cdot \frac{\partial x}{\partial z}+\frac{\partial f}{\partial y} \cdot \frac{\partial y}{\partial z}=\frac{1}{2} \frac{\partial f}{\partial x}-\frac{1}{2 i} \frac{\partial f}{\partial y} \\
\frac{\partial^{2} f}{\partial z \partial \bar{z}}=\frac{1}{2}\left[\frac{\partial^{2} f}{\partial x^{2}} \cdot \frac{\partial x}{\partial z}+\frac{\partial^{2}}{\partial x \partial y} \cdot \frac{\partial y}{\partial z}\right]-\frac{1}{2 i}\left[\frac{\partial^{2} f}{\partial x \partial y} \cdot \frac{\partial x}{\partial z}+\frac{\partial^{2} f}{\partial y^{2}} \cdot \frac{\partial y}{\partial z}\right]  \tag{19a-b}\\
=\frac{1}{4} f_{x x}+\frac{1}{4 i} f_{x y}-\frac{1}{4 i} f_{x y}+\frac{1}{4} f_{y y}=\frac{1}{4}\left(f_{x x}+f_{y y}\right)
\end{gather*}
$$

and consequently the condition equivalent to (16) is

$$
\begin{equation*}
\nabla^{2} \mathrm{f}=4 \frac{\partial^{2} \mathrm{f}}{\partial \mathrm{z} \partial \overline{\mathrm{z}}} \tag{20}
\end{equation*}
$$

A function $f(z)$ is said to be harmonic in $D$ if $f$ has continuous second derivatives in D and satisfies

$$
\begin{equation*}
\nabla^{2} \mathrm{f}=0, \forall \mathrm{z} \varepsilon \mathrm{D} \tag{21}
\end{equation*}
$$

Result 1 : If $\mathrm{f}=\mathrm{u}+\mathrm{iv}$ is analytic in a domain D , then $\frac{\partial \mathrm{f}}{\partial \overline{\mathrm{z}}}=0, \forall \mathrm{z} \varepsilon \mathrm{D}$.
Proof : u and v satisfy the Cauchy-Riemann equations and using (19b) we have,

$$
\begin{aligned}
\frac{\partial \mathrm{f}}{\partial \overline{\mathrm{z}}} & =\frac{1}{2}\left(\mathrm{u}_{\mathrm{x}}+\mathrm{iv}_{\mathrm{x}}\right)-\frac{1}{2 \mathrm{i}}\left(\mathrm{u}_{\mathrm{y}}+\mathrm{iv}_{\mathrm{y}}\right) \\
& =\frac{1}{2}\left(\mathrm{u}_{\mathrm{x}}+\mathrm{iv}_{\mathrm{x}}\right)-\frac{1}{2 \mathrm{i}}\left(-\mathrm{v}_{\mathrm{x}}+\mathrm{iu}_{\mathrm{x}}\right), \text { using } C-R \text { equations } \\
& =0
\end{aligned}
$$

Result 2 : The real and imaginary parts of an analytic function are harmonic.
Proof: Let $\mathrm{f}=\mathrm{u}+\mathrm{iv}$ be analytic in a domain D. By Cauchy-Riemann equations

$$
u_{x}=v_{y} \text { and } u_{y}=-v_{x}
$$

i.e. $u_{x x}=v_{x y}$ and $u_{y y}=-v_{x y}$ [since $v_{x y}=v_{y x}$, partial derivatives being continuous]
and on addition it proves that $u$ is harmonic in $D$. Likewise $v$ is also harmonic in $D$.
Harmonic conjugates : Let $u(x, y)$ and $v(x, y)$ be two harmonic functions in a domain $\mathrm{D} \subseteq \mathbb{C}$.

If they satisfy the Cauchy-Riemann equations :

$$
\frac{\partial u}{\partial x}=\frac{\partial v}{\partial y}, \frac{\partial u}{\partial y}=-\frac{\partial v}{\partial x} \text {, in } D, \text { then }
$$

we say that $v$ is a harmonic conjugate of $u$. It follows that $f(z)=u(x, y)+i v$ $(x, y)$ is analytic in a domain $D$ if and only if $v(x, y)$ is a harmonic conjugate of $u(x$, $y$ ) in $D$.

Remark : We know that the real part as well as the imaginary part of an analytic function are harmonic. Now the questions arise :

1. Can any real harmonic function be the real part of an analytic function?
2. Whether every real harmonic function has a harmonic conjugate?

## Existence of Harmonic conjugates

Theorem 2.1 Let $u(x, y)$ be a real-valued harmonic function in a simply connected domain $\mathrm{D} \subseteq \mathbb{C}$. Then there is an analytic function $f$ in $D$ such that $u=\operatorname{Ref}$ (or, equivalently there is a function $v$, a harmonic conjugate of $u$ ) which is unique to within addition of an arbitrary real constant.

Proof. Since the function $u(x, y)$ is harmonic in a simply connected domain D, we have

$$
\frac{\partial^{2} u}{\partial x^{2}}+\frac{\partial^{2} u}{\partial y^{2}}=0
$$

which can be rewritten as

$$
\frac{\partial}{\partial y}\left(-\frac{\partial u}{\partial y}\right)=\frac{\partial}{\partial x}\left(\frac{\partial u}{\partial x}\right) \text {, where }-\frac{\partial u}{\partial y} \text { and } \frac{\partial u}{\partial x} \text { are given functions with continuous }
$$ first partial derivatives. This implies that

$$
-\left(\frac{\partial u}{\partial y}\right) d x+\left(\frac{\partial u}{\partial x}\right) d y
$$

is exact. So there is a single-valued function $\mathrm{v}(\mathrm{x}, \mathrm{y})$ which is unique to within an additive arbitrary constant, i.e.

$$
\begin{equation*}
v(x, y)=\int_{\left(x_{0}, y_{0}\right)}^{(x, y)}-\frac{\partial u}{\partial y} d x+\frac{\partial u}{\partial x} d y+K \tag{22}
\end{equation*}
$$

$\mathrm{K} \equiv$ real constant,
where $\left(\mathrm{x}_{0}, \mathrm{y}_{0}\right)$ is an initial point and ( $\mathrm{x}, \mathrm{y}$ ) is any variable point lying in D and the integral on the curve connecting ( $\mathrm{x}_{0}, \mathrm{y}_{0}$ ) to ( $\mathrm{x}, \mathrm{y}$ ) is path independent.

From (22) we find that

$$
\frac{\partial v}{\partial x}=-\frac{\partial u}{\partial y}, \frac{\partial v}{\partial y}=-\frac{\partial u}{\partial x}
$$

which in turn ensures that $\mathrm{v}(\mathrm{x}, \mathrm{y})$ is harmonic in D and harmonic conjugate to $\mathrm{u}(\mathrm{x}, \mathrm{y})$ i.e. $\mathrm{f}=\mathrm{u}+\mathrm{iv}$ forms an analytic function in D .

Observation : If D is multiply connected then the integral in (22) may take different values for different paths connecting ( $\mathrm{x}_{0}, \mathrm{y}_{0}$ ), to ( $\mathrm{x}, \mathrm{y}$ ) giving $\mathrm{v}(\mathrm{x}, \mathrm{y})$ as a multi-valued function, unless the paths are restricted to a simply connected sub domain contained in D.

Example 1. Let D be the whole plane cut along the negative real axis including the origin $(y=0, x \leq 0)$. Show that $u(x, y)=\sin x \cosh y$ is harmonic in $D$, and find its harmonic conjugate. Also find the corresponding analytic function.

Solution : Here $u(x, y)$ possesses continuous second order partial derivatives in $D$ and also satisfies the Laplace equation : $u_{x x}+u_{y y}=0$. Hence $u(x, y)$ is harmonic in D .

Let $\mathrm{v}(\mathrm{x}, \mathrm{y})$ be its harmonic conjugate. Then according to the formula (22), we have $v(x, y)=\int_{(1,0)}^{(x, y)}\left(-\frac{\partial u}{\partial y} d x+\frac{\partial u}{\partial x} d y\right)+K, K \equiv$ real constant,
where $\mathrm{M}(1,0)$ is the initial point.


> Here, $\mathrm{u}(\mathrm{x}, \mathrm{y})=\sin \mathrm{x} \cosh \mathrm{y}$
> $\mathrm{u}_{\mathrm{x}}=\cos \mathrm{x} \cosh \mathrm{y}$
> $\mathrm{u}_{\mathrm{y}}=\sin \mathrm{x} \sinh \mathrm{y}$
Now let the point $\mathrm{Q}(\mathrm{x}, \mathrm{y})$ lie in the 1st quadrant of the right-half plane. Then integrating along MNQ, we find that

$$
\mathrm{v}(\mathrm{x}, \mathrm{y})=\int_{\mathrm{MN}}-\frac{\partial \mathrm{u}}{\partial \mathrm{y}} \mathrm{dx}+\int_{\mathrm{NQ}}-\frac{\partial \mathrm{u}}{\partial \mathrm{x}} \mathrm{dy}+\mathrm{K}_{1}
$$

$$
\begin{aligned}
& =-\int_{1}^{x} \sin x \sinh o d x+\int_{0}^{y} \cos x \cosh y d y+K_{1} \\
& =\cos x \sinh y+K_{1}
\end{aligned}
$$

Again, if the point ( $\mathrm{x}, \mathrm{y}$ ) lies in the 2nd quadrant of the left-half plane, then we obtain

$$
\begin{aligned}
v(x, y) & =\int_{M N^{1}} \frac{\partial u}{\partial x} d y+\int_{N^{1} Q}-\frac{\partial u}{\partial y} d x+K_{2} \\
& =\int_{0}^{y} \cos 1 \cosh y d y+\int_{1}^{x}-\sin x \sinh y d x+K_{2} \\
& =\cos 1 \sinh y+\cos x \sinh y-\cos 1 \sinh y+K_{2} \\
& =\cos x \sinh y+K_{2}
\end{aligned}
$$

The expression for $\mathrm{v}(\mathrm{x}, \mathrm{y})$ in both the cases turns out to be the same apart from an additive constant. It results from the fact that the two paths in determining the
integral lie in a simply connected domain. Thus, $\mathrm{v}(\mathrm{x}, \mathrm{y})=\cos \mathrm{x} \sinh \mathrm{y}+\mathrm{K}$ at all points of D . Therefore, an analytic function with the given real part will be of the form
$f(z)=\sin x \cosh y+i \cos x \sinh y+i K, K \equiv$ real constant
$=\sin (x+i y)+i K$
$=\sin \mathrm{z}+\mathrm{iK}$
As for uniqueness, if two analytic functions in D have the same real part, then their difference has derivative zero, by the Cauchy-Riemann equations. In that case the functions differ by a constant.

### 2.2 Gauss' Mean Value Theorem for harmonic functions

Let $\mathrm{u}(\mathrm{z})=\mathrm{u}(\mathrm{x}, \mathrm{y}), \mathrm{z}=\mathrm{x}+\mathrm{iy}$, be harmonic in the disk $\mathrm{K}:\left|\mathrm{z}-\mathrm{z}_{0}\right|<\mathrm{R}$ and continuous on the closed disk $\overline{\mathrm{K}}$. Then

$$
\begin{equation*}
\mathrm{u}\left(\mathrm{z}_{0}\right)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{u}\left(\mathrm{z}_{0}+\operatorname{Re}^{\mathrm{i} \theta}\right) \mathrm{d} \theta \tag{23}
\end{equation*}
$$

Proof. Let $f(z)$ be an analytic function defined in $K$ such that $\operatorname{Re} f(z)=u(z)$. It follows from Cauchy's integral formula that

$$
\mathrm{f}\left(\mathrm{z}_{0}\right)=\frac{1}{2 \pi \mathrm{i}} \int_{\mathrm{l}-\mathrm{z}_{0} \mid=\mathrm{r}} \frac{\mathrm{f}(\mathrm{z})}{\mathrm{z}-\mathrm{z}_{0}} \mathrm{dz}, 0<\mathrm{r}<\mathrm{R}
$$

using the parametric form of the circle $\left|\mathrm{z}-\mathrm{z}_{0}\right|=\mathrm{r}$.
$z=z_{0}+\mathrm{re}^{\mathrm{i} \theta}, 0 \leq \theta \leq 2 \pi$, so that $\mathrm{dz}=\operatorname{ire}^{\mathrm{i} \theta} \mathrm{d} \theta$. The integral then gives

$$
\mathrm{f}\left(\mathrm{z}_{0}\right)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{f}\left(\mathrm{z}_{0}+\mathrm{re}{ }^{\mathrm{i} \theta}\right) \mathrm{d} \theta, 0<\mathrm{r}<\mathrm{R}
$$

Equating the real parts, we obtain

$$
\mathrm{u}\left(\mathrm{z}_{0}\right)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{u}\left(\mathrm{z}_{0}+\mathrm{re}^{\mathrm{i} \theta}\right) \mathrm{d} \theta
$$

whence taking the limit $r \rightarrow R$, we obtain the desired result (23)

### 2.3 Inverse point of a given point with respect to a circle

Let $\gamma:|\mathrm{z}-\alpha|=\mathrm{R}$ and $\mathrm{z}_{0}$ be a given point. Let $\mathrm{z}_{1}$ be another point on the radius through $\mathrm{z}_{0}$ such that $\left|\mathrm{z}_{0}-\alpha\right|\left|\mathrm{z}_{1}-\alpha\right|=\mathrm{R}^{2}$. Then either of the points $\mathrm{z}_{0}$ and $\mathrm{z}_{1}$ is called the inverse point of the other with respect to $\gamma$. The centre of the circle $\gamma$ is called the centre of inversion.

It follows from the definition that (i) if $\mathrm{z}_{0}$ lies inside $\gamma$, then $\mathrm{z}_{1}$ must lie outside
$\gamma$, (ii) if $\mathrm{z}_{0}$ lies on $\gamma$, then $\mathrm{z}_{1}$ must also lie on $\gamma$ and it coincides with $\mathrm{z}_{0}$, (iii) if $\mathrm{z}_{0}$ lies outside $\gamma$, then $\mathrm{z}_{1}$ must lie inside $\gamma$.

Every point, except the centre of the circle, on the plane has a unique inverse point with respect to the circle. We associate the point at infinity to the inverse point of the centre.

Result : Let $\gamma:|z|=R$ and $z_{0}$ be a given point. Then the inverse point of $z_{0}$ with respect to $\gamma$ is given by $\frac{\mathrm{R}^{2}}{\overline{\mathrm{z}}_{0}}$.

Proof : Let $\mathrm{z}_{0}=r \mathrm{e}^{\mathrm{i} \theta}$. Then its inverse point with respect to $\gamma$ is given by $\mathrm{z}_{1}=\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \theta}$, where $\mathrm{rr}_{1}=R^{2}$. Hence $r_{1}=\frac{R^{2}}{r}$ and so

$$
z_{1}=\frac{R^{2}}{r} \cdot e^{i \theta}=\frac{R^{2}}{r e^{-i \theta}}=\frac{R^{2}}{\bar{z}_{0}}
$$

Poisson's integral formula : Theorem : Let $u(x, y)$ be a harmonic function in a simply connected region D and $\gamma:|\varsigma|=\mathrm{R}$ be a circle contained in D . Then for any $\mathrm{z}=\mathrm{re}^{\mathrm{i} \theta}$, $r<R$, $u$ can be written as $u(r, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(R^{2}-r^{2}\right) \cdot u(R, \phi) d \phi}{R^{2}+r^{2}-2 R r \cos (\phi-\theta)}$, where $R e^{i \phi}$ is a point on $\gamma$.

Proof : Since $u(x, y)$ is harmonic in $D$, there exists a conjugate harmonic function $\mathrm{v}(\mathrm{x}, \mathrm{y})$ in D so that $\mathrm{f}(\mathrm{z})=\mathrm{u}(\mathrm{x}, \mathrm{y})+\mathrm{iv}(\mathrm{x}, \mathrm{y})$ is analytic in D . Then $\mathrm{f}(\mathrm{z})$ is analytic within and on $\gamma$ and so for any z within $\gamma$, by Cauchy's integral formula,

$$
\begin{equation*}
\mathrm{f}(\mathrm{z})=\frac{1}{2 \pi \mathrm{i}} \int_{\gamma} \frac{\mathrm{f}(\varsigma)}{\varsigma-\mathrm{z}} \mathrm{~d} \varsigma \tag{24}
\end{equation*}
$$

The inverse point of $z$ with respect to $\gamma$ lies outside $\gamma$ and is given by $\frac{R^{2}}{\bar{z}}$. Hence by Cauchy-Goursat theorem,

$$
\begin{equation*}
0=\frac{1}{2 \pi \mathrm{i}} \int_{\gamma} \frac{\mathrm{f}(\varsigma)}{\varsigma-\frac{\mathrm{R}^{2}}{\overline{\mathrm{z}}}} \mathrm{~d} \varsigma \tag{25}
\end{equation*}
$$

Subtracting (25) from (24) we get,

$$
\mathrm{f}(\mathrm{z})=\frac{1}{2 \pi \mathrm{i}} \int_{\gamma}\left\{\frac{1}{\varsigma-\mathrm{z}}-\frac{1}{\varsigma-\frac{\mathrm{R}^{2}}{\overline{\mathrm{z}}}}\right\} \mathrm{f}(\varsigma) \mathrm{d} \varsigma
$$

$$
\begin{equation*}
=\frac{1}{2 \pi \mathrm{i}} \int_{\gamma} \frac{\left(\mathrm{z}-\mathrm{R}^{2} / \overline{\mathrm{z}}\right) \mathrm{f}(\varsigma) \mathrm{d} \varsigma}{(\varsigma-\mathrm{z})\left(\varsigma-\frac{\mathrm{R}^{2}}{\overline{\mathrm{z}}}\right)} \tag{26}
\end{equation*}
$$

Let $\varsigma=\operatorname{Re}^{\mathrm{i} \phi}$. Also, $\overline{\mathrm{z}}=\mathrm{re}^{-\mathrm{i} \theta}$. Then (26) becomes

$$
\begin{align*}
& f\left(r e^{i \theta}\right)=\frac{1}{2 \pi i} \int_{0}^{2 \pi} \frac{\left(r e^{i \theta}-\frac{R^{2}}{r} e^{i \phi}\right) f\left(\operatorname{Re}^{i \theta}\right) i \operatorname{Re}^{i \theta} d \phi}{\left(\operatorname{Re}^{i \theta}-r e^{i \theta}\right)\left(\operatorname{Re}^{i \phi}-\frac{R^{2}}{r} e^{i \theta}\right)} \\
& =\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(\mathrm{r}^{2}-\mathrm{R}^{2}\right) \mathrm{e}^{\mathrm{i}(\phi+\theta)} \mathrm{f}\left(\operatorname{Re}^{\mathrm{i} \phi}\right) \mathrm{d} \phi}{\left(\mathrm{Re}^{\mathrm{i} \phi}-\mathrm{re}^{\mathrm{i} \theta}\right)\left(\mathrm{re}^{\mathrm{i} \phi}-\mathrm{Re}^{\mathrm{i} \theta}\right)} \\
& =\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(\mathrm{R}^{2}-\mathrm{r}^{2}\right) \mathrm{f}\left(\mathrm{Re}^{\mathrm{i} \phi}\right) \mathrm{d} \phi}{\left(\mathrm{Re}^{\mathrm{i} \phi}-\mathrm{re}^{\mathrm{i} \theta}\right)\left(\mathrm{Re}^{-\mathrm{i} \phi}-\mathrm{re}^{-\mathrm{i} \theta}\right)} \\
& =\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(\mathrm{R}^{2}-\mathrm{r}^{2}\right) \mathrm{f}\left(\mathrm{Re}^{\mathrm{i} \phi}\right) \mathrm{d} \phi}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \mathrm{Rr}^{\cos (\phi-\theta)}} \tag{27}
\end{align*}
$$

Let $\quad \mathrm{f}\left(\mathrm{re}^{\mathrm{i} \theta}\right)=\mathrm{u}(\mathrm{r}, \theta)+\mathrm{iv}(\mathrm{r}, \theta)$. Then (27) becomes

$$
\begin{equation*}
\mathrm{u}(\mathrm{r}, \theta)+\mathrm{iv}(\mathrm{r}, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(\mathrm{R}^{2}-\mathrm{r}^{2}\right)\{\mathrm{u}(\mathrm{R}, \phi)+\mathrm{iv}(\mathrm{R}, \phi)\}}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \operatorname{Rr} \cos (\phi-\theta)} \mathrm{d} \phi \tag{28}
\end{equation*}
$$

Equating real parts in (28) we get,

$$
\begin{equation*}
\mathrm{u}(\mathrm{r}, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(\mathrm{R}^{2}-\mathrm{r}^{2}\right) \mathrm{u}(\mathrm{R}, \phi)}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \operatorname{Rr} \cos (\phi-\theta)} \mathrm{d} \phi \tag{29}
\end{equation*}
$$

Formula (29) is known as Poisson's integral formula.
Note : Let $\frac{R^{2}-r^{2}}{R^{2}+r^{2}-2 R r \cos (\phi-\theta)}=P(R, r, \phi-\theta)$. Then,
the function $P(R, r, \phi-\theta)$ is called the Poisson Kernel. Hence we can write (29) in the form

$$
\begin{equation*}
\mathrm{u}(\mathrm{r}, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\phi) \mathrm{u}(\mathrm{R}, \phi) \mathrm{d} \phi \tag{30}
\end{equation*}
$$

We can also get a formula similar to (29) for the imaginary part of $f(z)$ by equating the imaginary part in (28). The corresponding formula is

$$
\begin{equation*}
\mathrm{v}(\mathrm{r}, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(\mathrm{R}^{2}-\mathrm{r}^{2}\right) \mathrm{v}(\mathrm{R}, \phi) \mathrm{d} \phi}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \mathrm{Rr} \cos (\phi-\theta)}=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta) \mathrm{v}(\mathrm{R}, \phi) \mathrm{d} \phi \tag{31}
\end{equation*}
$$

Remark : Cauchy's integral formula expresses the values of an analytic function inside a circle in terms of its values on the boundary of the circle whereas Poisson's integral formula expresses the values of a harmonic function inside a circle in terms of its values on the boundary of the circle.

Result 3. $\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta) \mathrm{d} \phi=1$.
Proof : By Poisson's integral formula we have,

$$
\begin{aligned}
& \mathrm{u}(\mathrm{r}, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta) \mathrm{u}(\mathrm{R}, \phi) \mathrm{d} \phi \text { Taking } \mathrm{u}(\mathrm{r}, \theta) \equiv 1 \text { we get, } \\
& \frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta) \mathrm{d} \phi=1
\end{aligned}
$$

Result 4. $\mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta)=\operatorname{Re}\left(\frac{\varsigma+z}{\varsigma-z}\right)$
Proof: Let $\varsigma=R e^{i \phi}, z=r e^{i \theta}, r<R$. Then,

$$
\begin{aligned}
& \frac{\varsigma+\mathrm{z}}{\varsigma-\mathrm{z}}=\frac{\mathrm{Re}^{\mathrm{i} \phi}+\mathrm{re}^{\mathrm{i} \theta}}{\mathrm{Re}^{\mathrm{i} \phi}-\mathrm{re}} \mathrm{r} \theta \\
& =\frac{(\mathrm{R} \cos \phi+\mathrm{r} \cos \theta)+\mathrm{i}(\mathrm{R} \sin \phi+\mathrm{r} \sin \theta)}{(\mathrm{R} \cos \phi-\mathrm{r} \cos \theta)+\mathrm{i}(\mathrm{R} \sin \phi-\mathrm{r} \sin \theta)} \\
& =\frac{\{(\mathrm{R} \cos \phi+\mathrm{r} \cos \theta)+\mathrm{i}(\mathrm{R} \sin \phi+\mathrm{r} \sin \theta)\}\{(\mathrm{R} \cos \phi-\mathrm{r} \cos \theta)-\mathrm{i}(\mathrm{R} \sin \phi-\mathrm{r} \sin \theta)\}}{(\mathrm{R} \cos \phi-\mathrm{r} \cos \theta)^{2}+(\mathrm{R} \sin \phi-\mathrm{r} \sin \theta)^{2}}
\end{aligned}
$$

Simplifying we get, $\operatorname{Re}\left(\frac{\varsigma+z}{\varsigma-z}\right)=\frac{R^{2}-r^{2}}{R^{2}+r^{2}-2 \operatorname{Rr} \cos (\phi-\theta)}=P(R, r, \phi-\theta)$.
Result 5. Poisson Kernel $\mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta)$ is harmonic in $|\mathrm{z}|<\mathrm{R}$.
Proof : Let $f(z)=\frac{\varsigma+z}{\varsigma-z}$. Then $f(z)$ is analytic in $|z|<R$. By result 4, $P(R, r \phi-$ $\theta)=\operatorname{Ref}(z)$. Hence the Poisson Kernel is the real part of an analytic function. Hence $P(R, r, \phi-\theta)$ is harmonic in $|z|<R$.

Note : We can easily show that $\frac{R^{2}-r^{2}}{R^{2}+r^{2}-2 \operatorname{Rr} \cos (\phi-\theta)}=\frac{R^{2}-|z|^{2}}{\left|\operatorname{Re}^{i \phi}-z\right|^{2}}$
where $z=r e^{i \theta}, r<R$. Hence $\operatorname{Re}\left(\frac{\varsigma+z}{\varsigma-z}\right)=\frac{R^{2}-|z|^{2}}{\left|\operatorname{Re}^{i \phi}-z\right|^{2}}$ and Poisson's integral formula (29) can be written as

$$
\begin{equation*}
\mathrm{u}(\mathrm{r}, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\mathrm{R}^{2}-|\mathrm{z}|^{2}}{\left|\operatorname{Re}^{\mathrm{i} \mathrm{\phi}}-\mathrm{z}\right|^{2}} \mathrm{u}(\mathrm{R}, \phi) \mathrm{d} \phi \tag{32}
\end{equation*}
$$

The function $\frac{R^{2}-|z|^{2}}{\left|\operatorname{Re}^{i \phi}-z\right|^{2}}$ is the Poisson Kernel.
Theorem 2.2 Let $\mathrm{u}(\mathrm{x}, \mathrm{y}) \neq$ constant be harmonic on a simply connected domain D. Then $u(x, y)$ has neither a maximum nor a minimum at any point of $D$.

Proof. Let $\mathrm{z}_{0}=\mathrm{x}_{0}+\mathrm{iy} \mathrm{y}_{0}$ be an arbitrary point of D . Then following theorem 2.1 there is an analytic function $f(z)$ in a neighbourhood $N\left(z_{0}\right)$ of $z_{0}$ such that $\operatorname{Re} f=u$. Then

$$
\mathrm{g}(\mathrm{z})=\mathrm{e}^{\mathrm{f}(\mathrm{z})}
$$

is analytic on $N\left(z_{0}\right)$ and not equal to constant since $u(x, y) \neq$ constant and

$$
|\mathrm{g}(\mathrm{z})|=\mathrm{e}^{\mathrm{u}(\mathrm{x}, \mathrm{y})}
$$

Again exponential function is strictly increasing, so a maximum for u at $\left(\mathrm{x}_{0}, \mathrm{y}_{0}\right)$ is also a maximum for $e^{u}$, and hence also a maximum of $\left|e^{f}\right|$ i.e. of $|g(z)|$ at $z_{0}$. The function $u(x, y)$ cannot have a maximum at ( $\mathrm{x}_{0}, \mathrm{y}_{0}$ ), since otherwise $|\mathrm{g}(\mathrm{z})|$ would have a maximum at $\mathrm{z}_{0}$, thereby contradicting the maximum modulus principle. Likewise, following the minimum modulus principle $|\mathrm{g}(\mathrm{z})|$ cannot have a minimum value at $\mathrm{z}_{0}$ since $|g(z)| \neq 0$ on $D$. Therefore $u(x, y)$ cannot possess minimum value at $\left(x_{0}, y_{0}\right)$.

Corollary. Let $u(x, y)$ be harmonic on a domain $D$ and continuous on $\overline{\mathrm{D}}$. Then $\mathrm{u}(\mathrm{x}, \mathrm{y})$ attains its maximum and its minimum on the boundary of D .

Proof. Since $u(x, y)$ is continuous on the compact set $\overline{\mathrm{D}}$, it attains both its maximum and its minimum on $\overline{\mathrm{D}}$, but $\mathrm{u}(\mathrm{x}, \mathrm{y})$ cannot possess a maximum or a minimum at a point of D. Therefore the corollary follows.

Example 2. Given $\mathrm{u}(\mathrm{x}, \mathrm{y})$ harmonic in the disk $|\mathrm{z}|<\mathrm{R}$ and $\mathrm{A}\left(\mathrm{r}_{\mathrm{j}}\right)$ its maximum value on the circle $|z|=r_{j}, r_{j}<R, j=1,2,3$. Prove that

$$
A\left(r_{2}\right) \leq \frac{\log r_{2}-\log r_{1}}{\log r_{3}-\log r_{1}} A\left(r_{3}\right)+\frac{\log r_{3}-\log r_{2}}{\log r_{3}-\log r_{1}} A\left(r_{1}\right)
$$

for $\quad 0<r_{1}<r_{2}<r_{3}<R$.
Solution. Since $u(x, y)$ is harmonic in $|z|<R, u(x, y)+\alpha \log r, r=\sqrt{x^{2}+y^{2}}, \alpha \equiv a$ real constant to be fixed later, is also harmonic in the annulus $r_{1} \leq|z| \leq r_{3}$. Hence its
maximum is attained on the boundary of the annulus i.e. on $|z|=r_{1}$ or, $|z|=r_{3}$ or, on both. Either $\mathrm{A}\left(\mathrm{r}_{1}\right)+\alpha \log \mathrm{r}_{1}$ or, $\mathrm{A}\left(\mathrm{r}_{3}\right)+\alpha \log \mathrm{r}_{3}$ is maximum. We define $\alpha$ so that

$$
\begin{aligned}
& \mathrm{A}\left(\mathrm{r}_{1}\right)+\alpha \log \mathrm{r}_{1}=\mathrm{A}\left(\mathrm{r}_{3}\right)+\alpha \log \mathrm{r}_{3} \\
& \text { or, } \quad \alpha=\frac{\mathrm{A}\left(\mathrm{r}_{1}\right)-\mathrm{A}\left(\mathrm{r}_{3}\right)}{\log \mathrm{r}_{3}-\log \mathrm{r}_{1}}
\end{aligned}
$$

The circle $|\mathrm{z}|=\mathrm{r}_{2}$ lies inside the annulus $\mathrm{r}_{1} \leq|\mathrm{z}| \leq \mathrm{r}_{3}$ and according to corollary of the theorem 2.2 regarding maximum value of the harmonic function $u(x, y)+\alpha \log r$ we have

$$
\mathrm{A}\left(\mathrm{r}_{2}\right)+\alpha \log \mathrm{r}_{2} \leq \mathrm{A}\left(\mathrm{r}_{3}\right)+\alpha \log \mathrm{r}_{3}
$$

or, $\quad \mathrm{A}\left(\mathrm{r}_{2}\right) \leq \mathrm{A}\left(\mathrm{r}_{3}\right)+\alpha\left(\log \mathrm{r}_{3}-\log \mathrm{r}_{2}\right)$

$$
\begin{aligned}
& =A\left(r_{3}\right)+\frac{A\left(r_{1}\right)-A\left(r_{3}\right)}{\log r_{3}-\log r_{1}}\left(\log r_{3}-\log r_{2}\right) \\
& =\frac{\log r_{2}-\log r_{1}}{\log r_{3}-\log r_{1}} A\left(r_{3}\right)+\frac{\log r_{3}-\log r_{2}}{\log r_{3}-\log r_{1}} A\left(r_{1}\right)
\end{aligned}
$$

### 2.4 The Dirichlet Problem

Let D be a domain with boundary $\Gamma$ and let $\cup(x, y)$ be a continuous real function defined on $\Gamma$. The Dirichlet problem is to find a function $u(x, y)$, harmonic on $D$ and continuous on $\overline{\mathrm{D}}$, which coincides with $\cup(x, y)$ at every point of $\Gamma$.

## Existence of a solution of Dirichlet's problem for a disc

Theorem 2.3 Let $D$ be the disc $|z|<R$ with boundary $\Gamma:|z|=R$ and let $U(\phi)$ be a continuous real function on the interval $[0,2 \pi]$ such that $\mathrm{U}(0)=\mathrm{U}(2 \pi)$. Then the function $\mathrm{u}(\mathrm{r}, \theta)$ defined by the integral

$$
\begin{equation*}
\mathrm{u}(\mathrm{r}, \theta)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\left(\mathrm{R}^{2}-\mathrm{r}^{2}\right) \mathrm{U}(\phi)}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \mathrm{Rr} \cos (\phi-\theta)} \mathrm{d} \phi \tag{33}
\end{equation*}
$$

for any point $(r, \theta)$ on $D$ any by $u(R, \phi)=U(\phi)$
for any point $(\mathrm{R}, \phi)$ on $\Gamma$, solves the Dirichlet problem for the disc D . In otherwords, (i) $u$ is harmonic on $D$ and continuous on $\bar{D}$ and (ii) $\lim _{\mathrm{re}^{\mathrm{i} \theta} \rightarrow \mathrm{Re}^{\mathrm{if} \mathrm{i}_{0}}} \mathrm{u}(\mathrm{r}, \theta)=\mathrm{U}\left(\phi_{0}\right)$,
where $\mathrm{Re}^{\mathrm{i} \phi_{0}}$ is any fixed point on $\Gamma$.
Proof: To prove that $u(r, \theta)$ defined by (33) on $D$ is harmonic on $D$ we observe that

$$
\frac{\mathrm{R}^{2}-\mathrm{r}^{2}}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \mathrm{Rr} \cos (\phi-\theta)}=\mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta)
$$

$=\operatorname{Re}\left(\frac{\varsigma+z}{\varsigma-z}\right)$, where $P(R, r, \phi-\theta)$ is the Poisson Kernel and $\varsigma=\operatorname{Re}^{i \phi}, \mathrm{z}=\mathrm{re}^{\mathrm{i} \theta}, \mathrm{r}<\mathrm{R}$. The r.h.s. is the real part of the function $\frac{\varsigma+z}{\varsigma-z}$ which is analytic in D. Hence the Poisson Kernel $\mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta)$ is harmonic in D . So, differentiation under the sign of integration is valid. Applying the Laplacian $\nabla^{2}$ in (r, $\theta$ ) to both sides of (33) we get,

$$
\nabla^{2} \mathrm{u}=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{U}(\phi) . \nabla^{2} \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta) \mathrm{d} \phi=0 \quad[\text { Since } \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta)
$$

is harmonic in $\left.\mathrm{D} \Rightarrow \nabla^{2} \mathrm{P}(\mathrm{R}, \mathrm{r}, \phi-\theta)=0\right]$.
$\Rightarrow \mathrm{u}$ is harmonic on D .
Next we prove that the function $u(r, \theta)$ defined by the integral (33) approaches $\mathrm{U}\left(\phi_{0}\right)$ as the point $(\mathrm{r}, \theta)$ in D tends to any fixed point $\left(\mathrm{R}, \phi_{0}\right)$ on $\Gamma$.

Let $\left(r_{n}, \theta_{n}\right)$ be an arbitrary sequence of points in $D$ converging to the boundary point ( $\mathrm{R}, \phi_{0}$ ). We now consider the difference

$$
\begin{align*}
u\left(r_{n}, \theta_{n}\right)-U\left(\phi_{0}\right)= & \frac{1}{2 \pi} \int_{0}^{2 \pi} P\left(R, r_{n}, \phi-\theta_{n}\right) U(\phi) d \phi-U\left(\phi_{0}\right) \\
= & \frac{1}{2 \pi} \int_{0}^{2 \pi}\left\{U(\phi)-U\left(\phi_{0}\right)\right\} P\left(R, r_{n}, \phi-\theta_{n}\right) d \phi  \tag{35}\\
& \left(\text { Since } \frac{1}{2 \pi} \int_{0}^{2 \pi} P\left(R, r_{n}, \phi-\theta_{n}\right) d \phi=1\right)
\end{align*}
$$

Since $U(\phi)$ is continuous on $\Gamma$, for given $\epsilon>0$ there exists a $\delta(\epsilon)>0$ such that

$$
\begin{equation*}
\left|\mathrm{U}(\phi)-\mathrm{U}\left(\phi_{0}\right)\right|<\frac{\epsilon}{2} \tag{36}
\end{equation*}
$$

whenever $\quad\left|\phi-\phi_{0}\right|<2 \delta$
we choose $\delta$ so small that (36) is satisfied and $\phi_{0}-2 \delta>0, \phi_{0}+2 \delta<2 \pi$. We break the integral on r.h.s. of (35) as

$$
\begin{align*}
& \left|\mathrm{u}\left(\mathrm{r}_{\mathrm{n}}, \theta_{\mathrm{n}}\right)-\mathrm{U}\left(\phi_{0}\right)\right| \leq\left|\frac{1}{2 \pi} \int_{0}^{\phi_{0}-2 \delta} \mathrm{P}\left(\mathrm{R}, \mathrm{r}_{\mathrm{n}}, \phi-\theta_{\mathrm{n}}\right)\left\{\mathrm{U}(\phi)-\mathrm{U}\left(\phi_{0}\right)\right\} \mathrm{d} \phi\right| \\
& \quad+\left|\frac{1}{2 \pi} \int_{\phi_{0}-2 \delta}^{\phi_{0}+2 \delta} \cdots\right|+\left|\frac{1}{2 \pi} \int_{\phi_{\phi_{0}}+2 \delta}^{2 \pi} \underset{\cdots}{2}\right|=\left|\mathrm{I}_{1}\right|+\left|\mathrm{I}_{2}\right|+\left|\mathrm{I}_{3}\right| \tag{38}
\end{align*}
$$

Now, $\left.\quad\left|\mathrm{I}_{2}\right| \leq \frac{1}{2 \pi} \int_{\phi_{0}-2 \delta}^{\phi_{0}+2 \delta}\left|\mathrm{P}\left(\mathrm{R}, \mathrm{r}_{\mathrm{n}}, \phi-\theta_{\mathrm{n}}\right)\right| \mathrm{U}(\phi)-\mathrm{U}\left(\phi_{0}\right) \right\rvert\, \mathrm{d} \phi$

$$
\begin{equation*}
<\frac{\epsilon}{2} \cdot \frac{1}{2 \pi} \int_{0}^{2 \pi}\left|\mathrm{P}\left(\mathrm{R}, \mathrm{r}_{\mathrm{n}}, \phi-\theta_{\mathrm{n}}\right)\right| \mathrm{d} \phi=\frac{\varepsilon}{2} \tag{39}
\end{equation*}
$$

To estimate the other two terms we choose $n$ so large that
$\left|\phi_{0}-\theta_{\mathrm{n}}\right|<\delta$. Then, $\left.\left|\phi-\theta_{\mathrm{n}}\right|=\left|\phi-\phi_{0}+\phi_{0}-\theta_{\mathrm{n}}\right| \geq\left|\phi-\phi_{0}\right|-\mid \phi_{0}-\theta_{\mathrm{n}}\right)>2 \delta$ $-\delta=\delta$ since $\left|\phi-\phi_{0}\right|>2 \delta$ whenever $\phi$ belongs to either of the intervals $\left[0, \phi_{0}-\right.$ $2 \delta]$ or, $\left[\phi_{0}+2 \delta, 2 \pi\right]$.

Then, $\left|\mathrm{I}_{1}\right|+\left|\mathrm{I}_{3}\right| \leq 2 \mathrm{M} \cdot \frac{1}{2 \pi} \cdot \frac{\mathrm{R}^{2}-\mathrm{r}_{\mathrm{n}}^{2}}{\mathrm{R}^{2}+\mathrm{r}_{\mathrm{n}}^{2}-2 \mathrm{Rr}_{\mathrm{n}} \cos \delta}\left(\int_{0}^{\phi_{0}-2 \delta} \mathrm{~d} \phi+\int_{\phi_{0}+2 \delta}^{2 \pi} \mathrm{~d} \phi\right)$

$$
<2 \mathrm{M} \frac{\mathrm{R}^{2}-\mathrm{r}_{n}^{2}}{\mathrm{R}^{2}+\mathrm{r}_{\mathrm{n}}^{2}-2 \mathrm{Rr}_{\mathrm{n}} \cos \delta} \rightarrow 0 \text { as } \mathrm{r}_{\mathrm{n}} \rightarrow \mathrm{R}
$$

where

$$
\begin{equation*}
\mathrm{M}=\operatorname{Max}_{\phi[0,2 \pi]}\left|\mathrm{U}(\phi)-\mathrm{U}\left(\varphi_{0}\right)\right| \text { and } \cos \left(\phi-\theta_{\mathrm{n}}\right)<\cos \delta . \tag{40}
\end{equation*}
$$

Thus, for sufficiently large $n,\left|\mathrm{I}_{1}\right|+\left|\mathrm{I}_{3}\right|<\frac{\varepsilon}{2}$
Using (39) and (40) in (38) we get,
$\left|\mathrm{u}\left(\mathrm{r}_{\mathrm{n}}, \theta_{\mathrm{n}}\right)-\mathrm{U}\left(\phi_{0}\right)\right|<\varepsilon$ for sufficiently large $\mathrm{n} ;$
i.e. $\lim _{n \rightarrow \infty} u\left(r_{n}, \theta_{n}\right)=U\left(\phi_{0}\right)$
where $\left(r_{n}, \theta_{n}\right)$ is an arbitrary sequence of points in $D$ approaching ( $R, \phi_{0}$ ).
Equation (41) still holds if some or all the points $\left(r_{n}, \theta_{n}\right)$ lie on $\Gamma$ since in that case we can directly use the fact that $\mathrm{U}(\phi)$ is continuous on $\Gamma$. This implies $\mathrm{u}(\mathrm{r}, \theta)$ is continuous on $\overline{\mathrm{D}}$. This completes the proof.

## Uniqueness of the solution to the Dirichlet problem for a disc.

Let $u_{1}$ and $u_{2}$ be two solutions of the Dirichlet problem. Then their difference $u_{1}$ $-\mathrm{u}_{2}=\mathrm{h}$ is harmonic in D and continuous in the closed disk and takes the value zero on the boundary. Hence $h$ attains its upper bounds at some points of the closed disk. If $1>0$, the upper bound will occur in the open disk, since on the boundary $\Gamma \mathrm{h}$ is zero. This contradicts the conclusions of theorem 2.2. So then $l=0$. In the same way we can show that the lower bound of $h$ on $\overline{\mathrm{D}}$ is zero. Thus there is no alternative but h to be zero on $\overline{\mathrm{D}}$.

Theorem 2.4 Any continuous function $u(z)$ possessing the mean-value property in a domain D is harmonic in D .

Proof. Let $\overline{\mathrm{K}}$ be a closed disk contained in D. By hypothesis of the theorem u satisfies the mean value property in K . We shall prove that u is harmonic in K . By the theorem 2.3 on the Dirichlet problem for a disk there exists a continuous function $\tilde{\mathrm{u}}(\mathrm{z})$ in K , which is harmonic in the interior of K and coincides with $\mathrm{u}(\mathrm{z})$ on the boundary of $K$. The difference $u-\tilde{u}$ is continuous and satisfies the mean-value property in K. By the corollary to the theorem 3.7 [(14) page-58] u-ũ satisfies the maximum modulus prnciple in $K$. Now as $u-\tilde{u}$ is zero on the boundary of $K$, it will be identically zero in K . Therefore u coincides with the harmonic function $\tilde{\mathrm{u}}$ in the interior of K and since K is arbitrary, u is harmomic in the domain D .

The Harnack Inequality : Let $u$ be a non-negative Harmonic function on a closed disk $\overline{\mathrm{D}}(0, \mathrm{R})$. Then, for any point $\mathrm{z} \varepsilon \mathrm{D}(0, R)$

$$
\begin{equation*}
\frac{\mathrm{R}-|\mathrm{z}|}{\mathrm{R}+|\mathrm{z}|} \mathrm{u}(0) \leq \mathrm{u}(\mathrm{z}) \leq \frac{\mathrm{R}+|\mathrm{z}|}{\mathrm{R}-|\mathrm{z}|} \mathrm{u}(0) \tag{42}
\end{equation*}
$$

where $\mathrm{D}(0, \mathrm{R})$ denotes a disk with centre 0 and radius R .
Proof. From the Poisson's integral formula for $u$ on $\bar{D}(0, R)$ :

$$
\mathrm{u}(\mathrm{z})=\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{u}\left(\operatorname{Re}^{\mathrm{i} \phi}\right) \frac{\mathrm{R}^{2}-|\mathrm{z}|^{2}}{\left|\operatorname{Re}^{\mathrm{i} \phi}-\mathrm{z}\right|^{2}} \mathrm{~d} \phi
$$

$$
\text { Now, } \quad \frac{\mathrm{R}^{2}-|\mathrm{z}|^{2}}{\left|\mathrm{Re}^{\mathrm{i} \mathrm{\phi}}-\mathrm{z}\right|^{2}} \leq \frac{\mathrm{R}^{2}-|\mathrm{z}|^{2}}{(\mathrm{R}-|\mathrm{z}|)^{2}}=\frac{\mathrm{R}+|\mathrm{z}|}{\mathrm{R}-|\mathrm{z}|}
$$

Combining these two, we see that

$$
\mathrm{u}(\mathrm{z}) \leq \frac{\mathrm{R}+|\mathrm{z}|}{\mathrm{R}-|\mathrm{z}|} \frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{u}\left(\operatorname{Re}^{\mathrm{i} \phi}\right) \mathrm{d} \phi=\frac{\mathrm{R}+|\mathrm{z}|}{\mathrm{R}-|\mathrm{z}|} \mathrm{u}(0),
$$

where we make use of the mean value theorem. Similarly, the other inequality in
(42) will follow from $\frac{R^{2}-|z|^{2}}{\left|\operatorname{Re}^{\mathrm{iq}}-\mathrm{z}\right|^{2}} \geq \frac{\mathrm{R}^{2}-|\mathrm{z}|^{2}}{(\mathrm{R}+|\mathrm{z}|)^{2}}=\frac{\mathrm{R}-|\mathrm{z}|}{\mathrm{R}+|\mathrm{z}|}$

Corollary Let $u$ be a non-negative harmonic function on a closed disk $\overline{\mathrm{D}}(\varsigma, R)$. Then for any $z \in D(\varsigma, R)$,

$$
\begin{equation*}
\frac{R-|z-\varsigma|}{R+|z-\varsigma|} u(\varsigma) \leq u(z) \leq \frac{R+|z-\varsigma|}{R-|z-\varsigma|} u(\varsigma) \tag{43}
\end{equation*}
$$

### 2.5 Subharmonic \& Superharmonic Functions

Definition : A real-valued continuous function $u(x, y)$ in an open set $D$ of the complex plane $\mathbb{C}$ is said to be
(i) subharmonic if, for any $\varsigma \varepsilon D$

$$
u(\varsigma) \leq \frac{1}{2 \pi} \int_{0}^{2 \pi} u\left(\varsigma+r e^{i \theta}\right) d \theta
$$

hold for sufficiently small $\mathrm{r}>0$.
(ii) superharmonic if, for any a $\varepsilon \mathrm{D}$

$$
\mathrm{u}(\mathrm{a}) \geq \frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{u}\left(\mathrm{a}+\mathrm{re}^{\mathrm{i} \theta}\right) \mathrm{d} \theta
$$

hold for sufficiently small $\mathrm{r}>0$.
From the definition it follows that every harmonic function is subharmonic as well as superharmonic.

Example 3. If $f(z)$ is analytic on a domain D, then $|f(z)|$ is subharmonic but not harmonic in D unless $\mathrm{f}(\mathrm{z}) \equiv$ constant.

Solution : Using the Cauchy's integral formula

$$
\begin{equation*}
|f(a)| \leq \frac{1}{2 \pi} \int_{0}^{2 \pi}\left|f\left(a+r^{i \theta}\right)\right| d \theta \tag{44}
\end{equation*}
$$

for every a $\varepsilon \mathrm{D}$ and $\mathrm{r}(>0)$ is small enough. Here equality holds only if $\mathrm{f}(\mathrm{z}) \equiv$ constant. We now show that the integral

$$
\left.\mathrm{I}(\mathrm{r})=\frac{1}{2 \pi} \int_{0}^{2 \pi} \right\rvert\, \mathrm{f}\left(\mathrm{a}+\mathrm{re} \mathrm{e}^{\mathrm{i} \theta}\right) \mathrm{d} \theta
$$

is a strictly increasing function of $r$, if $f(z) \neq$ constant. Let $0<r_{1}<r_{2}<k(a)$ and $\mathrm{g}(\theta)$ be continuous on $[0,2 \pi]$ and $\mathrm{F}(\mathrm{z})$ be defined by
(i) $g(\theta) f\left(a+r_{1} e^{i \theta}\right)=\left|f\left(a+r_{1} e^{i \theta}\right)\right|, 0 \leq \theta \leq 2 \pi$
(ii) $F(z)=\frac{1}{2 \pi} \int_{0}^{2 \pi} f\left(a+z e^{i \theta}\right) g(\theta) d \theta,|z| \leq r_{2}$
(iii) $\mathrm{k}(\mathrm{a}) \equiv$ minmum distance between a and the boundary of D .
$\mathrm{F}(\mathrm{z})$ is regular for $|\mathrm{z}| \leq \mathrm{r}_{2}$ and attains its maximum of the boundary of the disc, say at $z=r_{2} e^{i \phi}$. Then

$$
\begin{aligned}
\mathrm{I}\left(\mathrm{r}_{1}\right) & =\frac{1}{2 \pi} \int_{0}^{2 \pi}\left|\mathrm{f}\left(\mathrm{a}+\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \theta}\right)\right| \mathrm{d} \theta \\
& =\frac{1}{2 \pi} \int_{0}^{2 \pi} \mathrm{f}\left(\mathrm{a}+\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \theta}\right) \mathrm{g}(\theta) \mathrm{d} \theta
\end{aligned}
$$

$$
\begin{aligned}
& =\mathrm{F}\left(\mathrm{r}_{1}\right) \\
& <\left|\mathrm{F}\left(\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \theta}\right)\right| \\
& \leq \frac{1}{2 \pi} \int_{0}^{2 \pi}\left|\mathrm{f}\left(\mathrm{a}+\mathrm{r}_{2} \mathrm{e}^{\mathrm{i}(\theta+\phi)}\right)\right| \mathrm{d} \theta \\
& =\frac{1}{2 \pi} \int_{\phi}^{2 \pi+\phi}\left|\mathrm{f}\left(\mathrm{a}+\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \psi}\right)\right| \mathrm{d} \psi, \text { taking } \phi+\theta=\psi \\
& =\frac{1}{2 \pi}\left\{\int_{0}^{2 \pi}-\int_{0}^{\phi}+\int_{2 \pi}^{2 \pi+\phi} \mid \mathrm{f}\left(\mathrm{a}+\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \psi}\right) \mathrm{d} \psi\right\} \\
& =\frac{1}{2 \pi} \int_{0}^{2 \pi}\left|\mathrm{f}\left(\mathrm{a}+\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \psi}\right)\right| \mathrm{d} \psi, \quad \text { substituting } \psi=2 \pi+\theta \text { in the third }
\end{aligned}
$$

integral, we find that it cancels the second term)
$=\mathrm{I}\left(\mathrm{r}_{2}\right)$. Hence equality in (44) is possible if and only if $\mathrm{f}(\mathrm{z}) \equiv$ constant. Therefore $|\mathrm{f}(\mathrm{z})|$ is subharmonic but not harmonic in D unless $\mathrm{f}(\mathrm{z}) \equiv$ constant.

Example 4. If $f(z) \neq 0$ is analytic in a domain $D$, then $\log |f(z)|$ is subharmonic in $D$.

Solution : Let $\Phi(\mathrm{z})=\log |\mathrm{f}(\mathrm{z})|$. Here at the zeros of $\mathrm{f}(\mathrm{z}), \Phi(\mathrm{z})$ has poles and takes the value $-\infty$ there. In every closed disk contained in $D$ there are at most a finite number of points where $\log f(z)=-\infty$.

Now let a $\varepsilon D$ be any point at which $f(z)$ is distinct from zero. Since $f(z)$ is analytic and not identically zero, there exists a small neighbourhood of a where $f(z)$ is distinct from zero. We find that

$$
\log \mathrm{f}(\mathrm{z})=\log |\mathrm{f}(\mathrm{z})|+\mathrm{i} \arg \mathrm{f}(\mathrm{z})
$$

is analytic in this neighbourhood and hence $\log |\mathrm{f}(\mathrm{z})|$ is harmonic there and we have the equality

$$
\begin{equation*}
\Phi(\mathrm{a})=\frac{1}{2 \pi} \int_{0}^{2 \pi} \Phi\left(\mathrm{a}+\mathrm{re}^{\mathrm{i} \theta}\right) \mathrm{d} \theta \tag{45}
\end{equation*}
$$

for all sufficiently small values of $r$. On the otherhand, if a is a zero of $f(z)$, we have

$$
\begin{equation*}
\Phi(\mathrm{a})=-\infty<\frac{1}{2 \pi} \int_{0}^{2 \pi} \Phi\left(\mathrm{a}+\mathrm{re}^{\mathrm{i} \theta}\right) \mathrm{d} \theta \tag{46}
\end{equation*}
$$

Combining (45) with (46) we obtain $\Phi(\mathrm{z})$ is subharmonic in D.

## Unit 3 - Conformal Mappings

## Structure

### 3.0 Objectives of this Chapter

### 3.1 Conformal Mappings

### 3.2 Basic Properties of Conformal Mapping

### 3.0 Objectives of this Chapter

This chapter deals with conformal mappings and their basic properties. Many examples are given to explain different concepts on conformal mappings. The inverse function theorem is also discussed.

### 3.1 Conformal Mappings

Let X be an open set in $\mathbb{C}$ and suppose a function $\mathrm{f}: \mathrm{X} \rightarrow \mathbb{C}$ is given. We know from functional analysis that if f is continuous, a compact set of X is mapped onto a compact set in $f(X)$ and a connected set of $X$ onto a connected set of $f(X)$. If moreover, f is single-valued and analytic there occur several interesting results. In this chapter we study mappings which transform different curves and regions from one complex plane to other complex plane with reference to magnitude and orientation. Such type of mappings play an important role in the study of various physical problems defined on domains and curves of arbitrary shape.

## Level Curves

Let $\mathrm{w}=\mathrm{f}(\mathrm{z})$ with $\mathrm{z}=\mathrm{x}+$ iy and $\mathrm{w}=\mathrm{u}+\mathrm{iv}$ where $\mathrm{f}(\mathrm{z})$ is analytic. $\mathrm{u}=\mathrm{u}(\mathrm{x}, \mathrm{y})$ $\mathrm{v}=\mathrm{v}(\mathrm{x}, \mathrm{y})$ satisfy Cauchy-Riemann equations

$$
u_{x}=v_{y}, u_{y}=-v_{x}
$$

from which it follows that

$$
\begin{aligned}
& u_{x x}+u_{y y}=0 \\
& v_{x x}=v_{y y}=0
\end{aligned}
$$

Also,
$\nabla_{\mathrm{u}} \cdot \nabla_{\mathrm{v}}=0$, where


$$
\nabla=\left(\frac{\partial}{\partial x}, \frac{\partial}{\partial y}\right)
$$

So that the level curves $\mathrm{u}(\mathrm{x}, \mathrm{y})=$ constant and $\mathrm{v}(\mathrm{x}, \mathrm{y})=$ constant are orthogonal.

$$
\begin{array}{ll}
\text { Now } & f^{1}(z)=u_{x}+i v_{x}=u_{x}-i u_{y}=v_{y}+i v_{x} \\
\text { so that } & \left|f^{1}(z)\right|^{2}=u_{x}^{2}+u_{y}^{2}=v_{x}^{2}+v_{y}^{2} .
\end{array}
$$

## Two basic results :

## No. 1


uv plane
Suppose that $w=f(z)$ maps $D$ into $D^{1}$.
Let $\psi(u, v)=\psi((u(x, y), v(x, y))=\phi(x, y)$.
To prove $\phi_{\mathrm{xx}}+\phi_{\mathrm{yy}}=\left|\mathrm{f}^{1}(\mathrm{z})\right|^{2}\left(\psi_{\mathrm{uu}}+\psi_{\mathrm{vv}}\right)$
we calculate $\quad \phi_{\mathrm{x}}=\psi_{\mathrm{u}} \mathrm{u}_{\mathrm{x}}+\psi_{\mathrm{v}} \mathrm{v}_{\mathrm{x}}$

$$
\begin{aligned}
& \phi_{\mathrm{xx}}=\psi_{u \mathrm{u}} \mathrm{u}_{\mathrm{x}}^{2}+\psi_{\mathrm{vv}} v_{\mathrm{x}}^{2}+2 \psi_{\mathrm{uv}} \mathrm{u}_{\mathrm{x}} \mathrm{v}_{\mathrm{x}}+\psi_{\mathrm{u}} \mathrm{u}_{\mathrm{xx}}+\psi_{\mathrm{v}} \mathrm{v}_{\mathrm{xx}} \\
& \phi_{\mathrm{yy}}=\psi_{\mathrm{uu}} \mathrm{u}_{\mathrm{y}}^{2}+\psi_{\mathrm{vv}} v_{\mathrm{y}}^{2}+2 \psi_{\mathrm{uv}} \mathrm{u}_{\mathrm{y}} \mathrm{v}_{\mathrm{y}}+\psi_{\mathrm{u}} \mathrm{u}_{\mathrm{yy}}+\psi_{\mathrm{v}} \mathrm{v}_{\mathrm{yy}}
\end{aligned}
$$

Thus, $\phi_{x x}+\phi_{y y}=\left(\mathrm{u}_{\mathrm{x}}^{2}+\mathrm{u}_{\mathrm{y}}^{2}\right) \psi_{\mathrm{uu}}+\left(\mathrm{v}_{\mathrm{x}}^{2}+\mathrm{v}_{\mathrm{y}}^{2}\right) \psi_{\mathrm{vv}}+2 \psi_{\mathrm{uv}} \nabla_{\mathrm{u}} \cdot \nabla_{\mathrm{v}}$, since $\mathrm{u}, \mathrm{v}$ satisfy Laplace equation. Again, $\nabla_{\mathrm{u}} . \nabla_{\mathrm{v}}=0$,
so we obtain $\phi_{\mathrm{xx}}+\phi_{\mathrm{yy}}=\left|\mathrm{f}^{1}(\mathrm{z})\right|^{2}\left(\psi_{\mathrm{uu}}+\psi_{\mathrm{vv}}\right)$
Therefore if $\mathrm{f}^{1}(\mathrm{z}) \neq 0$ inside D we have $\phi_{\mathrm{xx}}+\phi_{\mathrm{yy}}=0$ imples $\psi_{\mathrm{uu}}+\psi_{\mathrm{yv}}=0$ and vice-versa.


No. 2. Consider a level curve $\mathrm{F}(\mathrm{x}, \mathrm{y})=0$ upon $\nabla \phi \cdot \underline{\mathrm{n}}=0$.
Let under the analytic mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})$ the level curve map to $\mathrm{G}(\mathrm{u}, \mathrm{v})=0$.
We shall show that $\nabla \psi \cdot \underline{n}=0$ on $G(u, v)=0$

$F(x, y)=0$
Fig. 19


Consider the map $\mathrm{w}=\mathrm{f}(\mathrm{z}) \rightarrow \omega=\mathrm{u}+\mathrm{iv}$, so $\mathrm{u}=\mathrm{u}(\mathrm{x}, \mathrm{y}), \mathrm{v}=\mathrm{v}(\mathrm{x}, \mathrm{y})$.
Suppose $f(z)$ is analytic. Then,

$$
\left.\begin{array}{l}
\phi_{x}=\psi_{u} u_{x}+\psi_{v} v_{x} \\
\phi_{y}=\psi_{u} u_{y}+\psi_{v} v_{y}
\end{array}\right\} \quad \text { so, }\binom{\phi_{x}}{\phi_{y}}=S\binom{\psi_{u}}{\psi_{v}} \text { with } S=\binom{u_{x} v_{x}}{u_{y} v_{v}}
$$

Then, $\nabla \phi=S \nabla_{\psi}, \nabla F=S \nabla G$ and clearly, $S^{T} S=\left|f^{1}(\mathrm{z})\right|^{2} 1$
Now, $\frac{\partial \phi}{\partial \underline{\mathrm{n}}}=\nabla \phi \cdot \frac{\nabla \mathrm{F}}{|\nabla \mathrm{F}|}=\frac{\mathrm{S} \nabla \psi \cdot(\mathrm{S} \nabla \mathrm{G})}{|\mathrm{S} \nabla \mathrm{G}|}=\frac{(\nabla \psi)^{\mathrm{T}} \mathrm{S}^{\mathrm{T}} \mathrm{S} \nabla \mathrm{G}}{\left\{(\mathrm{S} \nabla \mathrm{G})^{\mathrm{T}}(\mathrm{S} \nabla \mathrm{G})\right\}^{1 / 2}}=\frac{(\nabla \psi)^{\mathrm{T}} \nabla \mathrm{G}\left|\mathrm{f}^{1}(\mathrm{z})\right|}{\left\{(\nabla \mathrm{G})^{\mathrm{T}} \nabla \mathrm{G}\right\}^{1 / 2}}$
(where the usual vector operations, $\underline{a} \cdot \underline{b}=a^{T} b$ and $(a \cdot a)^{1 / 2}=\left(a^{T} a\right)^{1 / 2}=|a|$ have been used)

So, $\quad \frac{\partial \phi}{\partial \underline{\mathrm{n}}}=\nabla \phi \cdot \frac{\nabla \mathrm{F}}{|\nabla \mathrm{F}|}=\left|\mathrm{f}^{1}(\mathrm{z})\right| \nabla \psi \frac{\nabla \mathrm{G}}{|\nabla \mathrm{G}|}=\left|\mathrm{f}^{1}(\mathrm{z})\right| \frac{\partial \psi}{\partial \underline{\mathrm{n}}}$
This shows that if $\frac{\partial \phi}{\partial \underline{\mathrm{n}}}=0$ on the boundary of D then $\frac{\partial \psi}{\partial \underline{\mathrm{n}}}=0$ on the boundary of $D^{1}$, provided $\left|\mathrm{f}^{1}(\mathrm{z})\right| \neq 0$ on the boundary of D .

Note : These give us a means of transforming the domain over which the Laplace's equation is to be solved comfortably. Such type of things is usually dealt in solving boundary value problems in potential theory.

## Angle of Rotation

Given a function of a complex variable $\mathrm{w}=\mathrm{f}(\mathrm{z})$ analytic in a domain D . Let $\mathrm{z}_{0}$ be any point lying within $\mathrm{D}, \gamma: \mathrm{z}=\sigma(\mathrm{t})$, $\mathrm{a} \leq \mathrm{t} \leq \mathrm{b}, \sigma\left(\mathrm{t}_{0}\right)=\mathrm{z}_{0}$, be a curve passing
through $\mathrm{z}_{0}$ (and lying within D ). The function $\sigma(\mathrm{t})$ has a non zero derivative $\sigma^{1}\left(\mathrm{t}_{0}\right)$ at the point $\mathrm{z}_{0}$ and the curve $\gamma$ has a tangent at this point with a slope equal to $\operatorname{Arg} \sigma^{1}\left(\mathrm{t}_{0}\right)$.


Fig. 21


Fig. 22

Under the mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})$ the curve $\gamma$ is transformed into a curve $\Gamma: \mathrm{w}=\mathrm{f}(\sigma(\mathrm{t}))$ $=\mu(\mathrm{t}), \mathrm{a} \leq \mathrm{t} \leq \mathrm{b}, \mu\left(\mathrm{t}_{0}\right)=\mathrm{f}\left(\mathrm{z}_{0}\right)=\mathrm{w}_{0}$ in the w-plane. $\mu(\mathrm{t})$ is differentiable at $\mathrm{t}=\mathrm{t}_{0}$ and the curve $\Gamma$ has a tangent at $w_{0}=f\left(z_{0}\right)$. Then following the chain rule for differentiation of composite functions, assuming $\mathrm{f}^{1}\left(\mathrm{z}_{0}\right) \neq 0$

$$
\mu^{1}\left(\mathrm{t}_{0}\right)=\mathrm{f}^{1}\left(\sigma\left(\mathrm{t}_{0}\right) \sigma^{1}\left(\mathrm{t}_{0}\right)\right.
$$

It follows that

$$
\begin{array}{ll} 
& \operatorname{Arg} \mu^{1}\left(\mathrm{t}_{0}\right)=\operatorname{Arg} \mathrm{f}^{1}\left(\mathrm{z}_{0}\right)+\operatorname{Arg} \sigma^{1}\left(\mathrm{t}_{0}\right) \\
\text { i.e., } & \operatorname{Arg} \mu^{1}\left(\mathrm{t}_{0}\right)=\operatorname{Arg} \sigma^{1}\left(\mathrm{t}_{0}\right)+\operatorname{Arg} \mathrm{f}^{1}\left(\mathrm{z}_{0}\right) \tag{47}
\end{array}
$$

This implies that change in slope of a curve at a point under a transformation depends only on the point and not on the particular curve through that point.

Example 1. Verify the result given in equation (47) for the curve $y=x^{2}$ under the transformation $\mathrm{f}(\mathrm{z})=\mathrm{z}^{2}$ at $\mathrm{z}=1+\mathrm{i}$.

Solution. First we calculate the change in slope of the curve $y=x^{2}$ at the given point under the transformation $w \equiv f(z)=z^{2}$. Following the formula given in eq. (47)

$$
\operatorname{Arg} f^{1}(1+i)=\operatorname{Arg} 2(1+i)=\tan ^{-1} 1
$$

A parametric form of the given curve $y=x^{2}$ is given by

$$
\gamma: \mathrm{z}=\mathrm{t}+\mathrm{it}^{2},-\infty<\mathrm{t}<\infty .
$$

Here $z_{0}=1+i$ at $t_{0}=1$ and $z^{1}(1)=1+2 \mathrm{i}$, so that slope of the curve $\gamma$ is $\tan ^{-1} 2$.

Now we find slope of the transformed curve.

$$
\mathrm{w}=\mathrm{f}(\mathrm{z}) \Rightarrow \mathrm{u}+\mathrm{iv}=(\mathrm{x}+\mathrm{iy})^{2}
$$

So, $u=x^{2}-y^{2}$ and $v=2 x y=2 x \cdot x^{2}=2 x^{3}$.

Then, $u=x^{2}-x^{4}=\left(\frac{v}{2}\right)^{2 / 3}-\left(\frac{v}{2}\right)^{4 / 3}$, which is the equation of the transformed curve $\Gamma$. The image of the point $(1+\mathrm{i})$ of z -plane is the point 2 i in the w-plane and the slope of the curve $\Gamma$ at $w=2 \mathrm{i}$ is

$$
\left.\frac{\mathrm{dv}}{\mathrm{du}}\right|_{\mathrm{w}=2 \mathrm{i}}=-3
$$

Thus the change in slope of the curve $\gamma$ under the transformation is

$$
\tan ^{-1}(-3)-\tan ^{-1}(2)=\tan ^{-1} \frac{-3-2}{1-6}=\tan ^{-1} 1
$$

which is the same as obtained earlier following equation (47).
Definition : A mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})$ is said to be conformal at a point $\mathrm{z}=\mathrm{z}_{0}$, if it preserves angles between oriented curves, passing through $\mathrm{z}_{0}$, in magnitude and in sense of rotation.

Theorem 3.1 : Let $\mathrm{f}(\mathrm{z})$ be an analytic function in a domain D containing $\mathrm{z}_{0}$. If $\mathrm{f}^{1}\left(\mathrm{z}_{0}\right) \neq 0$, then $\mathrm{f}(\mathrm{z})$ is conformal at $\mathrm{z}_{0}$.
Proof. Let $\mathrm{C}_{1}: \mathrm{z}=\mathrm{z}_{1}(\mathrm{t})$ and $\mathrm{C}_{2}: \mathrm{z}=\mathrm{z}_{2}(\mathrm{t}), \mathrm{t} \equiv$ parameter, be two curves which intersect at some $t=t_{0}$ where $z_{1}\left(\mathrm{t}_{0}\right)=\mathrm{z}_{2}\left(\mathrm{t}_{0}\right)=\mathrm{z}_{0}, \mathrm{C}_{1}^{1}, \mathrm{C}_{2}^{1}$ are their images under the mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})$.


Fig. 21
tangent lines are
$\mathrm{z}^{1}=\mathrm{z}_{1}{ }^{1}\left(\mathrm{t}_{0}\right), \mathrm{z}^{1}=\mathrm{z}_{2}{ }^{1}\left(\mathrm{t}_{0}\right)$ at $\mathrm{t}=\mathrm{t}_{0}$


Fig. 22
tangent lines are

$$
\begin{aligned}
& \mathrm{w}_{1}{ }^{1}\left(\mathrm{t}_{0}\right)=\mathrm{f}^{1}\left(\mathrm{z}_{1}\left(\mathrm{t}_{0}\right) \mathrm{z}_{1}{ }^{1}\left(\mathrm{t}_{0}\right)\right. \\
& \mathrm{w}_{2}^{1}\left(\mathrm{t}_{0}\right)=\mathrm{f}^{1}\left(\mathrm{z}_{2}\left(\mathrm{t}_{0}\right) \mathrm{z}_{2}\left(\mathrm{t}_{0}\right) \mathrm{z}_{2}{ }^{1}\left(\mathrm{t}_{0}\right)\right.
\end{aligned}
$$

Then following the result given in eq. (47)

$$
\begin{aligned}
& \operatorname{Arg}\left(w_{1}^{1}\left(\mathrm{t}_{0}\right)\right)-\operatorname{Arg}\left(\mathrm{z}_{1}^{1}\left(\mathrm{t}_{0}\right)\right)=\operatorname{Arg}\left(\mathrm{f}^{1}\left(\mathrm{z}_{1}\left(\mathrm{t}_{0}\right)\right)=\operatorname{Arg}\left(\mathrm{f}^{1} \mathrm{z}_{0}\right)\right) \\
& \operatorname{Arg}\left(\mathrm{w}_{2}^{1}\left(\mathrm{t}_{0}\right)\right)-\operatorname{Arg}\left(\mathrm{z}_{2}^{1}\left(\mathrm{t}_{0}\right)\right)=\operatorname{Arg}\left(\mathrm{f}^{1}\left(\mathrm{z}_{2}\left(\mathrm{t}_{0}\right)\right)=\operatorname{Arg}\left(\mathrm{f}^{1} \mathrm{z}_{0}\right)\right)
\end{aligned}
$$

and

Subtracting, $\operatorname{Arg}\left(\mathrm{w}_{1}^{1}\left(\mathrm{t}_{0}\right)\right)-\operatorname{Arg}\left(\mathrm{w}_{2}^{1}\left(\mathrm{t}_{0}\right)\right)-\left\{\operatorname{Arg}\left(\mathrm{z}_{1}^{1}\left(\mathrm{t}_{0}\right)\right)-\operatorname{Arg}\left(\mathrm{z}_{2}^{1}\left(\mathrm{t}_{0}\right)\right)\right\}=0$
i.e., $\theta=\phi$, where $\quad \theta=$ angle between the curves $\mathrm{C}_{1}$ and $\mathrm{C}_{2}$ at $\mathrm{z}_{0}$ and

$$
\phi=\text { angle between the curves } \mathrm{C}_{1}^{1} \text { and } \mathrm{C}_{2}^{1} \text { at } \mathrm{w}_{0} \text {. }
$$

Observation : From the basic results proved earlier we learn that if $f$ is a conformal mapping, then orthogonal curves are mapped onto orthogonal curves.

### 3.2 Basic Properties of conformal Mappings

Let $f(z)$ be an analytic function in a domain $D$, and let $z_{0}$ be a point in $D$. If $f\left(z_{0}\right)$ $=0$, then we can express $f(z)$ in the form

$$
f(z)=f\left(z_{0}\right)+\left(z-z_{0}\right) f^{1}\left(z_{0}\right)+\left(z-z_{0}\right) \eta(z)
$$

where $\eta(z) \rightarrow 0$ as $z \rightarrow z_{0}$. If $z$ is near $z_{0}$, then the transformation $w=f(z)$ has the linear approximation

$$
\mathrm{G}(\mathrm{z})=\mathrm{A}+\mathrm{B}\left(\mathrm{z}-\mathrm{z}_{0}\right) .
$$

where $A=f\left(z_{0}\right)$ and $B=f^{1}\left(z_{0}\right)$. As $\eta(z) \rightarrow 0$ when $z \rightarrow z_{0}$, for points near $z_{n}$ the transformation $w=f(z)$ has an effect much like the linear mapping $w=G(z)$. The effect of the linear mapping $G$ is a rotation of the plane through the angle $\alpha=\operatorname{Arg}$ $\left(\mathrm{f}^{1}\left(\mathrm{z}_{0}\right)\right.$ ), followed by a magnification by the factor $\left|\mathrm{f}\left(\mathrm{z}_{0}\right)\right|$, followed by a translation by the vector $\mathrm{A}+\mathrm{BZ}_{0}$.

Remark : If $\mathrm{f}^{1}\left(\mathrm{z}_{0}\right)=0$, the angle may not be preserved.
Let us consider, $w=f(z)=z^{2}$, then we have $f^{1}(0)=0$ and


Fig. 23


Fig. 24
the angle at $\mathrm{z}=0$ is not preserved but is doubled.
Definition : Let $f(z)$ be a nonconstant analytic function. If $f^{1}\left(z_{0}\right)=0$, the $z_{0}$ is called a critical point of $f(z)$, and the mapping $w=f(z)$ is not conformal at $z_{0}$. We shall see afterwards what happens at a critical point.

The Inverse Function theorem 3.2 Let $f(z)$ be analytic at $z_{0}$ and $\mathrm{f}^{1}\left(\mathrm{z}_{0}\right) \neq 0$. Then there exists a neighbourhood $N\left(w_{0}, \varepsilon\right)$ of $w_{0}=f\left(Z_{0}\right)$ in which the inverse function $\mathrm{z}=\mathrm{F}(\mathrm{w})$ exists and is analytic.

Moreover, $\mathrm{F}^{1}\left(\mathrm{w}_{0}\right)=1 / \mathrm{f}^{1}\left(\mathrm{z}_{0}\right)$.
Proof : Given $w=f(z),(z=x+i y, w=u+i v)$
is analytic in a neighbourhood of $\mathrm{z}_{0}, \mathrm{~K}:\left|\mathrm{z}-\mathrm{z}_{0}\right|<\rho$. We shall show that for each $w \in L:\left|w-w_{0}\right|<\in$ there is a unique solution $z=F(w)$, where $z \in K$.

We express the mapping $w=f(z)$ in terms of the set of equations

$$
\begin{equation*}
\mathrm{u}=\mathrm{u}(\mathrm{x}, \mathrm{y}) \text { and } \mathrm{v}=\mathrm{v}(\mathrm{x}, \mathrm{y}) \tag{49}
\end{equation*}
$$

which represents a transformation from the xy plane to the uv plane, $\mathrm{u}, \mathrm{v}$, possess continuous first-order partial derivatives satisfying C-R equations. The Jacobian determinant $\mathrm{J}(\mathrm{x}, \mathrm{y})$, is defined by

$$
J(x, y)=\left|\begin{array}{cc}
u_{x} & u_{y}  \tag{50}\\
v_{x} & v_{y}
\end{array}\right|
$$

The transformation in equations (49) has a local inverse in $L$ provided $\mathrm{J}(\mathrm{x}, \mathrm{y}) \neq 0$ in $K$ [(3) pp. 358-361]. Expanding r.h.s. of equation (50) and using the C-R equations, we obtain

$$
\begin{align*}
& \mathrm{J}\left(\mathrm{x}_{0}, \mathrm{y}_{0}\right)=\mathrm{u}_{\mathrm{x}}^{2}\left(\mathrm{x}_{0}, \mathrm{y}_{0}\right)+\mathrm{v}_{\mathrm{x}}^{2}\left(\mathrm{x}_{0}, \mathrm{y}_{0}\right) \\
& =\left|\mathrm{f}^{1}\left(\mathrm{z}_{0}\right)\right|^{2}  \tag{51}\\
& \neq 0 \text {, by the given hypothesis. }
\end{align*}
$$

Utilising the continuity of $\mathrm{J}(\mathrm{x}, \mathrm{y})$ in a small neighbourhood of ( $\mathrm{x}_{0}, \mathrm{y}_{0}$ ), equations (49) and (51) imply that a local inverse $\mathrm{z}=\mathrm{F}(\mathrm{w})$ exists in a neighbourhood of the point $w_{0}=f\left(z_{0}\right)$. The derivative of $F(w)$ is given by the familiar expression

$$
\begin{aligned}
& \begin{aligned}
\mathrm{F}^{1}(\mathrm{w}) & =\lim _{\Delta \mathrm{w} \rightarrow 0} \frac{\mathrm{~F}(\mathrm{w}+\Delta \mathrm{w})-\mathrm{F}(\mathrm{w})}{\Delta \mathrm{w}}=\lim _{\Delta \mathrm{w} \rightarrow 0} \frac{\Delta \mathrm{z}}{\Delta \mathrm{w}}=\lim _{\Delta \mathrm{z} \rightarrow 0} \frac{\Delta \mathrm{z}}{\mathrm{f}(\mathrm{z}+\Delta \mathrm{z})-\mathrm{f}(\mathrm{z})} \\
& =\lim _{\Delta \mathrm{z} \rightarrow 0} 1 /\left(\frac{\mathrm{f}(\mathrm{z}+\Delta \mathrm{z})-\mathrm{f}(\mathrm{z})}{\Delta \mathrm{z}}\right)=1 /\left(\lim _{\Delta z \rightarrow 0} \frac{\mathrm{f}(\mathrm{z}+\Delta \mathrm{z})-\mathrm{f}(\mathrm{z})}{\Delta \mathrm{z}}\right)
\end{aligned} \\
& \text { i.e., } \mathrm{F}^{\mathrm{l}}(\mathrm{w})=\frac{1}{\mathrm{f}^{1}(\mathrm{z})}
\end{aligned}
$$

holds in a neighbourhood of the point $\mathrm{w}_{0}$, as $\mathrm{f}(\mathrm{z})$ is analytic in K .
In particular, $\mathrm{F}^{1}\left(\mathrm{w}_{0}\right)=\frac{1}{\mathrm{f}^{1}\left(\mathrm{z}_{0}\right)}$
Theorem 3.3 Let $f(z)$ be analytic at the point $z_{0}$. If $\mathrm{f}^{1}\left(\mathrm{z}_{0}\right)=0, \mathrm{f}^{11}\left(\mathrm{z}_{0}\right)=0, \ldots$,
$\mathrm{f}^{(\mathrm{k}-1)}\left(\mathrm{z}_{0}\right)=0$ and $\mathrm{f}^{(\mathrm{k})}\left(\mathrm{z}_{0}\right) \neq 0$, then the mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})$ magnifies angles at $\mathrm{z}_{0}$ by k times.

Proof. By the given hypothesis, $\mathrm{f}(\mathrm{z})$ has the Taylor expansion in a neighbourhood of $z_{0}$ in the form

$$
\mathrm{f}(\mathrm{z})=\mathrm{f}\left(\mathrm{z}_{0}\right)+\mathrm{c}_{\mathrm{k}}\left(\mathrm{z}-\mathrm{z}_{0}\right)^{\mathrm{k}}+\mathrm{c}_{\mathrm{k}+1}\left(\mathrm{z}-\mathrm{z}_{0}\right)^{\mathrm{k}+1}+\ldots, \mathrm{c}_{\mathrm{k}} \neq 0
$$

so that we can express

$$
\begin{equation*}
\mathrm{f}(\mathrm{z})-\mathrm{f}\left(\mathrm{z}_{0}\right)=\left(\mathrm{z}-\mathrm{z}_{0}\right)^{\mathrm{k}}+\mathrm{h}(\mathrm{z}) \tag{52}
\end{equation*}
$$

where $h(z)$ is analytic at $z_{0}$ and $h\left(z_{0}\right) \neq 0$. Now let $w=f(z)$ and $w_{0}=f\left(z_{0}\right)$ and we obtain from (52)

$$
\operatorname{Arg}\left(\mathrm{w}-\mathrm{w}_{0}\right)=\mathrm{k} \operatorname{Arg}\left(\mathrm{z}-\mathrm{z}_{0}\right)+\operatorname{Arg}(\mathrm{h}(\mathrm{z}))
$$

Let $\mathrm{z} \rightarrow \mathrm{z}_{0}$ along a curve $\gamma$. Then $\mathrm{w} \rightarrow \mathrm{w}_{0}$ along the image curve $\Gamma$ and the slope of tangent to the curve $\gamma$ at $z_{0}$ and that of the tangent to the curve $\Gamma$ at $w_{0}$ are connected by the relation

$$
\begin{aligned}
& \text { lim } \lim _{\mathrm{w} \rightarrow \mathrm{w}_{0}} \operatorname{Arg}\left(\mathrm{w}-\mathrm{w}_{0}\right)=\mathrm{k} \lim _{\mathrm{z} \rightarrow \mathrm{z}_{0}} \operatorname{Arg}\left(\mathrm{z}-\mathrm{z}_{0}\right)+\lim _{\mathrm{z} \rightarrow \mathrm{z}_{0}} \operatorname{Arg}(\mathrm{~h}(\mathrm{z})) \\
& \text { i.e., }
\end{aligned} \theta_{0}=\mathrm{k} \phi_{0}+\operatorname{Arg}(\mathrm{h}(\mathrm{z}))
$$

Thus, if $\gamma_{1}$ and $\gamma_{2}$ be two curves passing through $\mathrm{z}_{0}$ and their images $\Gamma_{1}$ and $\Gamma_{2}$ under the mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})$, pass through $\mathrm{w}_{0}$, the difference of slopes of the curves $\gamma_{1}$ and $\gamma_{2}$ at $\mathrm{z}_{0}$ and that of the curves $\Gamma_{1}$ and $\Gamma_{2}$ at $\mathrm{w}_{0}$ are related as

$$
\theta_{2}-\theta_{1}=\mathrm{k}\left(\phi_{2}-\phi_{1}\right)
$$

with the sense remain unchanged.
Example 2. Show that the mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})=\mathrm{z}^{2}$ maps the rectangle
$R=\left\{x+\right.$ iy: $\left.-1 \leq x \leq 1,0 \leq y \leq \frac{1}{2}\right\}$ of unit area onto the region enclosed by the parabolas

$$
\mathrm{v}^{2}=\mathrm{u}+\frac{1}{4} \text { and } \mathrm{v}^{2}=-4(\mathrm{u}-1)
$$

Solution : Here $\mathrm{f}^{1}(\mathrm{z})=2 \mathrm{z}$ and the mapping $\mathrm{w}=\mathrm{z}^{2}$ is conformal for all $\mathrm{z} \neq 0$. We note that the right angles at the vertices $z_{1}=1, z_{2}=1+i / 2, z_{3}=-1+i / 2$ and $\mathrm{z}_{4}=-1$ are mapped into right angles at the vertices $\mathrm{w}_{1}=1, \mathrm{w}_{2}=\frac{3}{4}+\mathrm{i}, \mathrm{w}_{3}=\frac{3}{4}-\mathrm{i}$ and $\mathrm{w}_{4}=1$ respectively.


Fig. 26

The parabolas shown in the figure are obtained as follows :
Let $w=u+i v$. Then $\left.u=x^{2}-y^{2}, v=2 x y\right\} \ldots$
The line $\mathrm{x}=1$ corresponds to the curve $\mathrm{u}=1-\mathrm{y}^{2}, \mathrm{v}=2 \mathrm{y}$. Eliminating y , we get $v^{2}=-4(u-1)$, which is a parabola with vertex $(1,0)$ and opens towards the negative side of the $u$-axis in the w-plane. Also, the part of the line $x=1$ lying above the real axis corresponds to the part of the parabola lying above the $u$-axis in the w-plane. The same parabola in the w-plane is the image of the line $x=-1$. In this case, the part of the line $\mathrm{x}=-1$ lying above the real axis corresponds to the part of the parabola lying below the u -axis in the w -plane.

Again, when $\mathrm{y}=\frac{1}{2}$, from (53) we get $\mathrm{u}=\mathrm{x}^{2}-\frac{1}{4}$ and $\mathrm{v}=\mathrm{x}$. Eliminating x we get, $\mathrm{v}^{2}=\mathrm{u}+\frac{1}{4}$ which is also a parabola with vertex $\left(-\frac{1}{4}, 0\right)$ and opening towards the positive side of the $u$-axis in the w-plane. By similar argument as before we can say that the mapping $w=z^{2}$ maps the rectangle $R=\left\{x+i y:-1 \leq x \leq 1,0 \leq y \leq \frac{1}{2}\right\}$ onto the region enclosed by the parabolas $v^{2}=u+\frac{1}{4}$ and $v^{2}=-4(u-1)$.

Note : It is not hard to prove that the parabolas intersect each other orthogonally at $w_{2}$ and $w_{3}$.

At the point $\mathrm{z}_{0}=0$, we have $\mathrm{f}^{1}\left(\mathrm{z}_{0}\right)=\mathrm{f}^{1}(0)=0$ and $\mathrm{f}^{11}\left(\mathrm{z}_{0}\right)=2 \neq 0$. Hence the angles at the origin $\mathrm{z}_{0}=0$ are magnified by the factor $\mathrm{k}=2$. In particular the straight angle at $\mathrm{z}_{0}=0$ is mapped onto $2 \pi$ angle at $\mathrm{w}_{0}=0$.

## Unit 4 Multi-valued functions and Riemann Surface

## Structure

### 4.0 Objectives of this Chapter

4.1 Multi-valued functions
4.2 The logarithm function
4.3 Properties of $\log z$
4.4 Branch, Branch point and Branch cut
4.5 Integrals of Multi-valued function
4.6 Branch points at infinity
4.7 Detection of branch points
4.8 The Riemann Surface for $\mathbf{w}=\mathbf{z}^{1 / 2}$
4.9 Concept of neighbourhood
4.10 The Riemann Surface for $w=\log z$

### 4.11 The Inverse Trigonometric Functions

### 4.0 Objectives of this Chapter

In this chapter we shall study multi-valued functions and their Riemann surfaces. In particular, multi-valued logarithm function, the power function $z^{\alpha}$ both $\mathrm{z}, \alpha$ complex numbers, $\mathrm{z} \neq 0$ will be discussed. The ideas of branch, branch point, branch cut, branch point at infinity will be explained by means of different examples. A few contour integrations of multi-valued functions will be performed. Also Riemann surfaces for different multivalued functions will be constructed.

### 4.1 Multi-valued functions

So far we have considered single-valued functions i.e., one-to-one mapping or, many-to-one mapping. In the later case, under certain restrictions, inverse mappings give rise to multi-valued functions i.e., one-to-many.

For example,

$$
\mathrm{z}=\mathrm{e}^{\omega}, \mathrm{z}=\omega^{2}, \mathrm{z}=\sin \omega, \mathrm{z}=\cos \omega
$$

For each of these functions, a given value of z corresponds to more than one value of $\omega$.

$\omega=\mathrm{f}^{-1}(\mathrm{z})$ is multi-valued and $\mathrm{z}=\mathrm{f}(\omega)$ is single-valued, given $\omega$, there is a unique value of $z$.

The aim of this chapter is as follows :
(i) To determine all possible values of the inverse function $\omega$ and (ii) To construct an inverse function which is single-valued in some region of the complex plane.

Let $\omega=f(z)$ be a multi-valued function. A branch of $f$ is any single-valued function $\mathrm{f}_{0}$ that is continuous in some domain (except, perhaps, on the boundary). At each point $z$ in the domain, it assigns one of the values of $f(z)$.

Example 1 : We consider branches of the two-valued square-root function $f(z)$ $=z^{1 / 2}(z \neq 0)$. The principal branch of the square root function is

$$
\mathrm{f}_{1}(\mathrm{z})=|\mathrm{z}|^{1 / 2} \mathrm{e}^{\mathrm{i} \theta / 2}=\mathrm{r}^{1 / 2}\left(\cos \frac{\theta}{2}+\sin \frac{\theta}{2}\right), \theta=\operatorname{Arg}(\mathrm{z})
$$

where $\mathrm{r}=|\mathrm{z}|$ and $-\pi<\theta \leq \pi$. The function $\mathrm{f}_{1}$ is a branch of f . Using the same notation, we can find other branches of the function $f$. For example if we let

$$
f_{2}(z)=|z|^{1 / 2} e^{i(\theta+2 \pi) / 2}=r^{1 / 2}\left[\cos \left(\frac{\theta+2 \pi}{2}\right)+i \sin \left(\frac{\theta+2 \pi}{2}\right)\right]
$$

then

$$
\mathrm{f}_{2}(\mathrm{z})=\mathrm{r}^{1 / 2} \mathrm{e}^{\mathrm{i}(\theta+2 \pi) / 2}=\mathrm{r}^{1 / 2} \mathrm{e}^{\mathrm{i} \theta / 2} \cdot \mathrm{e}^{\mathrm{i} \pi}=-\mathrm{f}_{1}(\mathrm{z})
$$

So, $f_{1}$ and $f_{2}$ can be taken as the two branches of the multi-valued square root function. The negative real axis is called a branch cut for the functions $f_{1}$ and $f_{2}$. Each point on the branch cut is a point of discontinuity for both functions $f_{1}$ and $f_{2}$.

Result 1 : Show that the function $f_{1}$ is discontinuous on the negative real axis.

Solution : Let $\mathrm{z}_{0}=\mathrm{r}_{0} \mathrm{e}^{\mathrm{i} \pi}$ be any point on the negative real axis. We compute the limit as z approaches $\mathrm{z}_{0}$ through the upper half plane $\operatorname{lm} \mathrm{z}>0$ and the limit as z approaches $\mathrm{z}_{0}$ through the lower half plane $1 \mathrm{~m} \mathrm{z}<0$. The limits are

$$
\begin{aligned}
& \underset{(\mathrm{r}, \theta) \rightarrow\left(\mathrm{r}_{0}, \pi\right)}{\lim \mathrm{f}_{1}\left(\mathrm{re} \mathrm{i}^{\mathrm{i}}\right)}=\underset{(\mathrm{r}, \theta) \rightarrow\left(\mathrm{r}_{0}, \pi\right)}{ } \lim _{\mathrm{r}} \mathrm{r}^{1 / 2}\left[\cos \frac{\theta}{2}+\mathrm{i} \sin \frac{\theta}{2}\right]=\mathrm{ir}_{0}^{1 / 2} \text {, and } \\
& (\mathrm{r}, \theta) \xrightarrow[\left(\mathrm{r}_{0},-\pi\right)]{ } \mathrm{f}_{1}\left(\mathrm{re}^{\mathrm{i} \theta}\right)=(\mathrm{r}, \theta) \xrightarrow[\left(\mathrm{r}_{0},-\pi\right)]{ } \mathrm{r}^{1 / 2}\left[\cos \frac{\theta}{2}+\mathrm{i} \sin \frac{\theta}{2}\right]=-\mathrm{ir}_{0}^{1 / 2}
\end{aligned}
$$

The two limits are distinct, so the function $\mathrm{f}_{1}$ is discontinuous at $\mathrm{z}_{0}$. Since $\mathrm{z}_{0}$ is an arbitrary point on the negative real axis, $\mathrm{f}_{1}$ is discontinous there.

Note : Likewise, $\mathrm{f}_{2}$ is discontinuous at $\mathrm{z}_{0}$.


Fig. 28 a


Fig. 29 a


Fig. 28 b


Fig. 29 b

Figures : 28-29 The Branches $f_{1}$ and $f_{2}$ of $f(z)=z^{1 / 2}$

### 4.2 The logarithm function

Let us define the inverse function $\mathrm{f}^{-1}(\mathrm{z})$ for $\mathrm{z}=\mathrm{e}^{\omega}$ : Let $\mathrm{z}=\mathrm{re}^{\mathrm{i} \theta}$ and $\omega=\mathrm{u}+\mathrm{iv}$. Then

$$
\mathrm{re}^{\mathrm{i} \theta}=\mathrm{e}^{\mathrm{u}} \cdot \mathrm{e}^{\mathrm{iv}}
$$

So that

$$
\mathrm{r}=\mathrm{e}^{\mathrm{u}} \text { and } \mathrm{v}=\theta+2 \mathrm{k} \pi, \mathrm{k}=0, \pm 1, \pm 2, \ldots
$$

and

$$
\omega=\log \mathrm{r}+\mathrm{i}(\theta+2 \mathrm{k} \pi), \mathrm{k}=0, \pm 1, \pm 2, \ldots
$$

But $r=|z|$ and without loss of generality, we can take $\theta \in(-\pi, \pi)$. This motivates the definition of the inverse function $f^{-1}(z)$ for $z=e^{\omega}$

$$
\omega=\log \mathrm{z}=\log |\mathrm{z}|+\mathrm{i}(\operatorname{Arg} \mathrm{z}+2 \mathrm{k} \pi), \mathrm{k}=0, \pm 1, \pm 2, \ldots
$$

or, equivalently

$$
\omega=\log \mathrm{z}=\log |\mathrm{z}|+\mathrm{i} \arg \mathrm{z} .
$$

Mapping of the strip $|\operatorname{Im} \omega|<\pi$ under $z=e^{\omega}$


Fig. 30
I. Take $u=u_{0}>0, v \in(-\pi, \pi)$ for the line PQ :

$$
\begin{aligned}
& x+i y=e_{0}^{u}(\cos v+i \sin v) \\
& \left.\Rightarrow \begin{array}{l}
x=e^{u_{0}} \cos v \\
y=e^{u_{0}} \sin v
\end{array}\right\} \rightarrow x^{2}+y^{2}=e^{2 u_{0}}>1,
\end{aligned}
$$

a full circle in z-plane outside $|\mathrm{z}|=1$.
Now approach $\mathrm{Q} ; \mathrm{u}=\mathrm{u}_{0}>0, v=-\pi+\varepsilon$
$\mathrm{x}=\mathrm{e}^{\mathrm{u}_{0}} \cos (-\pi+\varepsilon) \rightarrow-\mathrm{e}^{\mathrm{u}_{0}}$ as $\varepsilon \rightarrow 0+$ and $-\mathrm{e}^{\mathrm{u}_{0}}<-1$ as $\mathrm{u}_{0}>0$
$\mathrm{y}=\mathrm{e}^{\mathrm{u}_{0}} \sin (-\pi+\varepsilon) \rightarrow 0-$ as $\varepsilon \rightarrow 0+$
Now approach $\mathrm{P}: \mathrm{u}=\mathrm{u}_{0}>0, \mathrm{v}=\pi-\varepsilon$

$$
\begin{gathered}
\mathrm{x}=\mathrm{e}^{\mathrm{u}_{0}} \cos (\pi-\varepsilon) \rightarrow-\mathrm{e}^{\mathrm{u}_{0}} \text { as } \varepsilon \rightarrow 0+ \\
\mathrm{y}=\mathrm{e}^{\mathrm{u}_{0}} \sin (\pi-\varepsilon) \rightarrow 0+\text { as } \varepsilon \rightarrow 0+
\end{gathered}
$$

II. Now take $u=u_{0}<0, v \in(-\pi, \pi)$ for the line RS :
$\left.\Rightarrow \quad \begin{array}{l}x=e^{-u_{0}} \cos v \\ y=e^{-u_{0}} \sin v\end{array}\right\} \rightarrow x^{2}+y^{2}=e^{-2 u_{0}}<1$
represents a full circle in z-plane inside $|\mathrm{z}|<1$.
Approach

$$
\begin{aligned}
& S: u=-\mathrm{u}_{0}<0, v=-\pi+\varepsilon \\
& \mathrm{x}=\mathrm{e}^{-\mathrm{u}_{0}} \cos (-\pi+\varepsilon) \rightarrow-\mathrm{e}^{-\mathrm{u}_{0}}>-1 \text { as } \varepsilon \rightarrow 0+ \\
& \mathrm{y}=\mathrm{e}^{-\mathrm{u}_{0}} \sin (-\pi+\varepsilon) \rightarrow 0-\text { as } \varepsilon \rightarrow 0+ \\
& \mathrm{R}: \mathrm{u}=-\mathrm{u}_{0}<0, \mathrm{v}=\pi-\varepsilon \\
& \mathrm{x}=\mathrm{e}^{-\mathrm{u}_{0}} \cos (\pi-\varepsilon) \rightarrow-\mathrm{e}^{-\mathrm{u}_{0}}>-1 \text { as } \varepsilon \rightarrow 0+ \\
& \mathrm{y}=\mathrm{e}^{-\mathrm{u}_{0}} \sin (\pi-\varepsilon) \rightarrow 0 \text { as } \varepsilon \rightarrow 0+
\end{aligned}
$$

Now approach $\quad \mathrm{R}: \mathrm{u}=-\mathrm{u}_{0}<0, \mathrm{v}=\pi-\varepsilon$

Observation : Points along the negative real axis in the z-plane yield multiple $w$ values. In order to obtain a single-valued inverse function for the fundamental strip $|\operatorname{lm} \omega|<\pi$ we require a cut in z-plane along $\operatorname{Re} \mathrm{z}<0$. The mapping $\mathrm{z}=\mathrm{e}^{\mathrm{w}}$ and $\mathrm{w}=\mathrm{f}^{-1}(\mathrm{z})$ will be single-valued in $|\operatorname{lm} \mathrm{w}|<\pi$ and $\mathrm{z} \in \mathbb{C} \backslash(\infty, 0)$.

Clearly the inverse function $\quad \mathrm{w}=\log \mathrm{z}=\log |\mathrm{z}|+\operatorname{iArg} \mathrm{z},-\pi<\operatorname{Arg} \mathrm{z} \leq \pi$
is single-valued. We call this function the


Fig. 31 principal value of $\log \mathrm{z}$.

The principal value of $\log \mathrm{z}$ is not defined at $\mathrm{z}=0$ and is discontinuous as z approach the negative real axis from top and bottom. Using the necessary and sufficient conditions for differentiability we find
$\frac{\mathrm{d}}{\mathrm{dz}} \log \mathrm{z}=\frac{1}{\mathrm{z}}, \mathrm{z} \neq 0, \mathrm{z} \notin(-\infty, 0)$
The point $\mathrm{z}=0$ is called a branch point of $\log \mathrm{z}$ since if we encircle the origin $\mathrm{z}=0$ by a closed contour then $\log \mathrm{z}$ changes by an amount proportional to $2 \pi \mathrm{i}$.

### 4.3 Properties of $\log z$

(i) $\log \left(\mathrm{z}_{1} \mathrm{z}_{2}\right)=\log \mathrm{z}_{1}+\log \mathrm{z}_{2}$
(means that the set of all values of $\log \mathrm{z}_{1}+\log \mathrm{z}_{2}$ is the same as the set of all values of $\left.\log \left(\mathrm{z}_{1} \mathrm{z}_{2}\right)\right)$.
(ii) $\mathrm{z}=\mathrm{e}^{\log \mathrm{z}}$, but $\log \left(\mathrm{e}^{\mathrm{z}}\right)=\mathrm{z}+2 \mathrm{k} \pi \mathrm{i}, \mathrm{k}=0 \pm 1, \pm 2, \ldots$

Let $\mathrm{z}=\mathrm{x}+\mathrm{iy}$

$$
\begin{aligned}
\log \mathrm{e}^{\mathrm{x}+\mathrm{iy}} & =\log \left(\mathrm{e}^{\mathrm{x}}\right)+\mathrm{i}\left(\tan ^{-1}\left(\frac{\sin \mathrm{y}}{\cos \mathrm{y}}\right)+2 \mathrm{k} \pi\right)+\mathrm{x}+\mathrm{iy}=2 \mathrm{k} \pi \mathrm{i} \\
& =\mathrm{z}+2 \mathrm{k} \pi \mathrm{i}, \mathrm{k}=0, \pm 1, \ldots
\end{aligned}
$$

(iii) $\log \mathrm{z}^{\mathrm{n}} \neq \mathrm{n} \log \mathrm{z}$ in general.

Let $\mathrm{z}=\mathrm{re}^{\mathrm{i} \theta}$
$\log \mathrm{z}^{\mathrm{n}}=\mathrm{n} \log \mathrm{r}+\mathrm{i}(\mathrm{n} \theta+2 \mathrm{k} \pi), \mathrm{k}=0, \pm 1, \ldots$
$\mathrm{n} \log \mathrm{z}=\mathrm{n} \log \mathrm{r}+\operatorname{in}(\theta+2 \mathrm{~m} \pi), \mathrm{m}=0, \pm 1, \ldots$
Let n be fixed. Then the set of values of $\{\mathrm{k}\}, \mathrm{k}=0, \pm 1, \pm 2, \ldots$
do not coincide with the set of values of $\{\mathrm{mn}\}, \mathrm{m}=0, \pm 1, \pm 2, \ldots$

$$
\Rightarrow \log \mathrm{z}^{\mathrm{n}} \neq \mathrm{n} \log \mathrm{z}
$$

(iv) $\log \left(\mathrm{z}^{1 / \mathrm{n}}\right)=\frac{1}{\mathrm{n}} \log \mathrm{z}$ (provided the set of values are the same) $\mathrm{n} \equiv+$ ve integer.

Now, $z=r e^{i \theta}, z^{1 / n}=r^{1 / n} e^{i(\theta+2 k \pi) / n}, k=0,1,2, \ldots, n-1$
$\log \mathrm{z}^{1 / \mathrm{n}}=\frac{1}{\mathrm{n}} \log \mathrm{r}+\mathrm{i}\left(\frac{\theta+2 \mathrm{k} \pi}{\mathrm{n}}+2 \ell \pi\right), \mathrm{k}=0,1, \ldots, \mathrm{n}-1 ; \ell=0, \pm 1, \pm 2, \ldots$
Again, $\quad \frac{1}{\mathrm{n}} \log \mathrm{z}=\frac{1}{\mathrm{n}} \log \mathrm{r}+\mathrm{i}\left(\frac{\theta}{\mathrm{n}}+\frac{2 \mathrm{~m} \pi}{\mathrm{n}}\right), \mathrm{m}=0, \pm 1, \pm 2, \ldots$
The set of values of $\log \left(z^{1 / n}\right)$ and $1 / n \log z$ are the same if the sets $\{k+\ln \}$, $\mathrm{k}=0,1, \ldots, \mathrm{n}-1 ; l=0, \pm 1, \pm 2, \ldots$ coincide with the set $\{\mathrm{m}\}, \mathrm{m}=0, \pm 1, \pm 2, \ldots$

## Complex exponents

If $\alpha$ is complex and $z \neq 0$ then
$\mathrm{z}^{\alpha}=\mathrm{e}^{\alpha \log \mathrm{z}}$ multi-valued.
$\left.\mathrm{z}^{\alpha}=\mathrm{e}^{\alpha[\log |z|}+\mathrm{i}(\operatorname{Argz}+2 \mathrm{k} \pi)\right], \mathrm{k}=0, \pm 1, \pm 2, \ldots$
$=\mathrm{e}^{\alpha[\log |z|+\mathrm{i}(\theta+2 \mathrm{k} \pi)]}$
agrees with our previous results if $\alpha=\mathrm{m}, \alpha=\frac{1}{\mathrm{~m}} ; \mathrm{m}=$ integer. If $\alpha$ is a rational number say $\mathrm{p} / \mathrm{q}$, then $\mathrm{z}^{\alpha}$ will have only q number of distinct values, occurred against $\mathrm{k}=0$, $1,2, \ldots, \mathrm{q}-1$ and the values of $\mathrm{e}^{\mathrm{i} 2 \mathrm{pk} \pi / \mathrm{q}}$ for $\mathrm{k}=-1,-2, \ldots,-(\mathrm{q}-1)$ coincide with
its values for $\mathrm{k}=\mathrm{q}-1, \mathrm{q}-2, \ldots, 2,1$ respectively, whereas the values of $\mathrm{e}^{\mathrm{i} 2 \mathrm{pk} \pi / \mathrm{q}}$ for $\mathrm{k}= \pm \mathrm{q}, \pm(\mathrm{q}+1), \ldots$ coincide with its values for $\mathrm{k}=0, \pm 1, \pm 2, \ldots$
$z^{\alpha}$ takes infinite number of values when $\alpha$ is irrational or complex. Clearly there is a distinct branch of $z^{\alpha}$ for each distinct branch of $\log z$ and the branch cuts are determined as in the case of $\log z$. Every branch of $z^{\alpha}$ is analytic except at the branch point $\mathrm{z}=0$ and on a branch cut.

Example 2. Find all distinct values of $\mathrm{i}^{-2 i}$.
Solution: $\quad i^{-2 i}=e^{-2 i \log i}=e^{2 i\left[\log i \left\lvert\,+i\left(\frac{\pi}{2}+2 k \pi\right)\right.\right]}, k=0, \pm 1, \ldots$

$$
=\mathrm{e}^{(4 \mathrm{k}+1) \pi}, \mathrm{k}=0, \pm 1, \pm 2, \ldots
$$

So, there are infinite number of values.
Example 3. Find all solutions of $z^{1-i}=6$.
Solution : $\mathrm{e}^{(1-\mathrm{i}) \log \mathrm{z}}=\mathrm{e}^{\log 6}$
$\Rightarrow \quad(1-\mathrm{i}) \log \mathrm{z}=\log 6+2 \mathrm{k} \pi \mathrm{i}, \mathrm{k}=0, \pm 1, \pm 2, \ldots$
or, $\quad 2 \log z=(1+i)[\log 6+2 k \pi i]$
or, $\quad \log \mathrm{z}=\frac{\log 6-2 \mathrm{k} \pi}{2}+\frac{\mathrm{i}}{2}(\log 6+2 \mathrm{k} \pi)$
Thus,

$$
\begin{aligned}
z & =e^{\log \sqrt{6}-k \pi}[\cos (k \pi+\log \sqrt{6})+i \sin (k \pi+\log \sqrt{6})] \\
& =\sqrt{6} e^{-k \pi}(-1)^{k}[\cos (\log \sqrt{6})-i \sin (\log \sqrt{6})]
\end{aligned}
$$

### 4.4 Branch, Branch point and Branch cut

Definition : $F(z)$ is a Branch of the multi-valued function $f(z)$ in a domain $D$ if $F(z)$ is single-valued and continuous in $D$ and has the property that for each $z$ in $D$ the value of $F(z)$ is one of the values of $f(z)$.

To determine $\mathrm{F}(\mathrm{z})$ we introduce a line imanating from a point (called a Branch Point) to ensure that F is single-valued in the cut plane by the line. A Branch Point is one for which if we enclose it with a curve the function changes discontinuously as the variable makes a complete round over the curve.

For instance, consider $\mathrm{w}=\mathrm{z}^{1 / 2}$. Let P be a point on the z -plane where $\mathrm{w}_{1}=\mathrm{z}_{1}^{1 / 2}$ and $\operatorname{Arg} \mathrm{z}_{1}=\phi_{1}, 0<\phi_{1}<2 \pi$.

Let $\mathrm{Z}_{1}=\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \phi_{1}}$, then at $\mathrm{P}, \mathrm{w}_{1}=\mathrm{r}_{1}^{1 / 2} \mathrm{e}^{\mathrm{i} \phi_{1} / 2}$. We now encircle the region along closed


Fig. 32

curve $C$ through $P$. Upon travelling anticlockwise once, we have $\phi=\phi_{1}+2 \pi$, i.e., $w=r_{1}^{1 / 2} e^{i\left(\phi_{1}+2 \pi\right) / 2}=-r_{1}^{1 / 2} e^{1 \phi_{1} / 2}$ at the point P.
$\Rightarrow \mathrm{w}=-\mathrm{w}_{1}$ at P . This shows that w has changed discontinuously after performing a loop about $\mathrm{z}=0$, which establishes $\mathrm{z}=0 \mathrm{a}$ Branch Point.

Now we consider a different loop, a closed curve $\Gamma$ around some point $\mathrm{z}^{*}$ which does not enclose the origin. As before, $\mathrm{Z}_{1}=\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \theta_{1}}$ and $\mathrm{w}_{1}=\mathrm{r}_{1}^{1 / 2} \mathrm{e}^{\mathrm{i} \mathrm{i}_{1} / 2}$ upon returning to P , travelling anticlockwise, we have $\phi=$ $\phi_{1}$ again. Hence $w$ is continuous after performing the loop. So $\mathrm{z}=\mathrm{z}^{*}$ is not a Branch Point for $\mathrm{z}^{1 / 2}=\mathrm{w}$.

Example 4. Discuss the multivaluedness of the function $f(z)=\left(z^{2}-1\right)^{1 / 2}$ and introduce cuts to obtain single-valued branches.
Solution : Let $\mathrm{z}-1=\mathrm{r}_{1} \mathrm{e}^{\mathrm{e} \theta}$ and $\mathrm{z}+1=\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \psi}$
Then $\mathrm{f}(\mathrm{z})=\sqrt{\mathrm{r}_{1} \mathrm{r}_{2}} \mathrm{e}^{\mathrm{i}(\theta+\psi) / 2}$
We choose a branch of $f(z)$ at a point $z_{0}$ by taking values of $\theta_{0}$ of $\theta$ and $\psi_{0}$ of $\psi$. Then at $\mathrm{z}_{0}, \mathrm{f}(\mathrm{z})$ takes the value

$$
\mathrm{f}_{0}=\sqrt{\mathrm{r}_{1} \mathrm{r}_{2}} \mathrm{e}^{\mathrm{i}\left(\theta_{0}+\psi_{0}\right) / 2}
$$

If now z traverses from the point $\mathrm{z}_{0}$, and form a simple closed contour (end point also $\mathrm{z}_{0}$ ) $\mathrm{C}_{0}$ enclosing the point $\mathrm{z}=1$, where the point $\mathrm{z}=-1$ lies outside $\mathrm{C}_{0}$, the value of $f(z)$ at $z_{0}$ changes to

$$
\sqrt{\mathrm{r}_{1} \mathrm{r}_{2}} \mathrm{e}^{\mathrm{i}\left(\theta_{0}+\psi_{0}+2 \pi\right) / 2}=-\mathrm{f}_{0}
$$



Fig. 34


Fig. 35
$\mathrm{f}(\mathrm{z})$ takes the same value $-\mathrm{f}_{0}$ while z travelling from $\mathrm{z}_{0}$ and returns to $\mathrm{z}_{0}$ itself forming a closed contour $C_{1}$ which encloses -1 , but not 1 . Hence it is clear that -1 and 1 are the branch points for the function $f(z)$.

In order to obtain single-valued branches we introduce two different set of branch cuts. (i) A branch cut between the points -1 and 1 on the real axis. In this case consider the closed contour $C$ enclosing the branch points -1 and 1 . Here $f(z)$ returns to the value (from its value $f_{0}$ at $z_{0}$ ).

$$
\sqrt{r_{1} r_{2}} e^{i\left(\theta_{0}+2 \pi+\psi_{0}+2 \pi\right) / 2}=\sqrt{r_{1} r_{2}} \mathrm{e}^{\mathrm{i}\left(\theta_{0}+\psi_{0}\right) / 2}=\mathrm{f}_{0}
$$

So, it is a single-valued branch.
(ii) Two branch cuts on the real-axis, $(-\infty,-1)$ and $(1, \infty)$.


Fig. 36


Fig. 37

Here the contour $\Gamma$ does not enclose any of the branch points, so $f(z)$ remains single-valued as z makes a complete round through $\Gamma$ initiating from $\mathrm{z}_{0}$.

Example 5. Construct a branch of $\log \left(\frac{z-1}{z+1}\right)$, which is analytic at the origin and takes the values $5 \pi \mathrm{i}$ there.

Solution : Let $g(z)=\log \left(\frac{z-1}{z+1}\right)$. The points $z= \pm 1$ are the branch points of $g(z)$ and the behaviour of $g(z)$ at these branch points are similar to $f(z)$ as shown in the previous example. We do not repeat these here.

Write both $\mathrm{z}-1$, and $\mathrm{z}+1$ in polar form :

$$
\mathrm{z}-1=\mathrm{re}^{\mathrm{i} \theta}, \quad \mathrm{z}+1=\rho \mathrm{e}^{\mathrm{i} \psi}
$$

Then we can express

$$
g(z)=\log \left(\frac{r e^{i \theta}}{\rho e^{i \psi}}\right)=\log \left[\frac{r}{\rho} e^{i(\theta-\psi)}\right]
$$

$$
=\log \left(\frac{r}{\rho}\right)+i(\theta-\psi)
$$

We consider the complex z-plane with two branch cuts $(-\infty,-1)$, and $(1, \infty)$. Here the principal branch of $g(z)$ is taken as

$$
\log \left(\frac{r}{\rho}\right)+\mathrm{i}(\theta-\psi), \quad 0 \leq \theta<2 \pi ; \quad-\pi \leq \psi<\pi
$$

Now, $g_{0}=g(0)=i \pi$
In the branch $4 \pi \leq \theta<6 \pi ; \pi \leq \psi<3 \pi, \mathrm{~g}(\mathrm{z})$ will take the value $5 \pi \mathrm{i}$ at the origin.
Example 6. Let $\mathrm{z}=\omega^{2}$ and consider $\operatorname{Re} \omega>0$.



Image is $\mathrm{z} \in \mathbb{C} \backslash(-\infty, 0)$
Note : Injective mapping if $\operatorname{Re} \omega>0$ and $\mathrm{z} \in \mathbb{C} \backslash(-\infty, 0)$. We need a Branch cut along negative real-axis in the z -plane.

Hence

$$
\mathrm{w}=\mathrm{z}^{1 / 2}, \quad \mathrm{z}=\mathrm{re}^{\mathrm{i} \phi}, \quad-\pi<\phi \leq \pi
$$

This ensures that Re $\omega>0$. Here the points on the cut go either to P or $\mathrm{Q} . \mathrm{P}$ and Q are arbitrary.

### 4.5 Integrals of Multi-valued functions

Example 7. Evaluate $\int_{0}^{\infty} \frac{x^{\alpha-1}}{1+x} d x, 0<\alpha<1$.
Let us consider the integal

$$
\int_{\mathrm{C}} \frac{\mathrm{Z}^{\alpha-1}}{1-\mathrm{z}} \mathrm{dz}
$$

where the contour C consists of a large Circle $\Gamma_{\mathrm{R}}$ with centre at the origin and radius R , a small circle $\gamma_{\varepsilon}$ with centre origin and radius $\varepsilon$ joined to the large circle
$\Gamma_{\mathrm{R}}$ along the negative side of the real axis from $\varepsilon$ to R by means of a cut as shown in the figure 39. Thus we have avoided the branch point $\mathrm{z}=0$.

We take principal branch of $z^{\alpha-1}$. Then

$$
\left|\int_{\Gamma_{R}} \frac{z^{\alpha-1}}{1-z} d z\right| \leq 2 \pi R \frac{R^{\alpha-1}}{1+R}=\frac{2 \pi R^{\alpha}}{1+R} \rightarrow 0 \text { as } R \rightarrow \infty,
$$

since $\alpha<1$,

$$
\left|\int_{\gamma_{\varepsilon}} \frac{\mathrm{z}^{\alpha-1}}{1-\mathrm{z}} \mathrm{dz}\right| \leq 2 \pi \varepsilon \frac{\varepsilon^{\alpha-1}}{1}=2 \pi \varepsilon^{\alpha} \rightarrow 0 \text { as } \varepsilon \rightarrow 0,
$$

since $\alpha>0$.
Thus, by residue theorem,

$$
\int_{C} \frac{z^{\alpha-1}}{1-z} d z=2 \pi i \operatorname{Res}\left[\frac{z^{\alpha-1}}{1-z} ; 1\right]
$$



Fig. 39

Observe that $\frac{\mathrm{z}^{\alpha-1}}{1-\mathrm{z}}$ has a simple pole at $\mathrm{z}=1$ which lies inside C .
or, $\lim _{R \rightarrow \infty} \int_{\Gamma_{R}} \frac{z^{\alpha-1}}{1-z} d z+\lim _{\varepsilon \rightarrow 0} \int_{\gamma_{\varepsilon}} \frac{z^{\alpha-1}}{1-z} d z+\int_{\gamma_{\alpha}} \frac{z^{\alpha-1}}{1-z} d z+\int_{\gamma_{\beta}} \frac{z^{\alpha-1}}{1-z} d z=-2 \pi i$
so,

$$
\begin{equation*}
\int_{\gamma_{\alpha}} \frac{z^{\alpha-1}}{1-z} d z+\int_{\gamma_{\beta}} \frac{z^{\alpha-1}}{1-z} d z=-2 \pi i \tag{54}
\end{equation*}
$$

On $\quad \gamma_{\alpha}: z=\rho \mathrm{e}^{\mathrm{i} \pi}, 0<\rho<\infty$
so $\quad 1-z=1+\rho$ and $d z=e^{i \pi} d \rho$
and

$$
\int_{\gamma_{\alpha}} \frac{z^{\alpha-1}}{1-z} d z=\int_{\infty}^{0} e^{i \pi} \frac{\rho^{\alpha-1}}{1+\rho} e^{i \pi(\alpha-1)} d \rho=e^{i \pi(\alpha-1)} \int_{0}^{\infty} \frac{\rho^{\alpha-1}}{1+\rho} d \rho=-e^{i \pi \alpha} \int_{0}^{\infty} \frac{\rho^{\alpha-1}}{1+\rho} d \rho
$$

On $\quad \gamma_{\beta}, \mathrm{z}=\rho \mathrm{e}^{-\mathrm{i} \pi}, 0<\rho<\infty$
so

$$
\begin{aligned}
& 1-z=1+\rho, d z=e^{-i \pi} d \rho, \text { then } \\
& \int_{\gamma_{\beta}} \frac{z^{\alpha-1}}{1-z} d z+\int_{0}^{\infty} e^{-i \pi} \frac{\rho^{\alpha-1}}{1+\rho} e^{-i \pi(\alpha-1)} d \rho=-e^{-i \pi(\alpha-1)} \int_{0}^{\infty} \frac{\rho^{\alpha-1}}{1+\rho} d \rho \\
& =e^{-i \pi \alpha} \int_{0}^{\infty} \frac{\rho^{\alpha-1}}{1+\rho} d \rho
\end{aligned}
$$

Substituting these integrals into (54), we get

$$
\left[-\mathrm{e}^{\mathrm{i} \pi \alpha}+\mathrm{e}^{-\mathrm{i} \pi \alpha}\right] \int_{0}^{\infty} \frac{\rho^{\alpha-1}}{1+\rho} \mathrm{d} \rho=-2 \pi \mathrm{i}
$$

i.e. $\quad \int_{0}^{\infty} \frac{\rho^{\alpha-1}}{1+\rho} \mathrm{d} \rho=\frac{2 \pi \mathrm{i}}{2 \mathrm{i} \sin \pi \alpha}$
take branch cut on the negative real-axis
or, $\quad \int_{0}^{\infty} \frac{x^{\alpha-1}}{1+x} d x=\frac{\pi}{\sin \pi \alpha}$
Example 8 : Evaluate $\int_{0}^{\infty} \frac{\mathrm{x}^{\alpha-1}}{1+\mathrm{x}^{3}} \mathrm{dx}, 0<\alpha<3$.
We take the contour integral


Fig. 40

$$
\int_{C} \frac{\mathrm{z}^{\alpha-1}}{1+\mathrm{z}^{3}} \mathrm{dz} \text {, where } \mathrm{C} \text { is the contour as shown in the fig. } 40 \text {. Take }
$$ principal branch of $z^{\alpha-1}$.

Then, $\quad\left|\int_{\gamma \varepsilon} \frac{\mathrm{z}^{\alpha-1}}{1+\mathrm{z}^{3}} \mathrm{dz}\right| \geq \frac{2 \pi}{3} \varepsilon \frac{\varepsilon^{\alpha-1}}{1}=\frac{2 \pi}{3} \varepsilon^{\alpha} \rightarrow 0$ as $\rightarrow \varepsilon \rightarrow 0$ since $\varepsilon>0$
and

$$
\left|\int_{\Gamma_{R}} \frac{\mathrm{z}^{\alpha-1}}{1+\mathrm{z}^{3}} \mathrm{dz}\right| \leq \frac{2 \pi \mathrm{R}}{3} \frac{\mathrm{R}^{\alpha-1}}{\mathrm{R}^{3}}=\frac{2 \pi}{3} \mathrm{R}^{\alpha-3} \rightarrow \infty \text { as } \mathrm{R} \rightarrow \infty \text { since } \alpha<3
$$

Now the function $z^{\alpha-1} / 1+z^{3}$ has only one simple pole $z=e^{\frac{i \pi}{3}}$ inside C. Thus

$$
\int_{c} \frac{z^{\alpha-1}}{1+z^{3}} d z=2 \pi i \operatorname{Res}\left[\frac{z^{\alpha-1}}{1+z^{3}} ; e^{i \pi / 3}\right]=2 \pi i \cdot \frac{e^{\frac{i \pi}{3}(\alpha-1)}}{3 e^{2 \pi i / 3}}=-\frac{2 \pi i}{3} e^{i \alpha \pi / 3}
$$

i.e., $\int_{\Gamma_{R}} \frac{\mathrm{z}^{\alpha-1}}{1+\mathrm{z}^{3}} \mathrm{dz}+\int_{\gamma \varepsilon} \frac{\mathrm{z}^{\alpha-1}}{1+\mathrm{z}^{3}}+\int_{\mathrm{R}}^{\varepsilon} \frac{\rho^{\alpha-1}}{1+\rho^{3}} \mathrm{e}^{2 \pi \mathrm{i}(\alpha-1) / 3} \mathrm{e}^{2 \pi \mathrm{i} / 3} \mathrm{~d} \rho+\int_{\varepsilon}^{\mathrm{R}} \frac{\rho^{\alpha-1}}{1+\rho^{3}} \mathrm{~d} \rho=-2 \pi \mathrm{i} \frac{\mathrm{e}^{\alpha \pi i / 3}}{3}$
[In the third integral, we used $z=\rho e^{2 \pi i / 3}, d z=e^{2 \pi i / 3} d \rho, 1+z^{3}=1+\rho^{3}$, and in the fourth integral, $z=\rho, d z=d \rho$ ]

Taking $\mathrm{R} \rightarrow \infty$ and $\varepsilon \rightarrow 0$ in the above integrals, we find using the earlier results

$$
-\mathrm{e}^{2 \alpha \pi i / 3} \int_{0}^{\alpha} \frac{\rho^{\alpha-1}}{1+\rho^{3}} \mathrm{~d} \rho+\int_{0}^{\alpha} \frac{\rho^{\alpha-1}}{1+\rho^{3}} \mathrm{~d} \rho=\frac{2 \pi \mathrm{ie}^{\alpha \pi \mathrm{i} / 3}}{3}
$$

So that,

$$
\int_{0}^{\infty} \frac{\rho^{\alpha-1}}{1+\rho^{3}} \mathrm{~d} \rho=\frac{2 \pi \mathrm{i}}{3} \cdot \frac{1}{\mathrm{e}^{\alpha \pi i / 3}-\mathrm{e}^{-\alpha \pi i / 3}}=\frac{\pi}{3 \sin \frac{\alpha \pi}{3}}
$$

or, $\quad \int_{0}^{\infty} \frac{\mathrm{x}^{\alpha-1}}{1+\mathrm{x}^{3}} \mathrm{dx}=\frac{\pi}{3 \sin \frac{\alpha \pi}{3}}$

## Riemann Surface

A Riemann surface is a generalization of the complex plane to a surface comprising several sheets so that a multi-valued function can have only one value corresponding to each point on that surface. Once such a surface is ascertained for a given multi-valued function, the function becomes single-valued on the surface and can be treated according to the theory of single-valued functions.

This topology removes artificial restrictions-Branch Cuts and gives us a more general notion of a domain so that a multi-valued analytic function becomes singlevalued if it is considered as a mapping to an appropriate generalized domain as suggested by G. F. B. Riemann (1826-1866) in 1851. The idea is ingenious-a geometric construction that permits surfaces to be the domain or range of a multivalued function.

### 4.6 Branch points at infinity

So far we have considered only branch points in the finite plane. Now we discuss about the possibility of a branch point at infinity. For this sake we map the point at infinity to the origin with the transformation $\varsigma=1 / \mathrm{z}$ and then examine the point $\varsigma=0$.

Example 9 : Again we consider the multi-valued function $f(z)=z^{1 / 2}$. Making the transformation $\varsigma=\frac{1}{\mathrm{z}}$, the point at infinity is mapped to the origin, we have $f(\varsigma)=(1 / \varsigma)^{1 / 2}$. For each value of $\varsigma$, there are two values of $\varsigma^{-1 / 2}$. Writing $\varsigma^{-1 / 2}$ in modulus-argument form

$$
\varsigma^{-1 / 2}=\frac{1}{\sqrt{|\zeta|}} \mathrm{e}^{-\mathrm{i} \operatorname{Arg}(\varsigma) / 2}
$$

we find that like $\mathrm{z}^{1 / 2}, \varsigma^{-1 / 2}$ possesses double sheeted Riemann surface. We see that each time we walk around the origin, the argument of $\varsigma^{-1 / 2}$ changes by $-\pi$. This means that the value of the function changes by the factor $\mathrm{e}^{-\mathrm{i} \pi}=-1$, i.e. the function changes sign. If we walk around the origin twice, the argument changes by $-2 \pi$, so that the value of the function does not change, $\mathrm{e}^{-2 \pi \mathrm{i}}=1$.

Now, since $\varsigma^{-1 / 2}$ has a branch point at zero, we conclude that $z^{1 / 2}$ has a branch point at infinity.

Example 10 : Again consider the multi-valued $\operatorname{logarithm}$ function $\mathrm{f}(\mathrm{z})=\log \mathrm{z}$. Mapping the point at infinity to the origin, we have

$$
f(\varsigma)=\log (1 / \varsigma)=-\log \varsigma
$$

But $\log \varsigma$ has a branch point at $\varsigma=0$. Thus $\log \mathrm{z}$ has a branch point at infinity.

## Branch points at infinity : Paths around infinity

We can also check for a branch point at infinity by considering a path that encloses the point at infinity and no other singularities. This can be done by drawing a simple closed curve that separates the complex plane into a bounded region that contains all the singularities of the function in the finite plane. Then, depending upon the orientation, the curve is a contour enclosing all the finite singularities, or the point at infinity and no other singularities.

Once again consider the function $\mathrm{z}^{1 / 2}$. We know that the function changes value on a curve that goes around the origin. Such a curve can be considered to be either a path around the origin or a path around the point at infinity. In either case the path encloses one branch point. Now consider a curve that does not go around the origin. Such a curve can be considered to be either a path around neither of the branch points or both of them. Thus we see that $z^{1 / 2}$ does not change value when we follow a path that encloses neither or both of its branch points.

Example 11 : Consider the multi-valued function $f(z)=\left(z^{2}-1\right)^{1 / 2}$. Rewriting the function $f(z)=(z-1)^{1 / 2}(z+1)^{1 / 2}$, we see that there are branch points at $z= \pm 1$. Now consider the point at infinity.

$$
f\left(\varsigma^{-1}\right)=\left(\varsigma^{-2}-1\right)^{1 / 2}= \pm \varsigma^{-1}\left(1-\zeta^{2}\right)^{1 / 2}
$$

which shows that $\mathrm{f}\left(\varsigma^{-1}\right)$ does not have a branch point at $\varsigma=0$ and $\mathrm{f}(\mathrm{z})$ does not have a branch point at infinity. We might reach the same conclusion by considering a path around the point at infinity. Consider a path that encircles the branch points at $\mathrm{z}= \pm 1$ once in the positive direction. Equivalently it encircles the point at infinity once in the negative direction. In traversing this path, the value of $f(z)$ is multiplied by the factor $\left(\mathrm{e}^{2 i \pi}\right)^{1 / 2}\left(\mathrm{e}^{2 i \pi}\right)^{1 / 2}=\mathrm{e}^{2 i \pi}=1$. Thus the value of the function remains unchanged. There is no branch point at infinity.

### 4.7 Detection of branch points

We have the definition of a branch point, but we do not have a convenient criterion for determining if a particular function has a branch point. We have noticed that $\log \mathrm{z}$ and $\mathrm{z}^{\mathrm{k}}$ for non-integer k have branch points at zero and infinity. The inverse trigonometric functions like $\sin ^{-1} \mathrm{z}, \cos ^{-1} \mathrm{z}$ etc. also have branch points, but they can be written in terms of the logarithm and the square root. In fact all the elementary functions with branch points can be written in terms of the functions $\log \mathrm{z}$ and $\mathrm{z}^{\mathrm{k}}$. Furthermore, note that the multi-valuedness of $\mathrm{z}^{\mathrm{k}}$ comes from the logarithm, $\mathrm{z}^{\mathrm{k}}=\mathrm{e}^{\mathrm{klog} z}$. This gives us a way of determining branch points of a function if there is any.

Result : Let $\mathrm{f}(\mathrm{z})$ be a single-valued function. Then $\log \mathrm{f}(\mathrm{z})$ and $(\mathrm{f}(\mathrm{z}))^{\mathrm{k}}$ may have branch points only where $f(z)$ is zero or singular.

Example 12 : Consider the functions

1. $\left(z^{2}\right)^{1 / 2}$
2. $\left(z^{1 / 2}\right)^{2}$
3. $\left(z^{1 / 2}\right)^{3}$

Are they multi-valued? Do they have branch points?

## Solution

1. 

$$
\left(\mathrm{z}^{2}\right)^{1 / 2}= \pm \sqrt{\mathrm{z}^{2}}= \pm \mathrm{z}
$$

Because of $(\cdot)^{1 / 2}$, the function is multi-valued. The only possible branch points are at zero and point at infinity. If $\left.\left(\mathrm{e}^{\mathrm{i} \theta}\right)^{2}\right)^{1 / 2}=1$, then as $\left(\left(\mathrm{e}^{2 \pi \mathrm{i}}\right)^{2}\right)^{1 / 2}=\left(\mathrm{e}^{4 \pi \mathrm{i}}\right)^{1 / 2}=\mathrm{e}^{2 \pi \mathrm{i}}=1$ the function does not change value when we walk around the origin. We can also consider this to be a path around infinity. This function is multi-valued, but has no branch points.

$$
\text { 2. } \quad\left(\mathrm{z}^{1 / 2}\right)^{2}=( \pm \sqrt{\mathrm{z}})^{2}=\mathrm{z}
$$

This function is single-valued.
3.

$$
\left(\mathrm{z}^{1 / 2}\right)^{3}=( \pm \sqrt{\mathrm{z}})^{3}= \pm(\sqrt{\mathrm{z}})^{3}
$$

This function is multi-valued. We consider the possible branch point at $\mathrm{z}=0$. If $\left.\left(\mathrm{e}^{\mathrm{i} 0}\right)^{1 / 2}\right)^{3}=1$, then as $\left(\left(\mathrm{e}^{2 \mathrm{i} \pi}\right)^{1 / 2}\right)^{3}=\left(\left(\mathrm{e}^{\mathrm{i} \pi 2}\right)^{1 / 2}\right)^{3}=\left(\mathrm{e}^{\mathrm{i} \pi}\right)^{3}=\mathrm{e}^{3 \pi \mathrm{i}}=-1$, the function changes value when we walk around the origin. So it has a branch point at $z=0$. Since this is also a path around infinity, there is a branch point at the point at infinity.

Example 13 : Consider the function $f(z)=\log (1 / z-1)$. Since $\frac{1}{z-1}$ has only zero at infinity and its only singularity (a pole here) is at $\mathrm{z}=1$, the only, possible branch points are at $\mathrm{z}=1$ and $\mathrm{z}=\infty$.

Here $f(z)=\log \left(\frac{1}{z-1}\right)=-\log (z-1)=\log \omega$, say
We know that $\log \omega$ has branch points at zero and infinity, so $f(z)$ has branch points at $\mathrm{z}=1$ and $\mathrm{z}=\infty$.

Example 14 : Consider the functions

1. $\mathrm{e}^{\log z} 2 . \log \mathrm{e}^{\mathrm{z}}$

Are they multi-valued? Do they have branch points?

## Solution :

1. 

$$
\begin{aligned}
& \mathrm{e}^{\log z}=\mathrm{e}^{\log z+\mathrm{i} 2 \pi \mathrm{k}}, \mathrm{k}=0, \pm 1, \ldots \\
& =\mathrm{e}^{\log z} \mathrm{e}^{\mathrm{i} 2 \pi \mathrm{k}}=\mathrm{z}
\end{aligned}
$$

The function is single-valued.
2.

$$
\operatorname{loge}^{\mathrm{z}}=\operatorname{Loge}^{\mathrm{z}}+\mathrm{i} 2 \pi \mathrm{k}=\mathrm{z}+\mathrm{i} 2 \pi \mathrm{k}, \mathrm{k}=0, \pm 1, \ldots
$$

This function is multi-valued. It may have branch points only where $e^{z}$ is zero or infinite. This occurs only at $\mathrm{z}=\infty$. Thus there are no branch points in the finite plane. The function does not change when traversing a simple closed path and since this path can be considered to enclose the point at infinity, there is no branch point at infinity.

Note : Let $f(z)$ be single-valued and have either a zero or a singularity at $z=z_{0}$. Then $\{f(z)\}^{k}$ may have a branch point at $z=z_{0}$. If $f(z)$ is not a power of $z$, then we are not sure whether $\{\mathrm{f}(\mathrm{z})\}^{\mathrm{k}}$ changes value when we walk around $\mathrm{z}_{0}$.

Now if $f(z)$ can be decomposed into factors $f(z)=h(z) g(z)$, where $h(z)$ is finite and non zero at $z_{0}$, then from $g(z)$ we know how fast $f(z)$ vanishes or tends to infinity. Again $\{\mathrm{f}(\mathrm{z})\}^{\mathrm{k}}=\{\mathrm{h}(\mathrm{z})\}^{\mathrm{k}}\{\mathrm{g}(\mathrm{z})\}^{\mathrm{k}}$ and $\{\mathrm{h}(\mathrm{z})\}^{\mathrm{k}}$ does not have a branch point at $\mathrm{z}_{0}$. So that $\{\mathrm{f}(\mathrm{z})\}^{\mathrm{k}}$ has a branch point at $\mathrm{z}_{0}$ if and only if $\{\mathrm{f}(\mathrm{z})\}^{\mathrm{k}}$ has a branch point there.

Similarly, we can decompose

$$
\log \{\mathrm{f}(\mathrm{z})\}=\log \{\mathrm{h}(\mathrm{z}) \mathrm{g}(\mathrm{z})\}=\log \{\mathrm{h}(\mathrm{z})\}+\log \{\mathrm{g}(\mathrm{z})\}
$$

to see that $\log \{\mathrm{f}(\mathrm{z})\}$ has a branch point at $\mathrm{z}_{0}$ if and only if $\log \{\mathrm{g}(\mathrm{z})\}$ has a branch point there.

Example 15 : Consider the functions :

1. $\sin z^{1 / 2}$
2. $(\sin \mathrm{z})^{1 / 2}$
3. $z^{1 / 2} \cos z^{1 / 2}$
4. $\left(\sin z^{2}\right)^{1 / 2}$.

Find the branch points and the number of branches.
Solution : 1. $\sin \mathrm{z}^{1 / 2}=\sin ( \pm \sqrt{z})= \pm \sin \sqrt{z}$
So it is multi-valued. It has two branches and the possible branch points are zero and infinity. Consider the unit circle $|\mathrm{z}|=1$ which is a path around the origin and infinity. If

$$
\sin \left(\mathrm{e}^{\mathrm{i} 0}\right)^{1 / 2}=\sin (1) \text {, then as }
$$

$$
\sin \left(\left(\mathrm{e}^{\mathrm{i} 2 \pi}\right)^{1 / 2}\right)=\sin \left(\mathrm{e}^{\mathrm{i} \pi}\right)=\sin (-1)=-\sin 1,
$$

there are branch points at the origin and infinity
2. $\quad(\sin \mathrm{z})^{1 / 2}= \pm \sqrt{\sin \mathrm{Z}}$

The function is multi-valued and has two branches. The sine function vanishes at $\mathrm{z}=\mathrm{n} \pi$ and is singular at infinity. These may be branch points of the function. Consider the point $z=n \pi$. We can express

$$
\sin \mathrm{z}=(\mathrm{z}-\mathrm{n} \pi) \frac{\sin \mathrm{z}}{\mathrm{z}-\mathrm{n} \pi}, \mathrm{n} \text { an integer. }
$$

But $\quad \lim _{z \rightarrow n \pi} \frac{\sin z}{z-n \pi}=\lim _{z \rightarrow n \pi} \frac{\cos z}{1}=(-1)^{n}$
So, $(\sin z)^{1 / 2}$ has branch points at $z=n \pi$ since $(z-n \pi)^{1 / 2}$ has a branch point at $\mathrm{z}=\mathrm{n} \pi$.

Here the branch points are $\mathrm{z}=\mathrm{n} \pi, \mathrm{n}=0, \pm 1, \ldots$ and they go to infinity. So it is not possible to make a path that encloses infinity and no other singularities. The point at infinity is a non-isolated singularity. A point can be a branch point only if it is an isolated singularity.
3. $\mathrm{z}^{1 / 2} \cdot \cos \mathrm{z}^{1 / 2}= \pm \sqrt{\mathrm{z}} \cos ( \pm \sqrt{\mathrm{z}})$

$$
= \pm \sqrt{\mathrm{z}} \cos \sqrt{\mathrm{z}}
$$

The function is multi-valued. It may possess branch points at $\mathrm{z}=0$ and $\mathrm{z}=\infty$. If $\left(\mathrm{e}^{\mathrm{i} 0}\right)^{1 / 2} \cos \left(\mathrm{e}^{\mathrm{i} 0}\right)^{1 / 2}=\cos (1)$, then as $\left(\mathrm{e}^{\mathrm{i} 2 \pi}\right)^{1 / 2} \cos \left(\left(\mathrm{e}^{\mathrm{i} 2 \pi}\right)^{1 / 2}\right)=(-1) \cos \left(\mathrm{e}^{\mathrm{i} \pi}\right)=-\cos (-1)$ $=-\cos 1$, there are branch points at the origin and infinity.
4. $\quad\left(\sin z^{2}\right)^{1 / 2}= \pm \sqrt{\sin z^{2}}$

The function is multi-valued. Now since siz $z^{2}=0$ at $z=(n \pi)^{1 / 2}$, there may be branch points there.

We consider first the point $\mathrm{z}=0$. We can write

$$
\begin{gathered}
\sin z^{2}=z^{2} \frac{\sin z^{2}}{z^{2}} \\
\text { but } \quad \lim _{z \rightarrow 0} \frac{\sin z^{2}}{z^{2}}=\lim _{z \rightarrow 0} \frac{2 z \cos z^{2}}{2 z}=1
\end{gathered}
$$

So, $\left(\sin z^{2}\right)^{1 / 2}$ does not have a branch point at $z=0$ as $\left(z^{2}\right)^{1 / 2}$ does not have a branch point there.

Next consider the point

$$
\mathrm{z}=\sqrt{\mathrm{n} \pi}
$$

$$
\sin z^{2}=(z-\sqrt{n \pi}) \frac{\sin z^{2}}{z-\sqrt{n \pi}}
$$

but $\quad \lim _{z \rightarrow \sqrt{n \pi}} \frac{\sin z^{2}}{z-\sqrt{n \pi}}=\lim _{z \rightarrow \sqrt{n \pi}} \frac{2 z \cos z^{2}}{1}=2 \sqrt{n \pi}(-1)^{n}$
Since $(z-\sqrt{n \pi})^{1 / 2}$ has a branch point at $z=\sqrt{n \pi},\left(\sin z^{2}\right)^{1 / 2}$, too as a branch point there.

Thus we see that $\left(\sin z^{2}\right)^{1 / 2}$ has branch points at $z=(n \pi)^{1 / 2}$ for $n \varepsilon Z \backslash\{0\}$. This is the set of numbers : $\{ \pm \sqrt{\pi}, \pm \sqrt{2 \pi}, \ldots, \pm \mathrm{i} \sqrt{\pi}, \pm \mathrm{i} \sqrt{2 \pi}, \ldots\}$. The point at infinity is a non-isolated singularity and hence it is not included in the set of branch points.

Example 16 : Find the branch points of

$$
f(z)=\left(z^{3}-z\right)^{1 / 3}
$$

and introduce the branch cuts. If $f(3)=2 \sqrt[3]{3}$, find $f(-3)$.
Solution : Here $f(z)=z^{1 / 3}(z-1)^{1 / 3}(z+1)^{1 / 3}$
So the branch points are at $\mathrm{z}=-1,0$ and 1 . We consider the point at infinity

$$
\begin{aligned}
f(1 / \varsigma) & =\left(\frac{1}{\zeta}\right)^{1 / 3}\left(\frac{1}{\zeta}-1\right)^{1 / 3}\left(\frac{1}{\zeta}+1\right)^{1 / 3} \\
& =\frac{1}{\varsigma}(1-\varsigma)^{1 / 3}(1+\varsigma)^{1 / 3}
\end{aligned}
$$

Since $f(1 / \varsigma)$ does not have a branch point at $\varsigma=0, f(z)$ does not have a branch point at infinity.

Here we give three possible branch cuts :




Fig. 41 Three possible branch cuts for $f(z)=\left(z^{3}-z\right)^{1 / 3}$
In the first and third the function is single-valued but in the second it is not. It is clear that the first branch cut does not allow us to walk around any of the branch points.

The second branch cut allows us to walk around the branch points at $\mathrm{z}= \pm 1$. If we walk around these two once in the positive direction, the value of the function would change by the factor $\mathrm{e}^{\mathrm{i} 4 \pi / 3}$.

The third branch cut allows us to walk around all the three branch points, the value of the function will not change (since $\mathrm{e}^{\mathrm{i} 6 \pi / 3}=\mathrm{e}^{\mathrm{i} 2 \pi}=1$ ).

To find $f(-3)$, we consider the third branch cut with $f(3)=2 \sqrt[3]{3}$.

$$
f(3)=\left(3 e^{i 0}\right)^{1 / 3}\left(2 e^{i 0}\right)^{1 / 3}\left(4 e^{i 0}\right)^{1 / 3}=2 \sqrt[3]{3}
$$

The value of $f(-3)$ is

$$
\mathrm{f}(-3)=\left(3 \mathrm{e}^{\mathrm{i} \pi}\right)^{1 / 3}\left(2 \mathrm{e}^{\mathrm{i} \pi}\right)^{1 / 3}\left(4 \mathrm{e}^{\mathrm{i} \pi}\right)^{1 / 3}=-2 \sqrt[3]{3}
$$

Example 17 : Determine the branch points of the function $f(z)=\left(z^{3}-1\right)^{1 / 2}$.
Construct branch cuts and define a branch so that $\mathrm{z}=0$ and $\mathrm{z}=-1$ do not lie on a cut, such that $\mathrm{f}(0)=-\mathrm{i}$; then what is $\mathrm{f}(-1 / 2)$ ?

Solution : The roots of the equation $z^{3}-1=0$ are

$$
\left\{1, \mathrm{e}^{\mathrm{i} 2 \pi / 3}, \mathrm{e}^{\mathrm{i} 4 \pi / 3}\right\}=\left\{1, \frac{-1+\mathrm{i} \sqrt{3}}{2}, \frac{-1-\mathrm{i} \sqrt{3}}{2}\right\}
$$

so that,

$$
\left(\mathrm{z}^{3}-1\right)^{1 / 2}=(\mathrm{z}-1)^{1 / 2}\left(\mathrm{z}+\frac{1-\mathrm{i} \sqrt{3}}{2}\right)^{1 / 2}\left(\mathrm{z}+\frac{1+\mathrm{i} \sqrt{3}}{2}\right)^{1 / 2}
$$

There are branch points at each of the cube roots of unity

$$
\mathrm{z}=\left\{1, \frac{-1+\mathrm{i} \sqrt{3}}{2}, \frac{-1-\mathrm{i} \sqrt{3}}{2}\right\}
$$

Now we examine the point at infinity. We make the change of variable $z=1 / \varsigma$

$$
f(1 / \varsigma)=\left(1 / \varsigma^{3}-1\right)^{1 / 2}=\zeta^{-3 / 2}\left(1-\zeta^{3}\right)^{1 / 2}
$$

$\varsigma^{-3 / 2}$ has a branch point at $\varsigma=0$, while $\left(1-\varsigma^{3}\right)^{1 / 2}$ is not singular there. Since $\mathrm{f}(1 / \varsigma)$ has a branch point at $\varsigma=0, \mathrm{f}(\mathrm{z})$ has a branch point at infinity.

There are several ways of introducing branch cuts to separate the branches of the function. The easiest approach is to put a branch cut from each of the three branch points in the finite complex plane out to the branch point at infinity (see Figure 42a). Clearly this makes the function single-valued as it is impossible to walk around any of the branch points. Another approach is to have a branch cut from one of the branch points in the finite plane to the branch point at infinity and a branch cut connecting the remaining two branch points (see Figure 42 bcd). In this case, in walking around
any one of the finite branch points (in the + ve direction), the argument of the function changes by $\pi$. This means that the value of the function changes by $\mathrm{e}^{\mathrm{i} \pi}$, which is to say, the value of the function changes sign. In walking around any two of the finite branch points (in the + ve direction), the argument of the function changes by $2 \pi$ i.e., the value of the function changes by $\mathrm{e}^{\mathrm{i} 2 \pi}$, that means the value of the function does not change.

Figure 42. Branch cuts for $\left(z^{3}-1\right)^{1 / 2}$





Now we choose the branch 42a, and introduce the variables $\mathrm{z}-1=\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \theta}, 0 \leq \theta_{1}<2 \pi$

$$
\begin{aligned}
& z+\frac{1-i \sqrt{3}}{2}=r_{2} \mathrm{e}^{\mathrm{i} \theta_{2}},-\frac{2 \pi}{3} \leq \theta_{2}<\frac{\pi}{3} \\
& z+\frac{1-\mathrm{i} \sqrt{3}}{2}=\mathrm{r}_{3} \mathrm{e}^{\mathrm{i} \theta_{3}},-\frac{\pi}{3} \leq \theta_{3}<\frac{2 \pi}{3}
\end{aligned}
$$

We compute $f(0)$ to see whether it has the desired value,

$$
\begin{aligned}
& \mathrm{f}(\mathrm{z})=\sqrt{\mathrm{r}_{1} \mathrm{r}_{2} \mathrm{r}_{3}} \mathrm{e}^{\mathrm{i}\left(\theta_{1}+\theta_{2}+\theta_{3}\right) / 2} \\
& \mathrm{f}(0)=\mathrm{e}^{\mathrm{i}(\pi-\pi / 3+\pi / 3) / 2}=\mathrm{e}^{\mathrm{i} \pi / 2}=\mathrm{i}
\end{aligned}
$$

Since it does not have the desired value, we change the range of $\theta_{1}$,

$$
\mathrm{z}-1=\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \theta_{1}}, 2 \pi \leq \theta_{1}<4 \pi
$$

$f(0)$ now has the desired value,

$$
f(0)=e^{i(3 \pi-\pi / 3+\pi / 3)}=-i
$$

We compute $\mathrm{f}\left(-\frac{1}{2}\right)$,

$$
f\left(-\frac{1}{2}\right)=\sqrt{\frac{3}{2} \cdot \frac{\sqrt{3}}{2} \cdot \frac{\sqrt{3}}{2}} e^{i\left(3 \pi-\frac{\pi}{2}+\frac{\pi}{2}\right) / 2}
$$

$$
=\sqrt{\frac{9}{8}} \mathrm{e}^{\mathrm{i} 3 \pi / 2}=\frac{-3 \mathrm{i}}{2 \sqrt{2}}
$$

Example 18 : Identify the branch points of the function

$$
\omega=f(z)=\left(z^{3}+z^{2}-6 z\right)^{1 / 2}
$$

in the extended complex plane. Specify the branch cuts and select a branch that gives a single-valued function where it is continuous at $\mathrm{z}=-1$ with $\mathrm{f}(-1)=-\sqrt{6}$.

Solution : First we factor the function

$$
f(z)=\left[z(z-2(z+3)]^{1 / 2}=z^{1 / 2}(z-2)^{1 / 2}(z+3)^{1 / 2}\right.
$$

There are branch points at $z=-3,0,2$. Now we examine the point at infinity.

$$
f(1 / \varsigma)=\left[\frac{1}{\varsigma}\left(\frac{1}{\zeta}-2\right)\left(\frac{1}{\varsigma}+3\right)\right]^{1 / 2}=\varsigma^{-3 / 2}[(1-2 \varsigma)(1+3 \varsigma)]^{1 / 2}
$$

Since $\varsigma^{-3 / 2}$ has a branch point at $\varsigma=0$ and the rest of the terms are analytic there, $f(\mathrm{z})$ has a branch point at infinity.

Now consider the branch cuts given in the figure 43. These cuts do not permit us to walk around any single branch point. We can walk around none of the branch points (or all of the branch points considering the cuts $[-3,2]$ and $x=0, y \leq 0$ ). The cuts can be


Fig. 43 used to define a single-valued branch of the function.
Now to define the branch, we choose $\mathrm{z}+3=\mathrm{r}_{\mathrm{i}} \mathrm{e}^{\mathrm{i} \theta_{1}},-\pi \leq \theta_{1}<\pi$; $\mathrm{z}=\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \theta_{2}}$, $\frac{-\pi}{2} \leq \theta_{2}<\frac{3 \pi}{2}$ and $z-2=r_{3} \mathrm{e}^{\mathrm{i} \theta 3}, 0 \leq \theta_{3}<2 \pi$.

The function is, $\mathrm{f}(\mathrm{z})=\left(\mathrm{r}_{1} \mathrm{r}_{2} \mathrm{r}_{3}\right)^{1 / 2} \mathrm{e}^{\mathrm{i}\left(\theta_{1}+\theta_{2}+\theta_{3}\right) / 2}$
Here $\mathrm{f}(-1)=[(2)(1)(3)]^{1 / 2} \mathrm{e}^{\mathrm{i}(0+\pi+\pi) / 2}=-\sqrt{6}$
So our choice of angles gave the desired branch.

### 4.8 The Riemann surface for $\omega=z^{1 / 2}$

We have seen that $\omega=z^{1 / 2}$ possesses two branch points $z=0$ and $z=\infty$. To utilize the developments made in Example 1, we introduce a branch cut along the negative real axis. The given function has two values for any $\mathrm{z} \neq 0$.

$$
\mathrm{f}_{1}(\mathrm{z})=\mathrm{r}^{1 / 2 \mathrm{e}^{\mathrm{i} \theta / 2},-\pi<\theta \leq \pi}
$$

and

$$
\mathrm{f}_{2}(\mathrm{z})=\mathrm{r}^{1 / 2} \mathrm{e}^{\mathrm{i} \theta / 2}, \pi<\theta \leq 3 \pi
$$

Each function $f_{1}$ and $f_{2}$ is single-valued on the domain formed by cutting the $z$ plane along the negative real-axis. Let $D_{1}$ and $D_{2}$ be the domains of $f_{1}$ and $f_{2}$


Fig. 44 respectively. The range set for $f_{1}$ is the set $R_{1}$ consisting of the right-half plane and the positive imaginary axis [see Figure 28b]; the range set for $f_{2}$ is the set $R_{2}$ consisting of the left-half plane and the negative imaginary axis [see Figure 29b]. The sets $\mathrm{R}_{1}$ and $\mathrm{R}_{2}$ are glued together along the positive imaginary axis and the negative imaginary axis to form the w-plane with the origin deleted. We stack $\mathrm{D}_{1}$ directly above $D_{2}$. The edge of $D_{1}$ in the upper-half plane is joined to the edge of $D_{2}$ in the lower-half plane, and the edge of $\mathrm{D}_{1}$ in the lower-half plane is joined to the edge of $\mathrm{D}_{2}$ in the upper-half plane (it is needless to mention that the line of joining is the negative real-axis). When these domains are glued together in this manner, they form a Riemann surface domain for the mapping $\mathrm{w}=\mathrm{f}(\mathrm{z})=\mathrm{z}^{1 / 2}$ shown in the figure 44 for some finite r .

### 4.9 Concept of neighbourhood

When a point lies on the boundary of two domains $D_{1}$ and $D_{2}$, we define a neighbourhood of that point in the following manner. Here the boundary of $D_{1}$ and $D_{2}$ is the negative real-axis. (i) Neighbourhood of a point $\varsigma \in D_{1}$ with $\operatorname{Im} \varsigma<0$, $\operatorname{Arg}$ $\varsigma=\pi,|z-\varsigma|<\varepsilon$ consists of points on : (a) $D_{1}$ if $\operatorname{Im} \varsigma \geq 0$ (b) $D_{2}$ if $\operatorname{Im} \varsigma<0$. (ii) Neighbourhood of a point $\eta \varepsilon D_{2}$ with $\operatorname{Im} \eta=0$, $\operatorname{Arg} \eta=3 \pi,|z-\eta|<\varepsilon$ consists of points on (a) $D_{1}$ if $\operatorname{Im} \eta<0$ and (b) $D_{2}$ if $\operatorname{Im} \eta \geq 0$. With these definitions of neighbourhood of a point, it becomes clear that $w=z^{1 / 2}$ is continuous and differentiable everywhere on the Riemann surface except at the origin and the point at infinity. The derivative is given by

$$
\frac{\mathrm{d}}{\mathrm{dz}} \mathrm{z}^{1 / 2}= \begin{cases}\frac{1}{2} & \frac{1}{\mathrm{f}_{1}} \text { on } \mathrm{D}_{1} \\ \frac{1}{2} & \frac{1}{\mathrm{f}_{2}} \text { on } \mathrm{D}_{2}\end{cases}
$$

### 4.10 The Riemann Surface for $w=\log z$

The Riemann surface for the multivalued function $\omega=\log \mathrm{z}$ is similar to the one we presented for the square root function. However, it requires infinitely many copies of the z -plane cut along the negative x -axis, which mark $\mathrm{S}_{\mathrm{k}}$ for $\mathrm{k}=\ldots,-\mathrm{n}, \ldots,-1,0$, $1, \ldots, \mathrm{n}, \ldots$. Now we stack these cut planes directly on each other so that the corresponding points have the same position. We join the sheet $S_{k}$ to $S_{k+1}$ as follows. For each integer $k$, the edge of the sheet $S_{k}$ in the upper half-plane is joined to the edge of the sheet $\mathrm{S}_{\mathrm{k}+1}$ in the lower-half plane. The Riemann surface for the domain of $\log \mathrm{z}$ looks like a spiral staircase that extends upward on the sheets $\mathrm{S}_{1}, \mathrm{~S}_{2} \ldots$, and downward on the sheets $S_{-1}, S_{-2}, \ldots$ as shown in figure 45 . For $S_{k}$, we use

$$
\begin{aligned}
& z=r e^{i \theta}=r(\cos \theta+i \sin \theta), \text { where } \\
& r=|z| \text { and } 2 \pi k-\pi<\theta \leq \pi+2 \pi k
\end{aligned}
$$

Again, for $\mathrm{S}_{\mathrm{k}}$, the correct branch of $\log \mathrm{z}$ on each sheet is

$$
\begin{aligned}
& \log \mathrm{z}=\log \mathrm{r}+\mathrm{i} \theta, \text { where } \\
& \mathrm{r}=|\mathrm{z}| \text { and } 2 \pi \mathrm{k}-\pi<\theta \leq \pi+2 \pi k
\end{aligned}
$$



Fig. 45


Fig. 46

Example 19 : Form a Riemann surface for $f(z)=(z-1)^{1 / 3}$ taking a branch cut along the line $y=0, x \geq 1$. Detect the point where the function takes the value $\sqrt{2}(\mathrm{i}-1)$.

Solution : Let $r=|z-1|$ and $\theta=\arg (z-1)$, where $0 \leq \theta<2 \pi$. Here the Riemann surface consists of three domains $\mathrm{D}_{1} \mathrm{D}_{2}$ and $\mathrm{D}_{3}$ :

$$
\begin{array}{ll}
\mathrm{f}_{1}(\mathrm{z})=\mathrm{r}^{1 / 3} \mathrm{e}^{\mathrm{i} \theta / 3}, & 0 \leq \theta<2 \pi\left(\mathrm{D}_{1}\right) \\
\mathrm{f}_{2}(\mathrm{z})=\mathrm{r}^{1 / 3} \mathrm{e}^{\mathrm{i} \theta / 3}, & 2 \pi \leq \theta<4 \pi\left(\mathrm{D}_{2}\right)
\end{array}
$$

$$
\mathrm{f}_{3}(\mathrm{z})=\mathrm{r}^{1 / 3} \mathrm{e}^{\mathrm{i} \theta / 3}, \quad 4 \pi \leq \theta<6 \pi\left(\mathrm{D}_{3}\right)
$$

Each function $f_{1}, f_{2}$ and $f_{3}$ is single-valued on the domain formed by cutting the z-plane along the line $y=0, x \geq 1$.

We place $D_{1}$ on the top, then $D_{2}$ and $D_{3}$. The edge of $D_{1}$ in the upper-half plane is joined to the edge of $D_{2}$ in the lower-half plane and the edge of $\mathrm{D}_{2}$ in the upperhalf plane is joined to the edge of $D_{3}$ in the lower-half plane and finally the edge of $D_{3}$ in the upper-half plane is joined to the edge of $D_{1}$ in the lower-half plane.

Say at $\mathrm{z}=\varsigma, \mathrm{f}(\varsigma)=\sqrt{ } 2(\mathrm{i}-1)$
i.e.

$$
\begin{aligned}
& \mathrm{f}(\varsigma)=-2\left(\frac{1}{\sqrt{2}}-\frac{\mathrm{i}}{\sqrt{2}}\right) \\
& =2 \mathrm{e}^{\mathrm{i} \pi} \mathrm{e}^{-\frac{\mathrm{i} \pi}{4}}=2 \mathrm{e}^{\mathrm{i} 3 \pi / 4} \\
& =2 \mathrm{e}^{\mathrm{i}\left(\frac{9 \pi}{4}\right) / 3}=2 \mathrm{e}^{\mathrm{i}\left(\frac{\pi}{4}+2 \pi\right) / 3}
\end{aligned}
$$



Fig. 47

So, $\varsigma-1=2^{3} \mathrm{e}^{\frac{\mathrm{i} \pi}{4}}, \varsigma=1+8 \mathrm{e}^{\frac{\mathrm{i} \pi}{4}}$ lying in the domain $\mathrm{D}_{2}$.
Example 20 : Form the Riemann surface for the function $f(z)=\left(z^{2}-1\right)^{1 / 2}$.
Solution : Here the given function $f(z)=\left(z^{2}-1\right)^{1 / 2}$ has branch points at $\mathrm{z}= \pm 1$. To examine the point at infinity, we substitute $\mathrm{z}=1 / \varsigma$ and examine the point $\varsigma=0$.

$$
\mathrm{f}(1 / \varsigma)=\left[\left(\frac{1}{\zeta}\right)^{2}-1\right]^{\mathrm{i} / 2}=\frac{1}{\left(\varsigma^{2}\right)^{1 / 2}}\left(1-\varsigma^{2}\right)^{1 / 2}
$$

Since there is no branch point at $\varsigma=0, \mathrm{f}(\mathrm{z})$ has no branch point at infinity.
Let $\mathrm{z}-1=\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \phi_{1}}$ and $\mathrm{z}+1=\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \phi_{2}}$,
where $\phi_{1}=\operatorname{Arg}(z-1)$ and $\phi_{2}=\operatorname{Arg}(z+1)$. Then $\omega=f(z)=\left(z^{2}-1\right)^{1 / 2}$

$$
=(\mathrm{z}-1)^{1 / 2}(\mathrm{z}+1)^{1 / 2}=\left(\mathrm{r}_{1} \mathrm{r}_{2}\right)^{1 / 2} \mathrm{e}^{\mathrm{i}\left(\phi_{1}+\phi_{2}\right)}
$$



Case-I $0 \leq \phi_{1}<2 \pi, 0 \leq \phi_{2}<2 \pi$

| on the <br> segment | $\phi_{1}$ | $\phi_{2}$ | $\mathrm{e}^{\mathrm{i}\left(\phi_{1}+\phi_{2}\right) / 2}$ | Continuity <br> of $\mathrm{f}(\mathrm{z})$ |
| :--- | :---: | :---: | :---: | :---: |
| B | $\pi$ | 0 | i | No |
| $\mathrm{B}^{\prime}$ | $\pi$ | $2 \pi$ | -i |  |
| C | 0 | 0 | 1 | Yes |
| $\mathrm{C}^{\prime}$ | $2 \pi$ | $2 \pi$ | 1 |  |
| D | $\pi$ | $\pi$ | -1 | Yes |
| $\mathrm{D}^{\prime}$ | $\pi$ | $\pi$ | -1 |  |

Fig. 49
Case-II $0 \leq \phi_{1}<2 \pi,-\pi \leq \phi_{2}<\pi$

| on the <br> segment | $\phi_{1}$ | $\phi_{2}$ | $\mathrm{e}^{\mathrm{i}\left(\phi_{1}+\phi_{2}\right) / 2}$ | Continuity <br> of $\mathrm{f}(\mathrm{z})$ |
| :--- | :---: | :---: | :---: | :---: |
| B | $\pi$ | 0 | i | Yes |
| $\mathrm{B}^{\prime}$ | $\pi$ | 0 | i |  |
| C | 0 | 0 | 1 | No |
| $\mathrm{C}^{\prime}$ | $2 \pi$ | 0 | -1 |  |
| D | $\pi$ | $\pi$ | -1 | No |
| $\mathrm{D}^{\prime}$ | $\pi$ | $-\pi$ | 1 |  |



Fig. 50 Branch cut $[-1,1]$


Fig. 51 Branch cuts $(-\infty,-1]$ and $[1, \infty)$

Two branches of $(z-1)^{1 / 2}$ can be taken as

$$
\mathrm{f}_{1}(\mathrm{z})=\sqrt{\mathrm{r}_{1} \mathrm{e}^{\mathrm{i} \phi_{1} / 2}} \text { and } \mathrm{f}_{2}(\mathrm{z})=\sqrt{\left.\mathrm{r}_{1} \mathrm{e}^{\mathrm{i}} \mathrm{i}_{1}+2 \pi\right) / 2}, 0 \leq \phi_{1}<2 \pi=-\mathrm{f}_{1}(\mathrm{z})
$$

Again two branches of $(\mathrm{z}+1)^{1 / 2}$ can be taken as

$$
\begin{aligned}
\mathrm{g}_{1}(\mathrm{z}) & =\sqrt{\mathrm{r}_{2} \mathrm{e}^{\mathrm{i} \phi_{2} / 2}} \text { and } \mathrm{g}_{2}(\mathrm{z})=\sqrt{\mathrm{r}_{2} \mathrm{e}^{\mathrm{i}\left(\phi_{2}+2 \pi\right) / 2}, 0 \leq \phi_{2}<2 \pi} \\
& =-\mathrm{g}_{1}(\mathrm{z})
\end{aligned}
$$

Let us construct a Riemann surface for $\omega=\left(z^{2}-1\right)^{1 / 2}$ considering case I.
Here a Riemann surface consists of two sheets So and $S_{1}$. Let $S_{0}$ be an extended complex plane cut along the real axis from $\mathrm{z}=-1$ to $\mathrm{z}=1$ and $\mathrm{S}_{1}$ be another extended complex plane cut of similar nature.

$$
S_{0}\left\{\begin{array} { l } 
{ 0 \leq \operatorname { A r g } ( z - 1 ) < 2 \pi } \\
{ 0 \leq \operatorname { A r g } ( z + 1 ) < 2 \pi }
\end{array} \quad S _ { 1 } \left\{\begin{array}{l}
2 \pi \leq \operatorname{Arg}(z-1)<4 \pi \\
2 \pi \leq \operatorname{Arg}(z+1)<4 \pi
\end{array}\right.\right.
$$

The sheets $S_{0}$ and $S_{1}$ are joined along the segment $[-1,1]$ in such a way that the lower edge of the slit in $S_{0}$ is joined to the upper edge of the slit in $S_{1}$, and the lower edge of the slit in $S_{1}$ is joined to the upper edge of the slit in $S_{0}$.

Let a point on the sheet $S_{0}$ move anticlockwise and form a simple closed curve which encloses the segment $[-1,1]$ once. Then both $\phi_{1}$ and $\phi_{2}$ change by an amount $2 \pi$ upon returning to their original position. i.e. $\left(\phi_{1}+\phi_{2}\right) / 2$ changes by an amount $2 \pi$, so the value of

$$
\omega=\left(r_{1} r_{2}\right)^{1 / 2} \mathrm{e}^{\mathrm{i}\left(\phi_{1}+2 \pi+\phi_{2}+2 \pi\right) / 2}=\left(\mathrm{r}_{1} \mathrm{r}_{2}\right)^{1 / 2} \mathrm{e}^{\mathrm{i}\left(\phi_{1}+\phi_{2}\right) / 2}
$$

remains unchanged.
Then $\omega=f_{1} g_{1}$ on $S_{0}$ and as well as on $S_{1}$.
If a point starting on the sheet $S_{0}$ travels a path which makes a complete round about only the branch point $z=1$, it crosses from the sheet $S_{0}$ to $S_{1}$. In this case, the value of $\phi_{1}$ changes by an amount $2 \pi$, while the value of $\phi_{2}$ does not change at all. The change in $\left(\phi_{1}+\phi_{2}\right) / 2$ is then $\pi$. The change in $\left(\phi_{1}+\phi_{2}\right) / 2$ remains the same if a point on the sheet $S_{0}$ makes a complete round about the branch point $z=-1$ only and enters on the $S_{1}$ sheet. This time.

$$
\omega=\left\{\begin{array}{l}
\mathrm{f}_{1} \mathrm{~g}_{1} \text { on } \mathrm{S}_{0} \\
-\mathrm{f}_{1} \mathrm{~g}_{1} \text { on } \mathrm{S}_{1}
\end{array}\right.
$$

Thus the double-valued function $\omega=\left(z^{2}-1\right)^{1 / 2}$ can now be considered as a single-valued function on the Riemann surface constructed above. Hence the transformation $\omega=\left(z^{2}-1\right)^{1 / 2}$ maps each of the sheets $S_{0}$ and $S_{1}$ forming the Riemann surface on the entire $\omega$-plane.

## Riemann surface for the case II

Here the Riemann surface is formed by two sheets $S_{0}$ and $S_{1}$. Each of these sheets is an extended complex plane cut along the line $(-\infty,-1) \cup[1, \infty)$

$$
S_{0}\left\{\begin{array} { l } 
{ 0 \leq \operatorname { A r g } ( z - 1 ) < 2 \pi } \\
{ - \pi \leq \operatorname { A r g } ( z + 1 ) < \pi }
\end{array} \quad S _ { 1 } \left\{\begin{array}{c}
2 \pi \leq \operatorname{Arg}(z-1)<4 \pi \\
\pi \leq \operatorname{Arg}(z+1)<3 \pi
\end{array}\right.\right.
$$

These sheets are joined along the line $(-\infty,-1] \cup[1, \infty)$ in such a way that the lower edge of the slit in $S_{0}$ is joined to the upper edge of the slit in $S_{1}$, and the lower edge of the slit in $S_{1}$ is joined to the upper edge of the slit in $\mathrm{S}_{0}$.

If a point traverses a simple closed curve on either of the sheets $S_{0}$ or $S_{1}$ not enclosing any of the branch points -1 or 1 , then the function $f(z)$ remains single-valued on the respective sheet, whereas if it encloses any one of the branch points the function changes the branch as explained in case I. In the same way the double-valued function $f(z)=\left(z^{2}-1\right)^{1 / 2}$ can be treated as a single-valued function on the Riemann surface formed earlier.

Example 21 : The power function $\omega=f(z)=[z(z-1)(z-2)]^{1 / 2}$ has two branches. Show that $f(-1)$ can be either $-\sqrt{ } 6 i$ or $\sqrt{ } 6$ i. Suppose the branch that corresponds to $f(-1)=-\sqrt{6} \mathrm{i}$ is chosen, find the value of the function at $\mathrm{z}=\mathrm{i}$.

Solution : The given power function can be expressed as

$$
\omega=\mathrm{f}(\mathrm{z})=\sqrt{|\mathrm{z}(\mathrm{z}-1)(\mathrm{z}-2)|} \mathrm{e}^{\mathrm{i}[\operatorname{Argz} z \operatorname{Arg}(\mathrm{z-1})+\operatorname{Arg}(\mathrm{z}-2) / / 2} \mathrm{e}^{\mathrm{ik} \pi}, \mathrm{k}=0,1
$$

where the two possible values of k correspond to the two branches of the doublevalued power function.

If figure 52a branch cuts are $y=0, x \leq 0$ and $y=0,1 \leq x \leq 2$ and in figure 52b branch cuts are $y=0,0 \leq x \leq 1$ and $y=0, x \geq 2$. In both the cases Riemann surface is formed by
$\qquad$
Fig. 52a
$\qquad$
Fig. 52b two branches.

At $\mathrm{z}=-1$, we note that

$$
\operatorname{Arg} z=\operatorname{Arg}(z-1)=\operatorname{Arg}(z-2)=\pi \text { and } \sqrt{|z(z-1)(z-2)|}=\sqrt{6}
$$

So, $\mathrm{f}(-1)$ can be either $\sqrt{6 \mathrm{e}}{ }^{\mathrm{i} 3 \pi / 2}=-\sqrt{6} \mathrm{i}$ or $\sqrt{6} \mathrm{e}^{\mathrm{i}(\pi+2 \pi+\pi+2 \pi+\pi+2 \pi)}=\sqrt{6} \mathrm{e}^{\mathrm{i} 3 \pi / 2} \mathrm{e}^{\mathrm{i} \pi}=\sqrt{6} \mathrm{i}$.
The branch that gives $f(-1)=\sqrt{6 i}$ corresponds to $k=0$. With the choice of that branch, we find

$$
\begin{aligned}
& \mathrm{f}(\mathrm{i})=\sqrt{|\mathrm{i}(\mathrm{i}-1)(\mathrm{i}-2)|} \mathrm{e}^{\mathrm{i} \mid \operatorname{Argi}+\operatorname{Arg}(\mathrm{i}-1)+\operatorname{Arg}(\mathrm{i}-2) / 2} \\
& =\sqrt{\sqrt{2 \sqrt{5}}} \mathrm{e}^{\mathrm{i}\left(\pi / 2+3 \pi / 4+\pi-\tan ^{-1} 1 / 2\right) / 2}=\sqrt[4]{10} \mathrm{e}^{\mathrm{i}\left(\frac{\pi}{4}-\tan ^{-\frac{1}{2}}\right)^{2} / 2^{2}} \mathrm{e}^{\mathrm{i} \pi} \\
& =-\sqrt[4]{10} \mathrm{e}^{1\left(\tan ^{-1} 1-\tan ^{-1} 1 / 2\right) / 2}=-\sqrt[4]{10} \mathrm{e}^{\mathrm{i}\left(\tan ^{-1} 1 / 3\right) / 2}
\end{aligned}
$$

### 4.11 The Inverse Trigonometric Functions

(i) The function $\sin ^{-1} \mathrm{z}$ is defined by the equation

$$
\mathrm{z}=\sin \omega
$$

Substituting $\frac{\mathrm{e}^{\mathrm{i} \omega}-\mathrm{e}^{-\mathrm{i} \omega}}{2 \mathrm{i}}$ for $\sin \omega$, we find that

$$
\left(\mathrm{e}^{\mathrm{i} \omega}\right)^{2}-2 \mathrm{ie}^{\mathrm{i} \omega} \mathrm{Z}-1=0
$$

i.e., $\quad e^{i \omega}=i z+\left(1-z^{2}\right)^{1 / 2}$
$\Rightarrow \quad i \omega=\log \left\{\mathrm{iz}+\left(1-\mathrm{z}^{2}\right)^{1 / 2}\right\}$
so that $\quad \sin ^{-1} \mathrm{Z}=-\mathrm{ilog}\left\{\mathrm{iz}+\left(1-\mathrm{z}^{2}\right)^{1 / 2}\right\}$
Similarly, we can have

$$
\cos ^{-1} \mathrm{z}=-\mathrm{i} \log \left\{\mathrm{z}+\left(\mathrm{z}^{2}-1\right)^{1 / 2}\right\}
$$

(ii) We take the function $\omega=\tan ^{-1} \mathrm{z}$, which is the inverse of $\mathrm{z}=\tan \omega$. Expressing $\tan \omega$ in terms of $\sin \omega$ and $\cos \omega$ and then converting to their exponential form, we get

$$
\begin{aligned}
& \qquad \begin{array}{l}
z=\frac{1 e^{i \omega}-e^{-i \omega}}{i e^{i \omega}+e^{-i \omega}} \\
\\
=\frac{1 e^{2 i \omega}-1}{i e^{2 i \omega}+1} \\
\text { i.e., } \\
\text { and finally, } \quad i z=\frac{e^{2 i \omega}-1}{e^{2 i \omega}+1} \Rightarrow e^{2 i \omega}=\frac{1+i z}{1-i z} \\
\end{array} \quad \omega=\frac{1}{2 \mathrm{i}} \log \frac{1+i z}{1-\mathrm{iz}}
\end{aligned}
$$

Note : When $z \neq \pm 1$, the quantity $\left(1-z^{2}\right)^{1 / 2}$ has two possible values. For each value, the logarithm generates infinitely many values. Therefore $\sin ^{-1} \mathrm{z}$ has two sets of infinite values. For example, consider

$$
\begin{aligned}
\sin ^{-1} \frac{1}{2} & =\frac{1}{\mathrm{i}} \log \left(\frac{\mathrm{i}}{2} \pm \frac{\sqrt{3}}{2}\right) \\
& =\frac{1}{\mathrm{i}}\left[\log \mathrm{e}^{\mathrm{i}\left(\frac{\pi}{6}+2 \mathrm{k} \pi\right)}\right] \text { or } \frac{1}{\mathrm{i}}\left[\log \mathrm{e}^{\mathrm{i}\left(\frac{5 \pi}{6}+2 \mathrm{k} \pi\right)}\right] \\
& =\frac{1}{\mathrm{i}}\left[\mathrm{i}\left(\frac{\pi}{6}+2 \mathrm{k} \pi\right)\right] \text { or } \frac{1}{\mathrm{i}}\left[\mathrm{i}\left(\frac{5}{6} \pi+2 \mathrm{k} \pi\right)\right] \\
& =\frac{\pi}{6}+2 \mathrm{k} \pi \text { or } \frac{5 \pi}{6}+2 \mathrm{k} \pi, \mathrm{k} \text { is any integer. }
\end{aligned}
$$

Likewise, the set of values for other inverse trigonometric functions can be ascertained.

Example 22 : Discuss the mapping $\omega=\sinh z$ that transforms the infinite strip $-\infty<\mathrm{x}<\infty, 0<\mathrm{y}<\pi$ into the $\omega$-plane. Find cuts in the $\omega$-plane which make the mapping continuous both ways. What are the images of the lines (a) $y=1 / \pi$ (b) $\mathrm{x}=1$ ?

Solution : First we express sinh z in cartesian form

$$
\omega=\sinh z=\sinh x \cos y+i \cosh x \sin y=u+i v
$$

We consider the line segment $\mathrm{x}=\mathrm{c}, \mathrm{y} \varepsilon(0, \pi)$. Its image is

$$
\{\sinh c \cos y+i \cosh c \sin y \mid y \varepsilon(0, \pi)\}
$$

Clearly, $u$ and $v$ then satisfy the equation for the ellipse

$$
\frac{\mathrm{u}^{2}}{\sinh ^{2} \mathrm{c}}+\frac{\mathrm{v}^{2}}{\cosh ^{2} \mathrm{c}}=1
$$

The ellipse starts at the point ( $\sinh \mathrm{c}, 0$ ), passes through the point $(0, \cosh \mathrm{c}$ ) and ends at ( $-\sinh \mathrm{c}, 0$ ). As c varies from zero to $\infty$ or from zero to $-\infty$, the semi-ellipses cover the upper-half of $\omega$-plane. Thus the mapping is $2-$ to -1 .

Now consider the infinite line $\mathrm{y}=\mathrm{c}, \mathrm{x} \in(-\infty, \infty)$.
It's image is $\{\sinh \mathrm{x} \cos \mathrm{c}+\mathrm{i} \cosh \mathrm{x} \sin \mathrm{c} \mid \mathrm{x} \in(-\infty, \infty)\}$.
Here $u$ and $v$ satisfy the equation for a hyperbola

$$
\frac{u^{2}}{\cos ^{2} c}-\frac{v^{2}}{\sin ^{2} c}=1
$$

As c varies from 0 to $\pi / 2$ or from $\pi / 2$ to $\pi$, the semi-hyperbola cover the upperhalf of $\omega$-plane. Thus the mapping is 2 -to- 1 .

We look for branch points of $\sinh ^{-1} \omega$

$$
\begin{aligned}
& \omega=\sinh \mathrm{z} \\
& \omega=\frac{\mathrm{e}^{\mathrm{z}}-\mathrm{e}^{-\mathrm{z}}}{2} \\
& \mathrm{e}^{2 \mathrm{z}}-2 \omega \mathrm{e}^{\mathrm{z}}-1=0 \\
& \mathrm{e}^{\mathrm{z}}=\omega+\left(\omega^{2}+1\right)^{1 / 2} \\
& \mathrm{z}=\log \left(\omega+(\omega-\mathrm{i})^{1 / 2}(\omega+\mathrm{i})^{1 / 2}\right)
\end{aligned}
$$

The branch points are at $\omega= \pm \mathrm{i}$. Since $\omega+\left(\omega^{2}+1\right)^{1 / 2}$ is non zero and finite in the finite complex plane, the logarithm does not introduce any branch in the finite plane. Thus the only branch point in the upper-half of $\omega$-plane is at $\omega=\mathrm{i}$. Any branch cut that connects $\omega=\mathrm{i}$ with the boundary of $\operatorname{Im} \omega>0$ will separate the branches under the inverse mapping.

We consider the line $\mathrm{y}=\pi / 4$. The image under the mapping is the upper-half of the hyperbola

$$
2 u^{2}-2 v^{2}=1
$$

Consider the segment $\mathrm{x}=1$. The image under the mapping is the upper-half of the ellipse.

$$
\frac{\mathrm{u}^{2}}{\sinh ^{2} 1}+\frac{\mathrm{v}^{2}}{\cosh ^{2} 1}=1
$$

Example 23 : Construct a Riemann Surface for $\cos ^{-1} \mathrm{z}$.
Solution : The function $\omega=\cos ^{-1} \mathrm{z}=-\mathrm{i} \log \left[\mathrm{z}+\left(\mathrm{z}^{2}-1\right)^{1 / 2}\right]$ has two sources of multi-valuedness; one due to the square root function $\left(z^{2}-1\right)^{1 / 2}$ and the other due to the logarithm, if any.

First we construct the branch of the square root

$$
\left(z^{2}-1\right)^{1 / 2}=(z+1)^{1 / 2}(z-1)^{1 / 2}
$$

We see that there are branch points at $\mathrm{z}=-1$ and $\mathrm{z}=1$. In particular we want the $\cos ^{-1} \mathrm{z}$ to be defined for $\mathrm{z}=\mathrm{x}, \mathrm{x} \in[-1,1]$. Hence we introduce the branch cuts on the lines $(-\infty,-1]$ and $[1, \infty)$. Let

$$
\mathrm{z}+1=\mathrm{re}^{\mathrm{i} \theta}, \mathrm{z}-1=\rho \mathrm{e}^{\mathrm{i} \phi}
$$

With the given branch cuts, the angles have the possible ranges

$$
-\pi \leq \theta<\pi, 0 \leq \phi<2 \pi
$$

Now we must determine if the logarithm introduces


Fig. 53 additional branch points. The only possibilities for branch points are where the argument of the logarithm is zero.

$$
z+\left(z^{2}-1\right)^{1 / 2}=0
$$

or, $z^{2}=z^{2}-1 \Rightarrow 0=-1$
We see that the argument of the logarithm can not be zero and thus there are no additional branch points. Here the Riemann surface consists of two sheets $\mathrm{S}_{0}$ and $\mathrm{S}_{1}$ joined on the real axis $(-\infty,-1] \cup[1, \infty)$ :

$$
\mathrm{S}_{0}\left\{\begin{array} { l } 
{ 0 \leq \phi < 2 \pi } \\
{ - \pi \leq \theta < \pi }
\end{array} \quad \mathrm { S } _ { 1 } \left\{\begin{array}{c}
2 \pi \leq \phi<4 \pi \\
\pi \leq \theta<3 \pi
\end{array}\right.\right.
$$

A particular branch of this function can be obtained by first taking

$$
\mathrm{z}+1=\mathrm{re}^{\mathrm{i} \theta},-\pi \leq \theta<\pi ; \mathrm{z}-1=\rho \mathrm{e}^{\mathrm{i} \phi}, 0 \leq \phi<2 \pi
$$

Then adding these relations, we find

$$
\mathrm{z}=\left(\mathrm{re}^{\mathrm{i} \theta}+\rho \mathrm{e}^{\mathrm{i} \phi}\right) / 2
$$

and the function $\mathrm{z}+\left(\mathrm{z}^{2}-1\right)^{1 / 2}$ reduces to

$$
\begin{aligned}
\mathrm{z}+\left(\mathrm{z}^{2}-1\right)^{1 / 2} & =\frac{\mathrm{re}^{\mathrm{i} \theta}+\rho \mathrm{e}^{\mathrm{i} \phi}}{2}+(\mathrm{r} \rho)^{1 / 2} \mathrm{e}^{\mathrm{i}(\theta+\phi) / 2} \\
& =\frac{\mathrm{re}^{\mathrm{i} \theta}}{2}\left(1+\frac{\rho}{\mathrm{r}} \mathrm{e}^{\mathrm{i}(\phi-\theta)}+2 \sqrt{\frac{\rho}{\mathrm{r}}} \mathrm{e}^{\mathrm{i}(\phi-\theta) / 2}\right)
\end{aligned}
$$

$$
=\frac{\mathrm{re}^{\mathrm{i} \theta}}{2}\left(1+\sqrt{\frac{\rho}{\mathrm{r}}}_{\mathrm{e}^{\mathrm{i}(\phi-\theta) / 2}}\right)^{2}
$$

Then $\cos ^{-1} \mathrm{z}=-\mathrm{i}\left\{\log \left(\frac{\mathrm{r}}{2} \mathrm{e}^{\mathrm{i} \theta}\right)+\log \left(1+\sqrt{\frac{\rho}{\mathrm{r}}} \mathrm{e}^{\mathrm{i}(\phi-\theta) / 2}\right)^{2}\right\}$ on $\mathrm{S}_{0}$. If a point lying on the sheet $\mathrm{S}_{0}$ is allowed to travel a path making a complete round about only the branch point $\mathrm{z}=1$, it enters to the sheet $S_{1}$ from the sheet $S_{0}$. In this case the value of $\phi$ changes by $2 \pi$ while the value of $\theta$ remains unchanged. The change in $(\phi-\theta) / 2$ is $\pi$. So in this case,

$$
\cos ^{-1} \mathrm{z}=-\mathrm{i}\left\{\log \left(\frac{\mathrm{r}}{2} \mathrm{e}^{\mathrm{i} \theta}\right)+\log \left(1-\sqrt{\frac{\rho}{\mathrm{r}}} \mathrm{e}^{\mathrm{i}(\phi-\theta) / 2}\right)^{2}\right\} \text { on } \mathrm{S}_{1} \text {. Similarly we can analyse }
$$

the case when the point on $\mathrm{S}_{0}$ encloses only the branch point $\mathrm{z}=-1$ while travelling a complete round.

Some standard branch cuts of elementary functions.

| Function | Branch cuts |
| :--- | :---: |
| $\mathrm{z}^{\mathrm{s}}$, non integral s with Re $\mathrm{s}>0$ | $(-\infty, 0)$ |
| $\mathrm{z}^{\mathrm{s}}$, non integral s with Re $\mathrm{s} \leq 0$ | $(-\infty, 0]$ |
| $\mathrm{e}^{\mathrm{Z}}$ | none |
| $\log \mathrm{z}$ | $(-\infty, 0]$ |
| $\sin ^{-1} \mathrm{z}, \cos ^{-1} \mathrm{z}$ | $(-\infty,-1]$ and $[1, \infty)$ |
| $\tan ^{-1} \mathrm{z}$ | $\mathrm{y} \leq-1, \mathrm{x}=0$ and $\mathrm{y} \geq 1, \mathrm{x}=0$ |
| $\operatorname{cosec}^{-1} \mathrm{z}, \sec ^{-1} \mathrm{z}$ | $(-1,1)$ |
| $\cot ^{-1} \mathrm{z}$ | $[-\mathrm{i}, \mathrm{i}]$ |
| $\sinh ^{-1} \mathrm{z}$ | $\mathrm{y}<-1, \mathrm{x}=0$ and $\mathrm{y}>1, \mathrm{x}=0$ |
| $\cosh ^{-1} \mathrm{z}$ | $(-\infty, 1)$ |
| $\operatorname{cosech}^{-1} \mathrm{z}$ | $-1<\mathrm{y}<1, \mathrm{x}=0$ |
| $\operatorname{sech}^{-1} \mathrm{z}$ | $(-\infty, 0]$ and $(1, \infty)$ |
| $\tanh ^{-1} \mathrm{z}$ | $\mathrm{y} \leq 1, \mathrm{x}=0$ and $\mathrm{y} \geq 1, \mathrm{x}=0$ |
| $\operatorname{coth}^{-1} \mathrm{z}$ | $[-1,1]$ |

## Exercises

1. Find the principal value of each of the following complex quantities :
(a) $(1-i)^{1+i}$
(b) $3^{3-\mathrm{i}}$
(c) $2^{2 i}$
2. Give the number of branches and locations of the branch points for the functions.
(a) $\cos \left(z^{1 / 2}\right)$
(b) $(\mathrm{z}+\mathrm{i})^{-\mathrm{z}}$
3. Determine the branch points of the function
$\omega=\left\{\left(z^{2}-z\right)(z+2)\right\}^{1 / 3}$
4. Find the branch points of $\left(z^{1 / 2}-1\right)^{1 / 2}$ in the finite complex plane. Introduce branch cuts to make the function single-valued.
5. Let D be the complex z -plane with a cut along the segment $[-1,1]$, determine the regular branches of the function

$$
f(z)=\left(\frac{1-z}{1+z}\right)^{1 / 2}
$$

6. Split the function $f(z)=\sqrt{\left(z^{2}-4\right)\left(z^{2}-9\right)}$ into two regular branches in the domain $\mathrm{D}: \mathbb{C} \backslash\{[-3,-2],[2,3]\}$
7. Evaluate
(i) $\int_{0}^{\infty} \frac{\mathrm{x}^{\alpha}}{\mathrm{x}^{2}-1} \mathrm{dx},-1<\alpha<1$
(ii) $\int_{0}^{\infty} \frac{\log x}{x^{2}+1} d x$
8. Prove that $\int_{0}^{\pi} \log \sin x d x=-\pi \log 2$.
9. Construct a Riemann surface for the following functions :
(i) $\omega=z^{1 / 3}$ (ii) $\omega=\left(z^{2}+1\right)^{1 / 2}$ (iii) $\omega=\log \frac{z+1}{z-1}$ (iv) $\omega=\sin ^{-1} z$.
10. Let $\mathrm{f}(\mathrm{z})$ have branch points at $\mathrm{z}=0$ and $\mathrm{z}= \pm \mathrm{i}$ but nowhere else in the extended complex plane. How does the value and argument of $f(z)$ change while traversing the contour given in the figures 51 (a) (b). Do the branch cuts make the function single valued?


Fig. 54 (a)


Fig. 54 (b)

## Unit 5 - Conformal Equivalence

## Structure

### 5.0 Objectives

### 5.1 Riemann Mapping Theorem

### 5.2 The Schwarz Reflection Principle

### 5.3 The Schwarz-Christoffel Transformation

### 5.4 Examples : Triangles / Rectangles

### 5.0 Objectives of this Chapter

The concept of conformal equivalence of two regions will be introduced in this chapter. The main theorem of this chapter is Riemann mapping theorem. Also Hurwitz's theorem, Schwarz lemma, Schwarz reflection principle, Schwarz-Christoffel transformation will be studied and their applications will be shown through a few examples.

### 5.1 Riemann Mapping Theorem

In the family of analytic functions that concern geometrical orientation, conformal mapping plays a leading role. As its consequences we shall present here a most important result named after G. F. B Riemann, known as "Riemann mapping theorem". Throughout $\mathrm{H}(\mathrm{G})$ will denote the family of analytic functions defined on the region G .

## Definition : Conformal Equivalence :

Two regions $R_{1}$ and $R_{2}$ are said to be conformally equivalent if there exists a $f \in H\left(R_{1}\right)$ such that $f$ is one-to-one in $R_{1}$ and $f\left(R_{1}\right)=R_{2}$ i.e. if there exists a conformal mapping one to one of $R_{1}$ onto $R_{2}$. Clearly, this is an equivalence relation (reflexive, symmetric and transitive).

Theorem 5.1 [Hurwitz's Theorem] Let $G$ be a region and $\left\{\mathrm{f}_{\mathrm{n}}\right\}$ be a sequence in $H(G)$ that converges uniformly to $f \in H(G)$. Suppose $f \neq 0, \bar{D}(a, R) \subset G$ and $f(z)$ $\neq 0$ on $\gamma:|z-a|=R$. Then there exists an integer $N$ such that for $n \geq N, f_{n}$ and $f$ have the same number of zeros in $\mathrm{D}(\mathrm{a}, \mathrm{R})$.

Proof. Since $\mathrm{f}(\mathrm{z})$ is never zero on the circle $\gamma$, we have

$$
\operatorname{Inf}_{\gamma}|f(z)|=\delta>0
$$

Again, $\mathrm{f}_{\mathrm{n}} \rightarrow \mathrm{f}$ uniformly on $\gamma$, so there is an integer N such that for $\mathrm{n} \geq \mathrm{N}$

$$
\sup _{\gamma}\left|\mathrm{f}_{\mathrm{n}}(\mathrm{z})-\mathrm{f}(\mathrm{z})\right|<\frac{\delta}{2}
$$

and thus on the circle $\gamma,\left|\mathrm{f}(\mathrm{z})-\mathrm{f}_{\mathrm{n}}(\mathrm{z})\right|<\frac{\delta}{2}<\delta \leq|\mathrm{f}(\mathrm{z})|$ for $\mathrm{n} \geq \mathrm{N}$. Using Rouche's theorem we find that $f_{n}$ and $f$ have the same number of zeros inside the circle $\gamma:|\mathrm{z}-\mathrm{a}|=\mathrm{R}$ for $\mathrm{n} \geq \mathrm{N}$.

By means of the above theorem, we can easily prove
Corollary 1. Let $G$ be a region and $\left\{f_{n}\right\}$ be a sequence in $H(G)$ such that each $f_{n}$ never vanishes in G. Suppose $f_{n} \rightarrow f$ uniformly in $H(G)$. Then $f(z)$ never vanishes in G , unless $\mathrm{f} \equiv 0$.

## Some useful results

(i) If $f(z)$ is analytic at $z_{0}$ and $f^{1}\left(z_{0}\right) \neq 0$, then there is a neighbourhood of $z_{0}$ in which $f(z)$ is univalent.
(ii) An univalent analytic function f on a domain G has a non-zero derivative at every point of G, i.e., $f^{1}(z) \neq 0$ on G.
(iii) The inverse of an univalent analytic function is analytic.
(iv) Any domain in $\mathbb{C}$, that is conformally equivalent to a simply connected domain must itself be simply connected.
(v) A domain D in $\mathbb{C}$ is simply connected if and only if every analytic function in D has a primitive in D.

## Schwarz Lemma

Let $\mathrm{f}: \mathrm{D}(0,1) \rightarrow \mathrm{D}(0,1)$ be an analytic function which maps the unit disc $\mathrm{D}(0,1)$ to itself. If $\mathrm{f}(0)=0$,
then
(i) $|\mathrm{f}(\mathrm{z})| \leq|\mathrm{z}|$ for $0 \leq|\mathrm{z}|<1$
(ii) $\left|\mathrm{f}^{1}(0)\right| \leq 1$
(iii) if equality holds in (i) for at least one $z \in D(0,1)-\{0\}$, or, if equality holds in (ii), then

$$
\mathrm{f}(\mathrm{z})=\lambda \mathrm{z},
$$

where $\lambda$ is a constant, $|\lambda|=1$.
Proof : Let us consider the function

$$
g(z)=\frac{f(z)}{z}
$$

which is analytic in the disc $\mathrm{D}(0,1)-\{0\}$ and it has removable singularity at z $=0$, since $f(0)=0$. It can be made analytic at $z=0$ if we define

$$
\begin{equation*}
g(0)=\lim _{z \rightarrow 0} \frac{f(z)}{z}=f^{1}(0) \tag{55}
\end{equation*}
$$

For $|z|=r$, where $0<r<1$

$$
|\mathrm{g}(\mathrm{z})|=\frac{|\mathrm{f}(\mathrm{z})|}{|\mathrm{z}|}<\frac{1}{\mathrm{r}}
$$

By the Maximum Modulus Principle, $|\mathrm{g}(\mathrm{z})|<1 / \mathrm{r}$ for the entire disc $|\mathrm{z}| \leq \mathrm{r}$. We fix $\mathrm{z} \in \mathrm{D}(0,1)-\{0\}$ and let $\mathrm{r} \rightarrow 1$. Then

$$
|\mathrm{g}(\mathrm{z})| \leq 1
$$

This is true for all $\mathrm{z} \in \mathrm{D}(0,1)-\{0\}$ and we get

$$
\begin{equation*}
\frac{|\mathrm{f}(\mathrm{z})|}{|\mathrm{z}|} \leq 1, \quad 0<|\mathrm{z}|<1 \tag{56}
\end{equation*}
$$

i.e. $|f(z)| \leq|z|, 0<|z|<1$. Since $f(0)=0$, we have $|f(z)| \leq|z|$ for $0 \leq|z|<1$. So,
(i) is proved and we find from (55) that $|\mathrm{g}(0)|=\left|\mathrm{f}^{1}(0)\right| \leq 1$ which proves (ii) To prove (iii), we observe that if at a point $\mathrm{z}_{0} \neq 0\left(\left|\mathrm{z}_{0}\right|<1\right)\left|\mathrm{g}\left(\mathrm{z}_{0}\right)\right| 1=1$ i.e. $|\mathrm{g}(\mathrm{z})|$ attains its maximum at an internal point and hence by the maximum modulus principle $\mathrm{g}(\mathrm{z})=\lambda$, a constant and that $|\lambda|=1$, so $\mathrm{f}(\mathrm{z})=\lambda \mathrm{z}$.

Theorem 5.2 Let $\mathrm{a} \in \mathrm{D}(0,1)$. Then $\phi_{\mathrm{a}}$ defined by

$$
\phi_{\mathrm{a}}(\mathrm{z})=\frac{\mathrm{z}-\mathrm{a}}{1-\overline{\mathrm{a}} \mathrm{z}}
$$

maps $\overline{\mathrm{D}}(0,1)$ onto $\overline{\mathrm{D}}(0,1)$.
Proof. Clearly, $\phi_{a}$ is a bilinear transformation, it is analytic in the whole complex plane except the point $\frac{1}{\overline{\mathrm{a}}}$ (which is the inverse point of the point a with respect to the circle $|z|=1$, and hence lies outside $|z|=1$ ). We observe that

$$
\begin{aligned}
\phi_{\mathrm{a}}\left(\phi_{-\mathrm{a}}(\mathrm{z})\right) & =\frac{\frac{\mathrm{z}+\mathrm{a}}{1+\overline{\mathrm{a}} \mathrm{z}}-\mathrm{a}}{1-\overline{\mathrm{a}} \frac{\mathrm{z}+\mathrm{a}}{1+\overline{\mathrm{a}} \mathrm{z}}} \\
& =\frac{\mathrm{z}\left(1-|\mathrm{a}|^{2}\right)}{1-|\mathrm{a}|^{2}} \\
& =\mathrm{z}=\phi_{-\mathrm{a}}\left(\mathrm{f}_{\mathrm{a}}(\mathrm{z})\right), \text { similarly. }
\end{aligned}
$$

Thus $\phi_{\mathrm{a}}$ maps $\mathrm{D}(0,1)$ onto $\mathrm{D}(0,1)$ in a one to one way. Now let $\theta$ be a real number. Then

$$
\begin{aligned}
& \left|\phi_{\mathrm{a}}\left(\mathrm{e}^{\mathrm{i} \theta}\right)\right|=\left|\frac{\mathrm{e}^{\mathrm{i} \theta}-\mathrm{a}}{1-\overline{\mathrm{a}} \mathrm{e}^{\mathrm{i} \theta}}\right| \\
= & \left|\frac{\mathrm{e}^{\mathrm{i} \theta}-\mathrm{a}}{\mathrm{e}^{-\mathrm{i} \theta}-\overline{\mathrm{a}}}\right|\left|\frac{1}{\mathrm{e}^{i \theta}}\right|=\left|\frac{\mathrm{e}^{\mathrm{i} \theta}-\mathrm{a}}{\overline{\mathrm{e}^{i \theta}-\mathrm{a}}}\right|=1
\end{aligned}
$$

i.e., $\phi_{\mathrm{a}}$ maps $|\mathrm{z}|=1$ on $|\mathrm{z}|=1$. Thus, $\phi_{\mathrm{a}}$ maps $\overline{\mathrm{D}}(0,1)$ onto $\overline{\mathrm{D}}(0,1)$.

## A maximal problem

Let $\alpha, \beta$ be two complex numbers with $|\alpha|<1,|\beta|<1$ and f be analytic on $D(0,1)$ satisfying $f(\alpha)=\beta$. What is the maximum possible value of $\left|f^{1}(\alpha)\right|$ among such mappings?

Solution : Let

$$
\begin{equation*}
\mathrm{g}=\phi_{\beta} 0 \mathrm{f} 0 \phi_{-\alpha} \text { where } \phi_{\beta} \text { is defined as in theorem } 5.2 \tag{57}
\end{equation*}
$$

Then g maps $\mathrm{D}(0,1)$ to $\mathrm{D}(0,1)$ and satisfies

$$
\begin{aligned}
\mathrm{g}(0) & =\phi_{\beta}\left\{\mathrm{f}\left(\phi_{-\alpha}(0)\right)\right\} \\
& =\phi_{\beta}\{\mathrm{f}(\alpha)\} \\
& =\phi_{\beta}(\beta) \\
& =0
\end{aligned}
$$

Thus $g$ satisfies all the conditions of Schwaz's lemma and hence $\left|g^{1}(0)\right| \leq 1$. To obtain an explicit form of $\mathrm{g}^{1}(0)$, we use (57) and apply the chain rule

$$
\begin{aligned}
g^{1}(0) \quad & =\left\{\left(\phi_{\beta} 0 f\right)^{1}\left(\phi_{-\alpha}(0)\right\} \phi^{1}{ }_{-\alpha}(0)\right. \\
& =\left(\phi_{\beta} 0 f\right)^{1}(\alpha)\left(1-|\alpha|^{2}\right) \\
& =\phi_{\beta}^{1}(\mathrm{f}(\alpha)) \mathrm{f}^{1}(\alpha)\left(1-|\alpha|^{2}\right) \\
& =\phi_{\beta}{ }^{1}(\beta) \mathrm{f}^{1}(\alpha)\left(1-|\alpha|^{2}\right) \\
& =\frac{1-|\alpha|^{2}}{1-|\beta|^{2}} \mathrm{f}^{1}(\alpha)
\end{aligned}
$$

But $\left|g^{1}(0)\right| \leq 1$, therefore

$$
\begin{equation*}
\left|\mathrm{f}^{1}(\alpha)\right| \leq \frac{1-|\beta|^{2}}{1-|\alpha|^{2}} \tag{58}
\end{equation*}
$$

Equality in (58) occurs only when $\left|g^{1}(0)\right|=1$. In that case by virtue of Schwarz
lemma there is a constant $\lambda,|\lambda|=1$ so that $g(z)=\lambda z$. Hence,

$$
\begin{equation*}
\mathrm{f}(\mathrm{z})=\phi_{-\beta}\left\{\lambda \phi_{\alpha}(\mathrm{z})\right\}, \quad \mathrm{z} \in \mathrm{D}(0,1) \tag{59}
\end{equation*}
$$

We now present an important consequence of Schwarz's lemma, which may be seen as the converse form of theorem 5.2.

Theorem 5.3 : Let $\mathrm{f}: \mathrm{D}(0,1) \rightarrow \mathrm{D}(0,1)$ be any conformal map of the unit disc onto itself and $\mathrm{f}(\mathrm{a})=0, \mathrm{a} \in \mathrm{D}(0,1)$. Then there is a constant $\lambda,|\lambda|=1$ such that $\mathrm{f}(\mathrm{z})=\lambda \phi_{\mathrm{a}}(\mathrm{z})$ where $\phi_{\mathrm{a}}$ is defined as in theorem 5.2.
Proof. Since f is a conformal map from $\mathrm{D}(0,1)$ to $\mathrm{D}(0,1)$, we can have inverse of $f, g$ defined by

$$
\mathrm{g}\{\mathrm{f}(\mathrm{z}))\}=\mathrm{z},
$$

which is analytic too. Applying the chain rule

$$
\begin{equation*}
g^{1}(0) f^{1}(a)=1 \tag{60}
\end{equation*}
$$

But according to inequality (58), $f$ and $g$ have to satisfy

$$
\begin{equation*}
\left|\mathrm{f}^{1}(\mathrm{a})\right| \leq \frac{1}{1-|\mathrm{a}|^{2}}, \quad\left|\mathrm{~g}^{1}(0)\right| \leq 1-|\mathrm{a}|^{2} \tag{61}
\end{equation*}
$$

(since, $f(a)=0$ and $g(0)=a)$.
From (60), (61) it follows that $\left|\mathrm{f}^{1}(\mathrm{a})\right|=\left(1-|\mathrm{a}|^{2}\right)^{-1}$. Hence applying the result (59) we find that

$$
\mathrm{f}(\mathrm{z})=\lambda \phi_{\mathrm{a}}(\mathrm{z})
$$

for some $\lambda$ with $|\lambda|=1$.
Lemma 5.1 : Let $G$ be a simply connected region and $\left\{f_{n}\right\}$ be a sequence of injective analytic mappings (conformal mappings) of $G$ into $\mathbb{C}$ which converges uniformly on every compact subset of G, then the limit function $f$ is either constant or injective.

Proof. Suppose f is not constant and not injective. Then there exist two points $\varsigma$ and $\eta \in G, \varsigma \neq \eta$ such that $f(\varsigma)=f(\eta)=\omega_{0}$, say.

Let $g_{n}(z)=f_{n}(z)-\omega_{0}$. We can find a positive $\delta, \delta<|\varsigma-\eta| / 2$ so that the discs $D(\varsigma, \delta)$ and $D(\eta, \delta)$ are included in $G$. Now $g(z)=f(z)-\omega_{0}$ never vanishes on the circles $|z-\varsigma|=\delta$ and $|z-\eta|=\delta$, where $g(z)=\lim _{n \rightarrow \infty} g_{n}(z)$. Applying Hurwitz's theorem, for large $n$, there exists $\varsigma_{n}$ lying inside the circle $|z-\varsigma|=\delta$ with $g_{n}\left(\varsigma_{n}\right)=$ 0 as $\mathrm{g}_{\mathrm{n}} \rightarrow \mathrm{g}$ uniformly in G . Similarly, for all large n , there is $\eta_{\mathrm{n}}$ within $|\mathrm{z}-\eta|=\delta$ with $\mathrm{g}_{\mathrm{n}}\left(\eta_{\mathrm{n}}\right)=0$. But by construction, $\mathrm{D}(\varsigma, \delta) \cap \mathrm{D}(\eta, \delta)=\phi$ and hence $\varsigma_{\mathrm{n}} \neq \eta_{\mathrm{n}}$. Thus
that is,

$$
\begin{gathered}
\mathrm{g}_{\mathrm{n}}\left(\varsigma_{\mathrm{n}}\right)=\mathrm{g}_{\mathrm{n}}\left(\eta_{\mathrm{n}}\right)=0, \varsigma_{\mathrm{n}} \neq \eta_{\mathrm{n}} \\
\mathrm{f}_{\mathrm{n}}\left(\varsigma_{\mathrm{n}}\right)=\mathrm{f}_{\mathrm{n}}\left(\eta_{\mathrm{n}}\right), \varsigma_{\mathrm{n}} \neq \eta_{\mathrm{n}}
\end{gathered}
$$

contradicting the injectivity of each $\mathrm{f}_{\mathrm{n}}$ and the proof follows.
NOTE : There is no conformal map f of the unit disc $\mathrm{D}(0,1)$ onto the whole complex plane $\mathbb{C}$ because then the inverse function $\mathrm{f}^{-1}: \mathbb{C} \rightarrow \mathrm{D}(0,1)$ would be a bounded entire function which is not constant, contradicting the Liouville's theorem.

Open mapping theorem : Let $G$ be a region and suppose that f is a non-constant analytic function on $G$. Then for any open set $U$ in $G, f(U)$ is open.

Proof : Omitted.
Uniform boundedness : A sequence of functions $\left\{\mathrm{f}_{\mathrm{n}}\right\}$ defined on a set D is said to be uniformly bounded on D if there exists a constant $\mathrm{M}>0$ such that $\left|\mathrm{f}_{\mathrm{n}}(\mathrm{z})\right| \leq \mathrm{M}$ for all n and for all $\mathrm{z} \in \mathrm{D}$.

Normal family : Let F be a family of functions in a region G . The family F is said to be normal in G if every sequence $\left\{f_{n}\right\}$ of functions $f_{n} \in F$ contains a subsequence $\left\{\mathrm{f}_{\mathrm{nk}}\right\}$ which converges uniformly on every compact subset of G .

Montel's theorem : A family F in $\mathrm{H}(\mathrm{G})$ is normal if and only if F is uniformly bounded on every compact subset of G.

Proof : Omitted.
Theorem 5.4 : [Riemann Mapping Theorem] Let G be a simply connected region, except for $\mathbb{C}$ itself and let $\mathrm{a} \in \mathrm{G}$. Then there is a unique conformal map $\mathrm{f}: \mathrm{G} \rightarrow \mathrm{D}$ $(0,1)$ of $G$ onto the unit disc which satisfies

$$
f(a)=0 \text { and } f^{1}(a)>0 .
$$

Proof. Let us first prove that f is unique. If there was another conformal map $\mathrm{g}: \mathrm{G} \rightarrow \mathrm{D}(0,1)$ with the given properties, then

$$
\mathrm{fog}^{-1}: \mathrm{D}(0,1) \rightarrow \mathrm{D}(0,1)
$$

would be a conformal map and also

$$
\left(\mathrm{fog}^{-1}\right)(0)=\mathrm{f}(\mathrm{a})=0
$$

Hence, applying Theorem 5.3, we find that there is a constant $\lambda$ with $|\lambda|=1$

$$
\left(\mathrm{fog}^{-1}\right)(\mathrm{z})=\lambda \mathrm{z}
$$

Deriving the derivative at the origin, we find

$$
\left(f^{-1}\right)^{-1}(0)=f^{\prime}\left(g^{-1}\right)(0)\left(g^{-1}\right)^{\prime}(0)=f^{\prime}(a) \frac{1}{g^{\prime}\left(g^{-1}(0)\right)}=\frac{f^{\prime}(a)}{g^{\prime}(a)}>0,
$$

from which it follows that $\lambda$ is positive. But also $|\lambda|=1$, so $\lambda=1$. Thus fog ${ }^{-1}$ is an identity map and $f=g$.

The proof of existence is divided into several stages.
Lemma 5.2 Let $G$ be a simply connected region other than $\mathbb{C}$. Then there exists an injective analytic map $f$ on $G$ with $f(G) \subset D(0,1)$.

Proof. We choose a point $b \in \mathbb{C} \backslash G$. Since $G$ is simply connected there exists a $\mathrm{g}: \mathrm{G} \rightarrow \mathbb{C}$ analytic with $\mathrm{g}^{2}(\mathrm{z})=\mathrm{z}-\mathrm{b}$.

Here $g$ is injective since

$$
\begin{array}{lc} 
& \\
& \mathrm{g}\left(\mathrm{z}_{1}\right)=\mathrm{g}\left(\mathrm{z}_{2}\right) \\
\Rightarrow & \mathrm{g}^{2}\left(\mathrm{z}_{1}\right)=\mathrm{g}^{2}\left(\mathrm{z}_{2}\right) \\
\text { i.e. } & \mathrm{z}_{1}-\mathrm{b}=\mathrm{z}_{2}-\mathrm{b} \\
\Rightarrow & \mathrm{z}_{1}=\mathrm{z}_{2} .
\end{array}
$$

By open mapping theorem $\mathrm{g}(\mathrm{G})$ is open. Let us pick $\omega_{0} \in \mathrm{~g}(\mathrm{G})$ and choose $\mathrm{r}>0$ so that $\mathrm{D}\left(\omega_{0}, \mathrm{r}\right) \subset \mathrm{g}(\mathrm{G})$. Then $\mathrm{D}\left(-\omega_{0}, \mathrm{r}\right) \subset \mathbb{C} \backslash \mathrm{g}(\mathrm{G})$. For, if there exists a point $\omega \in$ $D\left(-\omega_{0}, r\right) \cap g(G)$, then $\omega=g\left(z_{1}\right)$ for some $z_{1} \in G$ and also $-\omega \in D\left(\omega_{0}, r\right) \subset g(G)$, so that $-\omega=g\left(z_{2}\right)$ for some $z_{2} \in G$. Again,

$$
\begin{array}{ll} 
& g\left(z_{1}\right)=-g\left(z_{2}\right) \\
\Rightarrow & g^{2}\left(z_{1}\right)=g^{2}\left(z_{2}\right) \\
\text { or, } & z_{1}-b=z_{2}-b \\
\text { i.e. } & z_{1}=z_{2} \\
\text { or, } & g\left(z_{1}\right)=g\left(z_{2}\right)=-g\left(z_{1}\right) \\
\Rightarrow & g\left(z_{1}\right)=0 \\
\Rightarrow & 0=g^{2}\left(z_{1}\right)=z_{1}-b \\
\text { i.e. } & z_{1}=b \in \mathbb{C} \backslash G
\end{array}
$$

contradicting $\mathrm{z}_{1} \in \mathrm{G}$.

$$
\begin{equation*}
\text { We take } \quad \mathrm{f}(\mathrm{z})=\frac{\mathrm{r}}{2\left[\mathrm{~g}(\mathrm{z})+\omega_{0}\right]} \tag{62}
\end{equation*}
$$

Then $f$ is injective analytic map on $G$ (by construction $\left|g(z)+\omega_{0}\right| \geq r$ for $z \in G$ ) and also satisfies $|\mathrm{f}(\mathrm{z})| \leq \frac{1}{2}<1$ for $\mathrm{z} \in \mathrm{G}$.

Lemma 5.3 : Let G be a simply connected region other than $\mathbb{C}$ itself and let $\mathrm{a} \in \mathrm{G}$ be fixed. Then there exists a conformal map $\mathrm{f}: \mathrm{G} \rightarrow \mathrm{D}(0,1)$ of G onto the unit disc with the properties $f(z)=0$ and $f(a)>0$.

Proof : Let F denote the family of analytic functions $f: G \rightarrow \mathbb{C}$ such that either $\mathrm{f} \equiv 0$ or f is injective, and $\mathrm{f}(\mathrm{G}) \subset(0,1), \mathrm{f}(\mathrm{a})=0$ and $\mathrm{f}^{\prime}(\mathrm{a})>0$.

Let us consider the function

$$
\psi(z)=\frac{f(z)-f(a)}{1-\bar{f}(a)} f(z)
$$

where $f(z)$ is given by (62) of lemma 5.2 and we find that $\psi(G) \subset D(0,1), \psi(a)$ $=0$ and $\psi^{1}(\mathrm{a})>0$. So F is non empty and by Montel's theorem it is normal. Applying Lemma 1 we see that all functions in the closure of $F$ in $H(G)$ are either constant or injective. Now since all functions in F take the value zero at a, the same is true for all functions in the closure of F . Likewise the only constant function in the closure is

0 while the other functions in the closure satisfy $f(G) \subset \bar{D}(0,1)$. Since $f(G)$ is open, by open mapping theorem, $f(G) \subset D(0,1)$. Again since the $f \rightarrow f^{1}(a)$ is continuous, all functions in the closure of $F$ must satisfy $\mathrm{f}^{1}(\mathrm{a}) \geq 0$. The functions in the closure, that are not identically zero, are injective, so $\mathrm{f}^{1}(\mathrm{a})>0$ unless $\mathrm{f} \equiv 0$. These observations prove that the set F is closed in $\mathrm{H}(\mathrm{G})$. Hence $F$ is compact in $\mathrm{H}(\mathrm{G})$.

Since the map $f \rightarrow f^{\prime}(a): F \rightarrow R$ is a continuous function on a compact set, it must attain its maximum value, as we are not considering constant function (here it is zero). Let $f \in F$ be a function with $f^{\prime}(a)$ maximum.

We now show that $f(G)=D(0,1)$. On the contrary, suppose that $f(G) \neq D(0,1)$ and choose $\mathrm{w} \in \mathrm{D}(0,1) \backslash \mathrm{f}(\mathrm{G})$. Using the property that every non-vanishing analytic function in a simply connected region has an analytic square root, we take a function $h \in H(G)$ with

$$
\begin{equation*}
[\mathrm{h}(\mathrm{z})]^{2}=\frac{\mathrm{f}(\mathrm{z})-\omega}{1-\bar{\omega} \mathrm{f}(\mathrm{z})} \tag{63}
\end{equation*}
$$

Now as the bilinear transformation $\phi_{\mathrm{a}}(\mathrm{z})=\frac{\mathrm{z}-\mathrm{a}}{1-\overline{\mathrm{a}} \mathrm{z}}$ maps $\mathrm{D}(0,1)$ onto $\mathrm{D}(0,1)$ and as $f \in F, h(G) \subset D(0,1)$.

Let $\mathrm{g}: \mathrm{G} \rightarrow \mathbb{C}$ defined by

$$
\mathrm{g}(\mathrm{z})=\frac{\left|\mathrm{h}^{\prime}(\mathrm{a})\right|}{\mathrm{h}^{\prime}(\mathrm{a})} \cdot \frac{\mathrm{h}(\mathrm{z})-\mathrm{h}(\mathrm{a})}{1-\overline{\mathrm{h}(\mathrm{a})} \mathrm{h}(\mathrm{z})}
$$

Then clearly, $\mathrm{g}(\mathrm{G}) \subset \mathrm{D}(0,1), \mathrm{g}(\mathrm{a})=0$ and g is analytic injective and $\mathrm{g}^{\prime}(\mathrm{a})>$ 0 , since

$$
\begin{align*}
\mathrm{g}^{\prime}(\mathrm{a}) & =\frac{\left|\mathrm{h}^{1}(\mathrm{a})\right|}{\mathrm{h}^{1}(\mathrm{a})} \cdot \frac{\mathrm{h}^{1}(\mathrm{a})\left[1-|\mathrm{h}(\mathrm{a})|^{2}\right]}{\left[1-|\mathrm{h}(\mathrm{a})|^{2}\right]^{2}} \\
& =\frac{\left|\mathrm{h}^{1}(\mathrm{a})\right|}{1-|\mathrm{h}(\mathrm{a})|^{2}}>0 \tag{64}
\end{align*}
$$

So, $g \in F$.
Again, differentialing (63) we find that

$$
2 h(a) h^{1}(a)=f^{1}(a)\left(1-|\omega|^{2}\right)
$$

So, from (64)

$$
g^{1}(a)=\frac{|h(a)| h^{1}(a) \mid}{|h(a)|\left(1-|h(a)|^{2}\right.}=\frac{f^{1}(a)\left(1-|\omega|^{2}\right.}{2 \sqrt{\omega}(1-|\omega|)} \text {, as }|h(a)|^{2}=|\omega|
$$

$$
=\frac{\mathrm{f}^{1}(\mathrm{a})(1+|\omega|)}{2 \sqrt{\omega}}>\mathrm{f}^{1}(\mathrm{a}) .
$$

contradicting the choice of $f \in F$ as maximising $f^{1}(a)$. Thus $f(G)=D(0,1)$.
Note : The Riemann mapping theorem is one of the most celebrated results of complex analysis. It is the beginning of the study of complex analysis from a geometric view point. G. F. B. Riemann in 1851 correctly formulated the theorem, but unfortunately his proof of the theorem was lacking. According to various accounts, he assumed but did not prove that a certain maximal problem had a solution. A final proof was definitely known by the early 20th century, different sources attributed to it particularly, W. F. Osgood, P. Koebe, L Bieberbach etc.

### 5.2 The Schwarz Reflection Principle

Let f be analytic in the domains $\mathrm{D}_{1}, \mathrm{D}_{2}$ which have a common piece of boundary, a smooth curve $\gamma$. Assume further that f is continuous across $\gamma$. Then, by Morera's theorem, f is analytic in $\mathrm{D}_{1} \cup \mathrm{D}_{2}$. This allows us to perform analytic continuation in some cases.

Theorem 5.5 [The Schwarz reflection principle] Given a function $f(z)$ analytic in a domain D lying in the upper half plane whose boundary contains a segment $\mathrm{I} \subset \mathrm{IR}$, assume $f$ is continuous on $D \cup I$ and real-valued on $I$. Then $f$ has analytic continuation across $I$, in a domain $D \cup I U D^{*}$, where $D^{*}=\{\bar{z}: z \in D\}$.

Proof. Let us consider the function

$$
f(z)= \begin{cases}\frac{f(z)}{f(\bar{z})}, & z \in D \in D * U I\end{cases}
$$

It is clear that F is analytic in D . We shall show that F is also analytic in $\mathrm{D}^{*}$.
Let z and $\mathrm{z}+\mathrm{h}$ lie within $\mathrm{D}^{*}$. Then $\overline{\mathrm{z}}$ and $\overline{\mathrm{z}}+\overline{\mathrm{h}}$ lie within D and we can express.

$$
\lim _{h \rightarrow 0} \frac{F(z+h)-F(z)}{h}=\lim _{h \rightarrow 0} \frac{\overline{f(\bar{z}+\bar{h})}-\overline{f(\bar{z})}}{h}=\lim _{h \rightarrow 0}\left[\frac{\overline{f(\bar{z}+\bar{h})-f(\bar{z})}}{h}\right]=\overline{f^{\prime}(\bar{z})} .
$$

So, F is analytic in $\mathrm{D}^{*}$. F is also continuous on $\mathrm{D}^{*} \mathrm{U}$ I.
For, $\mathrm{z} \in \mathrm{I}$

$$
\lim _{z \rightarrow x} F(z)=\lim _{z \rightarrow x} \overline{f(\bar{z})}=\overline{f(x)}=f(x),
$$

by hypothesis. Thus F is continuous on D U I U D*. To prove F is also analytic there, we consider the function

$\phi(\mathrm{z})=\frac{1}{2 \pi \mathrm{i}} \int_{\Gamma} \frac{\mathrm{F}(\varsigma)}{\varsigma-\mathrm{z}} \mathrm{d} \varsigma$
It is analytic in D U I U D* [as (i) $\frac{F(\varsigma)}{\varsigma-z}$ is continuous function of both variables when z lies within $\Gamma$ and $\varsigma$ on $\Gamma$.
(ii) for each such $\varsigma, \frac{F(\varsigma)}{\varsigma-z}$ is analytic in z in D U I U D*. [see (14)].
To complete the proof, we try to establish $\phi(z)=F(z)$ for all $z \varepsilon D U I D^{*}$.
Breaking the integral in (65) and adding the two integrals along I, which are in opposite directions, we write

$$
\begin{equation*}
\phi(\mathrm{z})=\frac{1}{2 \pi \mathrm{i}} \int_{\Gamma_{1}} \frac{\mathrm{~F}(\varsigma)}{\varsigma-\mathrm{z}} \mathrm{~d} \rho+\frac{1}{2 \pi \mathrm{i}} \int_{\Gamma_{2}} \frac{F(\varsigma)}{\zeta-\mathrm{z}} \mathrm{~d} \rho \tag{66}
\end{equation*}
$$

where $\Gamma_{1}$ and $\Gamma_{2}$ are the boundary of D U I and D* U I respectively. When zeDUI, the second integral in (66) vanishes and $\phi(\mathrm{z})=\mathrm{F}(\mathrm{z})$. Again, the first integral vanishes when $z \varepsilon D^{*} U$ I and $\phi(z)=F(z)$ in this case too. Thus $\phi(z)=F(z)$ for all $z \in D U$ I U D* and we have found a function F(z), analytic in D U I U D*, and coincides with $f(z)$ in D U I.

### 5.3 The Schwarz-Christoffel Transformation

We know from Riemann's mapping theorem that there is a conformal mapping which maps a given simply connected domain onto another simply connected domain, or equivalently onto the unit disc. But it does not help us to determine such mappings.

Many applications in boundary-value problem requires construction of one-to-one conformal mapping from the upper half plane $\operatorname{Im} \mathrm{z}>0$ onto a polygon $\Omega$ in the w-plane. Two German mathematicians H. A. Schwarz and E. B. Christoffel independently discovered a method for finding such mappings during the years 1864-1869.

Theorem 5.6 [Schwarz and Christoffel] Let $P$ be a polygon with vertices $w_{1}$, $\ldots \mathrm{w}_{\mathrm{k}}$ in the anticlockwise direction and interior angles $\alpha_{1} \pi, \ldots, \alpha_{k} \pi$ respectively, where $-1<\alpha_{1}, \ldots, \alpha_{k}<1$. Then there exists a one-to-one conformal mapping of the form

$$
\begin{equation*}
f(z)=A \int_{z_{0}}^{z}\left(s-x_{1}\right)^{\alpha_{1}-1}\left(s-x_{2}\right)^{\alpha_{2}-1} \ldots\left(s-x_{k-1}\right) \alpha_{k-1}-1 d s+B \tag{67}
\end{equation*}
$$

where $A, B \in \mathbb{C}$, that maps the upper plane $\operatorname{Im} z>0$ onto the interior of $P$, with

$$
\begin{equation*}
f\left(x_{1}\right)=w_{1} \ldots \ldots \ldots, f\left(x_{k-1}\right)=w_{k-1}, f(\infty)=w_{k} . \tag{68}
\end{equation*}
$$

Remarks: (i) We do not need to have specific information on $W_{k}$ and $\alpha_{k}$. While travelling the polygon anticlockwise direction we made a left turn of an angle $\pi-\alpha_{j} \pi$ at the vertex $\omega_{\mathrm{j}}$.
(ii) Sometimes certain infinite regions can be thought of as infinite polygons. In this case it is convenient to take $\mathrm{w}_{\mathrm{k}}$ as the point at infinity, as we need no information on $\alpha_{k}$.
(iii) It can be shown that Schwarz-Christoffel transformation can be uniquely determined by three points as in the case of bilinear transformation. One of these is used by taking $\mathrm{f}(\infty)=\omega_{\mathrm{k}}$. We can therefore have the freedom to choose two points say, $x_{1}$ and $x_{2}$ satisfying $-\infty<x_{1}<x_{2}<\infty$.
(iv) Note that the integral involved may be impossible to calculate theoretically. In practical problems numerical techniques are often used to evaluate the integral. In first part of the proof we take $f\left(x_{k}\right)=\omega_{k}, x_{k}=$ finite.

Proof. By Riemann mapping theorem such a mapping exists. We shall prove that its form is given by (67). So $f(z)$ is analytic for $\operatorname{Im} z>0$ and $f^{1}(z) \neq 0$ in the upper half plane. From these it is clear that

$$
\frac{\mathrm{d}}{\mathrm{dz}} \log \mathrm{f}^{\prime}(\mathrm{z})=\frac{\mathrm{f}^{\prime \prime}(\mathrm{z})}{\mathrm{f}^{\prime}(\mathrm{z})}
$$

is analytic in the upper half plane. To construct the function $f(z)$ our aim is to


Fig. 55 establish that $f^{\prime \prime}(z) / f^{\prime}(z)$ is analytic for $\operatorname{Im} z \geq 0$ save for the pre-image points of the vertices of the polygon lying on the real axis.

Let $l$ be a side of the polygon P , which makes an angle $\theta$ (positive sense) with the realaxis and $\varsigma$ be any point on $l$ but not a vertex of the polygon P . Then for any $\omega$ on $l,(\omega-\varsigma) \mathrm{e}^{-\mathrm{i} \theta}$ is real and there is a point $z$ on the real axis of the $z$-plane so that $f(z)=\omega$ and a corresponding point $\mathrm{z}=\mathrm{a}$ for $\varsigma$ on the same line. Hence

$$
\{\mathrm{f}(\mathrm{z})-\varsigma\} \mathrm{e}^{-\mathrm{i} \theta}
$$

is real and continuous on the segment $\gamma$ of the real axis of the z -plane corresponding to the straight line $l$ of the $\omega$-plane. Moreover, this function is also analytic for $\operatorname{Im} \mathrm{z}>0$, thus following the Schwarz reflection principle we can continue this function analytically across $\gamma$ to the lower half plane $\operatorname{Im} \mathrm{z}<0$. In particular, this function is analytic in a neighbourhood of the point $\mathrm{z}=\mathrm{a}$ and can be expanded in the form of the Taylor series.

$$
\{\mathrm{f}(\mathrm{z})-\varsigma\} \mathrm{e}^{-\mathrm{i} \theta}=\sum_{\mathrm{k}=1}^{\infty} \mathrm{c}_{\mathrm{k}}(\mathrm{z}-\mathrm{a})^{\mathrm{k}}
$$

where $c_{1}=f^{\prime}(a) \neq 0$, maintaining the status quo that $f(a)=\varsigma$ and the function $f$ maps the segment $\gamma$ onto the straight line $l$. Now

$$
\begin{gathered}
f^{\prime}(z)=e^{i \theta}\left\{c_{1}+c_{2} 2(z-a)+\ldots\right\} \\
\log f^{\prime}(z)=i \theta+\log \left\{c_{1}+2 c_{2}(z-a)+\ldots\right\}
\end{gathered}
$$

and
So, $\frac{d}{d z} \log f^{1}(z)$ is analytic in a neighbourhood of $z=a$ and real on a real line segment intercepted by the neighbourhood.

Let us consider the case when the point $\varsigma$ is the corresponding point at infinity on $\gamma$ (in this case $\gamma$ is divided into two parts, each of infinite lenght). Here the Taylor series expansion in the neighbourhood of point at infinity

$$
\{\mathrm{f}(\mathrm{z})-\varsigma\} \mathrm{e}^{-\mathrm{i} \theta}=\sum_{\mathrm{k}=1}^{\infty} \mathrm{c}_{\mathrm{k}} / \mathrm{z}^{\mathrm{k}}
$$

where each $c_{R}$ is real and $c_{1} \neq 0$ (with the same reason mentioned in the finite case). So

$$
\begin{gathered}
\mathrm{f}^{\prime}(\mathrm{z}) \mathrm{e}^{-\mathrm{i} \theta}=-\frac{\mathrm{c}_{1}}{\mathrm{z}^{2}}-\frac{2 \mathrm{c}_{2}}{\mathrm{z}^{3}}-\frac{3 \mathrm{c}_{3}}{\mathrm{z}^{4}}-\ldots \\
\mathrm{f}^{\prime \prime}(\mathrm{z}) \mathrm{e}^{-\mathrm{i} \theta}=\frac{2 \mathrm{c}_{1}}{\mathrm{z}^{3}}+\frac{6 \mathrm{c}_{2}}{\mathrm{z}^{4}}+\frac{12 \mathrm{c}_{3}}{\mathrm{z}^{5}}+\ldots .
\end{gathered}
$$

and we find that

$$
\begin{align*}
\frac{\mathrm{f}^{\prime \prime}(\mathrm{z})}{\mathrm{f}^{\prime}(\mathrm{z})}=\frac{\mathrm{z}^{-3}\left\{2 \mathrm{c}_{1}+\frac{6 \mathrm{c}_{2}}{\mathrm{z}}+\frac{12 \mathrm{c}_{3}}{\mathrm{z}^{2}}+\ldots\right\}}{-\mathrm{c}_{1} \mathrm{z}^{-2}\left\{1+\frac{2 \mathrm{c}_{2} / \mathrm{c}_{1}}{\mathrm{z}}+\ldots\right\}} & =-\frac{1}{\mathrm{c}_{1}}\left\{2 \mathrm{c}_{1}+\frac{6 \mathrm{c}_{2}}{\mathrm{z}}+\ldots\right\}\left\{1-\frac{2 \mathrm{c}_{2} / \mathrm{c}_{1}}{\mathrm{z}}+\ldots\right\} \\
& =-\frac{2}{\mathrm{z}}+\sum_{\mathrm{k}=2}^{\infty} \frac{\tilde{\mathrm{c}}_{\mathrm{k}}}{\mathrm{z}^{k}} \tag{69}
\end{align*}
$$

$\frac{d}{d z} \log f^{1}(z)$ is analytic in a neighbourhood of the point at infinity and is real when z is real.

In the polygon P , let $\ell^{1}$ be an adjacent side to $\ell$ making on angle $\alpha_{1} \pi$ at their point of intersection $\omega_{1}$. The corresponding point of $\omega_{1}$ on the real axis is $x_{1}$. Here
the function $f(z)$ is not analytic in a neighbourhood of $x_{1}$, we choose the branch of the argument so that

$$
\frac{\pi}{2}<\operatorname{Arg}\left(z-x_{1}\right)<\frac{3 \pi}{2}
$$

introducing a branch cut along the axis $\left\{\mathrm{x}_{1}+\mathrm{iy}: \mathrm{y} \leq 0\right\}\left[\mathrm{f}^{\prime}(\mathrm{z})\right.$ is not continuous on this branch cut].


Fig. 56


Fig. 57

Here $\operatorname{Arg}\left\{\left(\omega_{1}-\omega\right) \mathrm{e}^{-\mathrm{i} \theta}\right\}$ is equal to zero or $\alpha_{1} \pi$ according as $\omega$ lies on $\ell$ or $\ell^{1}$. So the function

$$
\left[\left\{\omega_{1}-\mathrm{f}(\mathrm{z}) \mathrm{e} \mathrm{e}^{-\mathrm{i} \theta}\right]^{1 / \alpha_{1}}\right.
$$

is real and continuous on the segment of the real axis corresponding to the consecutive sides $\ell$ and $\ell^{1}$. Again this function is analytic for $\operatorname{Im} \mathrm{z}>0$ since $\mathrm{f}(\mathrm{z})-\omega_{1}$ is analytic and non zero there.

Expanding $\left[\left\{\omega_{1}-f(z)\right\} e^{-i \theta}\right]^{1 / \alpha_{1}}$ in Taylor's series in a neighbourhood of $x_{1}$ we find

$$
\left[\left\{\omega_{1}-f(z)\right\} e^{-i \theta}\right]^{1 / \alpha_{1}}=\sum_{k=1}^{\infty} c_{k}\left(z-x_{1}\right)^{k}
$$

where each $c_{k}$ is real and $c_{1} \neq 0$. On simplifying, we find

$$
\begin{aligned}
f(z) & =\omega_{1}-e^{i \theta}\left(z-x_{1}\right)^{\alpha_{1}}\left[c_{1}+c_{2}\left(z-x_{1}\right)+\ldots\right]^{\alpha_{1}} \\
& =\omega_{1}+e^{i \theta}\left(z-x_{1}\right)^{\alpha_{1}} \sum_{k=0}^{\infty} c_{k}^{1}\left(z-x_{1}\right)^{k}
\end{aligned}
$$

where $c_{0}{ }^{1}$ is a constant multiple of $c_{1}$, hence not equal to zero. Now we have

$$
\begin{aligned}
\mathrm{f}^{\prime}(\mathrm{z}) & =\mathrm{e}^{\mathrm{i} \mathrm{\theta}}\left(\mathrm{z}-\mathrm{x}_{1}\right)^{\alpha_{1}-1}\left[\alpha_{1} \mathrm{c}_{0}{ }^{1}+\left(\alpha_{1}+1\right) \mathrm{c}_{1}{ }^{1}\left(\mathrm{z}-\mathrm{x}_{1}\right)+\ldots\right] \\
& =\left(\mathrm{z}-\mathrm{x}_{1}\right)^{\alpha_{1}-1} \mathrm{~F}(\mathrm{z})
\end{aligned}
$$

where $F(z)$ is analytic and not zero in a neighbourhood of $z=x_{1}$ and we obtain

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{dz}} \log \mathrm{f}^{1}(\mathrm{z})=\frac{\alpha_{1}-1}{\mathrm{z}-\mathrm{x}_{1}}+\frac{\mathrm{F}^{1}(\mathrm{z})}{\mathrm{F}(\mathrm{z})} \tag{70}
\end{equation*}
$$

This shows that if the polygon $P$ has an angle $\alpha_{1} \pi$ at a point $\omega_{1}$ then $\frac{d}{d z} \log f^{1}(z)$ will have a simple pole of residue $\alpha_{1}-1$ at its corresponding point $\mathrm{x}_{1}$.

Now if the point at infinity be the corresponding point to $\omega_{1}$ at which the polygon $P$ has an angle $\alpha_{1} \pi$, then we can express

$$
\begin{align*}
& {\left[\left\{\omega_{1}-\mathrm{f}(\mathrm{z})\right\} \mathrm{e}^{-\mathrm{i} \theta}\right]^{1 / \alpha_{1}}=\frac{\mathrm{c}_{1}}{\mathrm{z}}+\frac{\mathrm{c}_{2}}{\mathrm{z}^{2}}+\ldots .} \\
& \text { or, } \quad f(z)=\omega_{1}-e^{i \theta( }\left(\frac{c_{1}}{z}\right)^{\alpha_{1}}\left(1+\alpha_{1} \frac{c_{2}}{z_{1}}+\ldots\right) \\
& f^{\prime}(z)=+e^{i \theta} \alpha_{1} \frac{c_{1}^{\alpha_{1}}}{z^{\alpha_{1}+1}}\left(1+\alpha_{1} \frac{c_{2}}{z c_{1}}+\ldots\right)-e^{i \theta}\left(\frac{c_{1}}{z}\right)^{\alpha_{1}}\left(-\frac{\alpha_{1} c_{2}}{z^{2} c_{1}}-\ldots\right) \\
& =\mathrm{e}^{\mathrm{i} \theta} \mathrm{c}_{1}^{\alpha_{1}} \frac{\alpha_{1}}{\mathrm{z}^{\alpha_{1}+1}}\left[1+\left(\alpha_{1}+1\right) \frac{\mathrm{c}_{2}}{\mathrm{zc}_{1}}+\ldots\right] \\
& \mathrm{f}^{\prime \prime}(\mathrm{z})=-\mathrm{e}^{\mathrm{i} \theta} \mathrm{c}_{1}{ }_{1}^{\alpha_{1}} \frac{\alpha_{1}\left(\alpha_{1}+1\right)}{\mathrm{z}^{\alpha_{1}+2}}\left\{1+\left(\alpha_{1}+1\right) \frac{\mathrm{c}_{2}}{\mathrm{zc}_{1}}+\ldots\right\}+\mathrm{e}^{\mathrm{i} \theta} \mathrm{c}_{1}{ }^{\alpha} \frac{\alpha_{1}}{\mathrm{z}^{\alpha_{1}+1}}\left\{-\left(\alpha_{1}+1\right) \frac{\mathrm{c}_{2}}{\mathrm{z}^{2} \mathrm{c}_{1}}-\ldots\right\} \\
& =-\mathrm{e}^{\mathrm{i} \theta} \mathrm{c}_{1}{ }_{1}^{\alpha_{1}} \frac{\alpha_{1}\left(\alpha_{1}+1\right)}{\mathrm{z}^{\alpha_{1}+2}}\left[1+\left(\alpha_{1}+2\right) \frac{\mathrm{c}_{2}}{\mathrm{zc}_{1}}+\ldots\right] \\
& \frac{d}{d z} \log f^{\prime}(z)=\frac{f^{\prime \prime}(z)}{f^{\prime}(z)}=-\frac{\alpha_{1}+1}{z}\left\{1+\left(\alpha_{1}+2\right) \frac{c_{2}}{{z c_{1}}^{\prime}}+\ldots\right\}\left\{1-\left(\alpha_{1}+1\right) \frac{c_{2}}{{z c_{1}}^{\prime}}+\ldots\right\} \\
& =-\frac{\alpha_{1}+1}{\mathrm{z}}\left\{1+\left(\alpha_{1}+2-\alpha_{1}-1\right) \frac{\mathrm{c}_{2}}{\mathrm{zc}_{1}}+\ldots .\right\} \\
& =-\frac{\alpha_{1}+1}{z}+\sum_{k=2}^{\infty} \frac{\tilde{c}_{k}}{z^{k}} \tag{71}
\end{align*}
$$

Now since $\mathrm{x}_{2}, \mathrm{x}_{3} \ldots, \mathrm{x}_{\mathrm{k}}$ are the corresponding points lying on the real-axis of the z-plane, to the vertices $w_{2}, w_{3}, \ldots w_{k}$ respectively of the polygon $P$ with angles $\alpha_{2} \pi$,
$\alpha_{3} \pi, \ldots \alpha_{k} \pi$ there, the function $\frac{d}{d z} \log f^{1}(z)$ will have simple poles with residue $\alpha_{j}-$ 1 at $x_{j}, j=2, \ldots, k$. Thus we see that this function is analytic for $\operatorname{Im} z>0$ and continuous on $\operatorname{Im} \mathrm{z}=0$ except the points $\mathrm{x}_{1}, \mathrm{x}_{2}, \ldots, \mathrm{x}_{\mathrm{k}}$ and using the Schwarz reflection principle it can be continued analytically across the real axis. Hence $\frac{d}{d z} \log f^{1}(z)$ possesses only simple poles at $\mathrm{x}_{1}, \mathrm{x}_{2}, \ldots \mathrm{x}_{\mathrm{k}}$ as its only singularities and can be expressed as

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{dz}} \log \mathrm{f}^{1}(\mathrm{z})=\frac{\alpha_{1}-1}{\mathrm{z}-\mathrm{x}_{1}}+\frac{\alpha_{2}-1}{\mathrm{z}-\mathrm{x}_{2}}+\ldots+\frac{\alpha_{\mathrm{k}}-1}{\mathrm{z}-\mathrm{x}_{\mathrm{k}}}+\mathrm{G}(\mathrm{z}) \tag{72}
\end{equation*}
$$

where $G(z)$ is a polynomial.
When $|\mathrm{z}|$ is large enough

$$
\begin{gather*}
\frac{\alpha_{i}-1}{z-x_{i}}=\frac{\alpha_{i}-1}{z}\left(1+\frac{x_{i}}{z}+\frac{x_{i}{ }^{2}}{z^{2}}+\ldots\right), i=1, \ldots, k \\
\text { So, } \frac{d}{d z} \log f^{1}(z)=\sum_{i=1}^{k}\left(\alpha_{i}-1\right) / z+\sum_{i=1}^{k} x_{1}\left(\alpha_{i}-1\right) / z^{2}+\sum_{i=1}^{k} x_{1}^{2}\left(\alpha_{1}-1\right) / z^{3}+\ldots+G(z) \\
=  \tag{73}\\
=-\frac{2}{z}+\sum_{i=2}^{\infty} \frac{d_{i}}{z_{i}}+G(z)
\end{gather*}
$$

Using the property of the sum of the exterior angles of a polygon, $\left(1-\alpha_{1}\right) \pi+$ $\left(1-\alpha_{2}\right) \pi+\ldots\left(1-\alpha_{k}\right) \pi=2 \pi$. Comparing (73) with (69) we get $\mathrm{G}(\mathrm{z})$ identically zero.

Finally integrating equation (72), we find the desired mapping $f(z)$ as

$$
\begin{equation*}
f(z)=A \int_{z 0}^{z}\left(s-x_{1}\right)^{\alpha_{1}-1}\left(s-x_{2}\right)^{\alpha_{2}-1} \ldots\left(s-x_{k}\right)^{\alpha_{k-1}-1} d s+B \tag{74}
\end{equation*}
$$

## Role of constants $A$ and $B$

(i) $|\mathrm{A}|$ controls the size of the polygon
(ii) $\operatorname{Arg} \mathrm{A}$ and B help to select the position, if any, in determining orientation and translation respectively.

## An useful observation

In some occasions we urge to make the evaluation process of the integral in (74) simple. For this sake, we consider the point at infinity corresponds to the vertex $w_{k}$ where the polygon $P$ has an angle $\alpha_{k} \pi$. Then we can express [see eq. (71)]

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{dz}} \log \mathrm{f}^{1}(\mathrm{z})=\frac{\alpha_{\mathrm{k}}-1}{\mathrm{z}}+\sum_{2}^{\infty} \frac{\tilde{c}_{i}}{z^{i}} \tag{75}
\end{equation*}
$$

in the neighbourhood of the point at infinity.
Again considering the expression of $\frac{d}{d z} \log f^{1}(z)$ in the neighbourhood of the points corresponding to the vertices $\mathrm{w}_{1}, \mathrm{w}_{2} \ldots, \mathrm{w}_{\mathrm{k}-1}$ [see eq. (70)].

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{dz}} \log \mathrm{f}^{1}(\mathrm{z})=\frac{\alpha_{1}-1}{\mathrm{z}-\mathrm{x}_{1}}+\frac{\alpha_{2}-1}{\mathrm{z}-\mathrm{x}_{2}}+\ldots+\frac{\alpha_{\mathrm{k}-1}-1}{\mathrm{z}-\mathrm{x}_{\mathrm{k}-1}}+\mathrm{G}(\mathrm{z}) \tag{1}
\end{equation*}
$$

where $G(z)$ is a polynomial. If $|z|$ is large enough, proceeding as earlier

$$
\begin{align*}
\frac{\mathrm{d}}{\mathrm{dz}} \log \mathrm{f}^{1}(\mathrm{z}) & =\sum_{1}^{\mathrm{k}-1}\left(\alpha_{\mathrm{i}}-1\right) / \mathrm{z}+\sum_{i}^{k-1} x_{i}\left(\alpha_{i}-1\right) / \mathrm{z}^{2}+\sum_{1}^{k-1} x_{i}{ }^{2}\left(\alpha_{i}-1\right) / z^{3}+G(z) \\
& =-\frac{\alpha_{\mathrm{k}}+1}{z}+\sum_{2}^{\infty} \frac{\tilde{\mathrm{d}}_{\mathrm{i}}}{\mathrm{z}^{1}}+G(\mathrm{z}) \tag{76}
\end{align*}
$$

Comparing (76) with (75), $G(z)$ turns out to be identically zero and hence integrating (751) we obtain

$$
f(z)=A \int_{z_{0}}^{z}\left(s-x_{1}\right)^{\alpha_{1}-1}\left(s-x_{2}\right)^{\alpha_{2}-1} \ldots\left(s-x_{k-1}\right)^{\alpha_{k-1}-1} d s+B
$$

where the role of the constants A and B remain as before.

### 5.4 Examples : Triangles / Rectangles

The Schwarz-Christoffel transformation is expressed in terms of the points $\mathrm{x}_{\mathrm{j}}$, not in terms of their images i.e., the vertices of the polygon. Not more than three points $\left(\mathrm{x}_{\mathrm{j}}\right)$ can be chosen arbitrarily. If the point at infinity be one of the $\mathrm{x}_{\mathrm{j}}$ 's then only two finite points on the real-axis are free to be chosen, whether the polygon is a triangle or a rectangle etc.

## Triangle

Let the polygon be a triangle with vertices $\mathrm{w}_{1}, \mathrm{w}_{2}$ and $\mathrm{w}_{3}$. The S-C transformation is written as

$$
\begin{equation*}
\mathrm{w}=\mathrm{A} \int_{\mathrm{z}_{0}}^{\mathrm{z}}\left(\mathrm{~s}-\mathrm{x}_{1}\right)^{\alpha_{1}-1}\left(\mathrm{~s}-\mathrm{x}_{2}\right)^{\alpha_{2}-1}\left(\mathrm{~s}-\mathrm{x}_{3}\right)^{\alpha_{3}-1} \mathrm{ds}+\mathrm{B} \tag{77}
\end{equation*}
$$

where $\alpha_{1}, \pi, \alpha_{2} \pi$ and $\alpha_{3} \pi$ are the internal angles at the respective vertices.


Fig. 58


Fig. 59

Here we have chosen all the three finite points $\mathrm{x}_{1}, \mathrm{x}_{2}, \mathrm{x}_{3}$ on the real-axis.
The constants A, B control the size and position of the triangle respectively.
If we take the vertex $w_{3}$ as the image of the point at infinity, the $\mathrm{S}-\mathrm{C}$ transformation becomes

$$
\begin{equation*}
\mathrm{w}=\mathrm{A} \int_{\mathrm{z}_{0}}^{z}\left(\mathrm{~s}-\mathrm{x}_{1}\right)^{\alpha_{1}-1}\left(\mathrm{~s}-\mathrm{x}_{2}\right)^{\alpha_{2}-1} \mathrm{ds}+\mathrm{B} \tag{78}
\end{equation*}
$$

Here $\mathrm{x}_{1}$ and $\mathrm{x}_{2}$ can be chosen arbitrarily.
Example 1 : Find a Schwarz-Christoffel transformation that maps the upper halfplane to the inside of the triangle with vertices $-1,1$ and $\sqrt{ } 3 \mathrm{i}$.

## Solution :



Fig. 60


Fig. 61

Following our notation, we write $\mathrm{w}_{1}=-1, \mathrm{w}_{2}=1$ and $\mathrm{w}_{3}=\sqrt{3 i}$ so that $\alpha_{1}=$ $\alpha_{2}=\alpha_{3}=1 / 3$. We choose the form (78) of S-C transformation and consider the mapping.

$$
f(z)=A \int_{0}^{z}\left(s-x_{1}\right)^{-2 / 3}\left(s-x_{2}\right)^{-2 / 3} d s+B, \quad[\text { here } f(\infty)=\sqrt{ } 3 i]
$$

We may choose $x_{1}=-1$ and $x_{2}=1$, so that $f(-1)=-1$ and $f(1)=1$. Therefore

$$
\begin{aligned}
f(z) & =A \int_{0}^{z}(s+1)^{-2 / 3}(s-1)^{-2 / 3} d s+B \\
& =A \int_{0}^{z}\left(s^{2}-1\right)^{-2 / 3} d s+B
\end{aligned}
$$

It then follows that

$$
=\mathrm{A} \int_{0}^{-1}\left(\mathrm{~s}^{2}-1\right)^{-2 / 3} \mathrm{ds}+\mathrm{B}=-1, \mathrm{~A} \int_{0}^{1}\left(\mathrm{~s}^{2}-1\right)^{-2 / 3} \mathrm{ds}+\mathrm{B}=1 .
$$

Rewriting these as

$$
-\mathrm{AL}+\mathrm{B}=-1 \text { and } \mathrm{AL}+\mathrm{B}=1 \text {, where } \mathrm{L}=\int_{0}^{1}\left(\mathrm{~s}^{2}-1\right)^{-2 / 3} \mathrm{ds}
$$

We obtain $\quad \mathrm{A}=\frac{1}{\int_{0}^{1}\left(\mathrm{~s}^{2}-1\right)^{-2 / 3} \mathrm{ds}}$ and $\mathrm{B}=0$. Hence

$$
\mathrm{f}(\mathrm{z})=\frac{1}{\int_{0}^{1}\left(\mathrm{~s}^{2}-1\right)^{-2 / 3} \mathrm{ds}} \int_{0}^{z}\left(\mathrm{~s}^{2}-1\right)^{-2 / 3} \mathrm{ds} .
$$

Example 2 : Using Schwarz-Christoffel transformation map the upper half-plane onto an equilateral triangle of side 5 units.

## Solution :



Fig. 62


Fig. 63

It is convenient to choose three arbitrary points $x_{1}=-1, x_{2}=1$ and $x_{3}=\infty$ which are mapped into the vertices of the equilateral triangle, so we take S-C transformation (78).

$$
\mathrm{f}(\mathrm{z})=\mathrm{A} \int_{1}^{\mathrm{z}}(\mathrm{~s}+1)^{-2 / 3}(\mathrm{~s}-1)^{-2 / 3} \mathrm{ds}
$$

Here, $f(-1)=w_{1}=0$ and $f(1)=w_{2}=5$. So that

$$
\mathrm{A}=5 / \int_{-1}^{1}\left(\mathrm{~s}^{2}-1\right)^{-2 / 3} \mathrm{ds}
$$

Hence the desired transformation is

$$
f(z)=\frac{5 \int_{1}^{z}\left(s^{2}-1\right)^{2 / 3} d s}{\int_{-1}^{1}\left(s^{2}-1\right)^{2 / 3} d s}
$$

Alternative : We take $\mathrm{z}_{0}=-1, \mathrm{~A}=1, \mathrm{~B}=0$ and find $\mathrm{S}-\mathrm{C}$ transformation as, (choosing one of $\mathrm{x}_{\mathrm{i}}$ 's as point at infinity)

$$
\begin{equation*}
\mathrm{w}=\int_{1}^{2}(\mathrm{~s}+1)(\mathrm{s}-1)^{2 / 3} \mathrm{ds} \tag{79}
\end{equation*}
$$

taking $\mathrm{x}_{1}=-1$ and $\mathrm{x}_{2}=1$.
Then $\tilde{f}(1)=\tilde{\mathrm{w}}_{2}$, say, and the image of the point $\mathrm{z}=-1$ is the point $\tilde{\mathrm{w}}_{1}=0$. When $\mathrm{z}=1$ in the integral we can write $\mathrm{s}=\mathrm{x}$, where $-1<\mathrm{x}<1$. Then $\mathrm{x}+1>0$ and $\operatorname{Arg}$ $(x+1)=0$, while $|x-1|=1-x$ and $\operatorname{Arg}(x-1)=\pi$. Hence

$$
\tilde{w}_{2}=\int_{-1}^{1}(x+1)^{2 / 3}(1-x)^{2 / 3} e^{-\mathrm{i} 2 \pi / 3} d x
$$

$$
\begin{aligned}
& =-\mathrm{e}^{\mathrm{i} \pi / 3} \int_{-1}^{1} \frac{\mathrm{dx}}{\left(1-\mathrm{x}^{2}\right)^{2 / 3}}=-\mathrm{e}^{\mathrm{i} \pi / 3} \int_{0}^{1} \frac{2}{\left(1-\mathrm{x}^{2}\right)^{2 / 3}} \mathrm{dx} \\
& =-\mathrm{e}^{\mathrm{i} \pi / 3} \int_{0}^{1} \frac{\mathrm{dt}}{\sqrt{\mathrm{t}}(1-\mathrm{t})^{2 / 3}} \text {, substituting } \mathrm{x}=\sqrt{ } \mathrm{t} \\
& =-\mathrm{e}^{\mathrm{i} \pi / 3} \mathrm{~B}\left(\frac{1}{2}, \frac{1}{3}\right) . \text { We choose } \mathrm{w}_{2} \text { as, } \mathrm{w}_{2}=\mathrm{k} \tilde{\mathrm{w}}_{2}=5 \text { where }
\end{aligned}
$$

$$
\mathrm{k}=-5 \mathrm{e}^{-\mathrm{i} \pi / 3} / \mathrm{B}\left(\frac{1}{2}, \frac{1}{3}\right)
$$

To find $\mathrm{w}_{3}$ let us first calculate for $\widetilde{\mathrm{w}}_{3}$.

$$
\begin{aligned}
\tilde{\mathrm{w}}_{3} & =\int_{-1}^{\infty}(\mathrm{x}+1)^{-2 / 3}(\mathrm{x}-1)^{-2 / 3} \mathrm{dx} \\
& =\int_{-1}^{1}(\mathrm{x}+1)^{-2 / 3}(\mathrm{x}-1)^{-2 / 3} \mathrm{dx}+\int_{1}^{\infty}(\mathrm{x}+1)^{-2 / 3}(\mathrm{x}-1)^{-2 / 3} \mathrm{dx} \\
& =-\mathrm{e}^{\mathrm{i} \pi / 3} \mathrm{~B}\left(\frac{1}{2}, \frac{1}{3}\right)+\mathrm{e}^{-\mathrm{i} \pi} \int_{-1}^{\infty}(|\mathrm{x}+1||\mathrm{x}-1|)^{-2 / 3} \mathrm{dx} \\
& =-\mathrm{e}^{-\mathrm{i} \pi / 3} \mathrm{~B}\left(\frac{1}{2}, \frac{1}{3}\right)+\mathrm{e}^{-\mathrm{i} \pi} \int_{-\mathrm{i}}^{\infty}(|\mathrm{x}+1||\mathrm{x}-1|)^{-2 / 3} \mathrm{dx} \\
& =-+\mathrm{e}^{-\mathrm{i} \pi+\mathrm{i} \frac{2 \pi}{3}+\mathrm{i} \frac{2 \pi}{3}} \int_{-1}^{-\infty}|\mathrm{x}+1|^{-2 / 3} \mathrm{e}^{-\mathrm{i} \frac{2 \pi}{3}}|\mathrm{x}-1|^{-2 / 3} \mathrm{e}^{-2 \pi \mathrm{i} / 3} \mathrm{dx} \\
& =-+\mathrm{e}^{1 \pi / 3} \int_{-1}^{-\infty}(\mathrm{x}+1)^{-2 / 3}(\mathrm{x}-1)^{-2 / 3} \mathrm{dx}
\end{aligned}
$$

Now, the value of $\tilde{w}_{3}$ can also be represented by the integral $\int_{-i}^{-\infty}(x+1)^{-2 / 3}(x-1)^{-2 / 3} d x$ when $z$ tends to infinity along the negative real axis. Thus from the above relation, we have

$$
\begin{array}{ll}
\qquad \tilde{w}_{3}=-e^{i \pi / 3} B\left(\frac{1}{2}, \frac{1}{3}\right)+\mathrm{e}^{\mathrm{i} \pi / 3} \tilde{\mathrm{w}}_{3} \\
\text { i.e., } & \tilde{\mathrm{w}}_{3}=-\mathrm{e}^{\mathrm{i} \pi / 3} \cdot \mathrm{e}^{\mathrm{i} \pi / 3} \mathrm{~B}\left(\frac{1}{2}, \frac{1}{3}\right) \\
\text { So, } & \mathrm{w}_{3}=\mathrm{k} \tilde{\mathrm{w}}_{3}=5 \mathrm{e}^{\frac{i \pi}{3}}
\end{array}
$$

Therefore, the three vertices of the equilateral triangle are $\mathrm{w}_{1}=0, \mathrm{w}_{2}=5$ and $\mathrm{w}_{3}=5 \mathrm{e}^{\mathrm{i} \pi / 3}$. Clearly each of it's side is of length 5 unit. The desired transformation is then

$$
\begin{aligned}
f(z) & =K \tilde{f}(z) \\
& =\frac{-5 \mathrm{e}^{-\mathrm{i} \pi / 3}}{\mathrm{~B}\left(\frac{1}{2}, \frac{1}{3}\right)} \int_{-1}^{z}(\mathrm{~s}+1)^{-2 / 3}(\mathrm{~s}-1)^{-2 / 3} \mathrm{ds}
\end{aligned}
$$

which is same as obtained in the first process.
Remark : Following the above technique we can determine a S-C transformation from $\operatorname{Im} \mathrm{z} \geq 0$ onto a triangle, in particular, whose one side opposite to an angle is given.

## Rectangle :

Example 3 : Find a S-C transformation that maps the upper half of the z-plane to the inside of the rectangle in the w-plane with vertices $-\mathrm{a}, \mathrm{a}, \mathrm{a}+\mathrm{ib}$ and $-\mathrm{a}+\mathrm{ib}$ which are the preimages of $-1,1, \alpha$ and $-\alpha$ respectively.

## Solution :



Fig. 64


Fig. 65

Let us first make the identification of the vertices of the rectangle

$$
\begin{aligned}
& \mathrm{w}_{1}=-\mathrm{a}+\mathrm{ib}, \mathrm{w}_{2}=-\mathrm{a}, \mathrm{w}_{3}=\mathrm{a}, \mathrm{w}_{4}=\mathrm{a}+\mathrm{ib} \\
& \alpha_{1}=\alpha_{2}=\alpha_{3}=\alpha_{4}=1 / 2
\end{aligned}
$$

We choose

$$
\mathrm{x}_{1}=-\alpha, \mathrm{x}_{2}=-1, \mathrm{x}_{3}=1, \mathrm{x}_{4}=\alpha
$$

where $\alpha>1$ will be determined later. We are attempting to benefit from the symmetry here, which requires the image $\mathrm{z}=0$ to be $\mathrm{w}=0$. So taking $\mathrm{z}_{0}=0$ we get $\mathrm{B}=0$ in the formula (74) for S-C transformation, which reduces to

$$
\left.f(z)=A \int_{0}^{z}[s+\alpha)(s+1)(s-1)(s-\alpha)\right]^{-1 / 2} d s
$$

$$
\begin{equation*}
=\mathrm{A} \int_{0}^{\mathrm{z}} \frac{\mathrm{ds}}{\sqrt{\left[\left(1-\mathrm{s}^{2}\right)\left(\alpha^{2}-\mathrm{s}^{2}\right)\right]}}(\equiv \phi(\mathrm{z}, \alpha)) \tag{80}
\end{equation*}
$$

The constant A may be found by using the fact that $f(1)=$ a i.e.,

$$
\begin{gather*}
a=A \int_{0}^{1} \frac{d s}{\sqrt{\left[\left(1-s^{2}\right)\left(\alpha^{2}-s^{2}\right)\right]}} \text { or } A=\mathrm{a} / \int_{0}^{1} \frac{d s}{\sqrt{\left[\left(1-s^{2}\right)\left(\alpha^{2}-s^{2}\right)\right]}} \\
=\mathrm{a} / \phi(\alpha), \text { say } \tag{81}
\end{gather*}
$$

To find $\alpha$, we apply $f(\alpha)=a+i b$,

$$
\begin{aligned}
& a+i b=\frac{a}{\phi(\alpha)} \int_{0}^{a} \frac{d s}{\sqrt{\left[\left(1-s^{2}\right)\left(\alpha^{2}-s^{2}\right)\right]}} \\
& =\frac{a}{\phi(\alpha)}\left\{\int_{0}^{1} \frac{d s}{\left.\sqrt{\left[\left(1-s^{2}\right)\left(\alpha^{2}-s^{2}\right)\right]}+i \int_{1}^{\alpha} \frac{d s}{\sqrt{\left[\left(s^{2}-1\right)\left(\alpha^{2}-s^{2}\right)\right]}}\right\}}\right.
\end{aligned}
$$

from which, equating imaginary parts, we arrive at

$$
\mathrm{b} \phi(\alpha)=\alpha \int_{1}^{\alpha} \frac{\mathrm{ds}}{\sqrt{\left[\left(s^{2}-1\right)\left(\alpha^{2}-\mathrm{s}^{2}\right)\right]}}
$$

Since a and bare known, this equation determines $\alpha$, which gives rise to the evaluation of $\phi(\alpha)$ i.e. A is completely known.

Note : The function $\phi(\mathrm{z}, \alpha)$, given in (80), which involves z as the upper limit of an integral, is called an elliptic integral of the first kind and it is not an elementary function. The real definite integral $\phi(\alpha)$ in (81) is called a complete elliptic integral of the first kind.

Example 4 : Find a Schwarz-Christoffel transformation that maps the upper half of the z-plane to the vertical semi-infinite strip $-\pi / 2<u<\pi / 2, v>0$ of the w-plane.

## Solution :



Fig. 66


Fig. 67

Here we take $\mathrm{x}_{1}=-1, \mathrm{x}_{2}=1$ and $\mathrm{x}_{3}=\infty$ and the image points are $\mathrm{w}_{1}=-\pi / 2$ and $\mathrm{w}_{2}=\pi / 2$ respectively, so that a S-C transformation can be written as

$$
\begin{aligned}
f(z) & =A \int_{z_{0}}^{z}(s+1)^{-1 / 2}(s-1)^{-1 / 2} d s+B \\
& =A \int_{z_{0}}^{\tau} \frac{1}{\left(s^{2}-1\right)^{1 / 2}} d s+B \\
& =\tilde{A} \log \left(i z \sqrt{1-z^{2}}\right)+\tilde{B}
\end{aligned}
$$

Using $f(-1)=-\frac{\pi}{2}$ and $f(1)=\frac{\pi}{2}$, we find

$$
\mathrm{f}(\mathrm{z})=-\mathrm{i} \log \left(\mathrm{iz}+\sqrt{1-\mathrm{z}^{2}}\right),
$$

Choosing a suitable branch of the logarithm.
Unit 6 - Entire and Meromorphic Functions
Structure
6.0 Objectives
6.1 Entire function
6.2 Infinite Products
6.3 Infinite product of functions
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6.0 The Objectives of the Chapter

In this chapter we shall study entire functions, their growth properties and meromorphic functions. Infinite products and their convergence will be discussed. Properties of zeros of
an entire function, convex functions, gamma function and its important properties will also be discussed.

### 6.1 Entire function

A function $f(z)$ analytic in the finite complex plane is said to be entire (or sometimes integral) function. Clearly, the sum, difference and product of two or more entire functions are entire functions.

Examples : The polynomial function $P(z)=a_{0}+a_{1} Z+\ldots+a_{n} z^{n}$, exponential function $\mathrm{e}^{\mathrm{z}}, \sin \mathrm{z}, \cos \mathrm{z}$ etc. are entire functions.

Let us consider the first example, the polynomial function. It is evident that $\mathrm{P}(\mathrm{z})$ can be uniquely expressed as a product of linear factors in the form

$$
A_{0}\left(1-\frac{z}{z_{1}}\right)\left(1-\frac{z}{z_{2}}\right) \cdots\left(1-\frac{z}{z_{n}}\right), \text { if } a_{0} \neq 0
$$

or,

$$
\begin{equation*}
A_{p} z^{p}\left(1-\frac{z}{\zeta_{1}}\right)\left(1-\frac{z}{\zeta_{2}}\right) \cdots\left(1-\frac{z}{\zeta_{n-p}}\right) \text {, if } a_{0}=a_{1}=\cdots a_{p-1}=0, a_{p} \neq 0, \tag{82}
\end{equation*}
$$

where $A_{0}\left(\right.$ or, $\left.A_{p}\right)$ is constant and $z=z_{1}, z_{2}, \ldots, z_{n}\left(\right.$ or, $\left.z=0, \varsigma_{1}, \varsigma_{2}, \ldots, \varsigma_{n-p}\right)$ are the zeros of $\mathrm{P}(\mathrm{z})$, multiple zeros are counted according to their multiplicities. There arises a natural question : whether any entire function can be expressed in a similar manner in terms of its zeros. The observations are as follows :
(i) There may exist entire function which never vanishes,
(ii) If an entire function possesses finite number of zeros, then it is always possible to express it in the form (82) stated above. But when the number of zeros are infinite the form (82) reduces to a product of infinite number of linear factors which need not always be convergent. We first consider infinite products of complex numbers and functions.

### 6.2 Infinite Products

An infinite product is an expression of the form

$$
\begin{equation*}
\prod_{n=1}^{\infty} p_{n} \tag{83}
\end{equation*}
$$

where $\mathrm{p}_{1}, \mathrm{p}_{2}, \ldots, \mathrm{p}_{\mathrm{n}}, \ldots$ are non-zero complex factors. If we allow any of the factors be zero, it is evident that the infinite product would be zero regardless of the behaviour of the other terms.

Let $\quad \mathrm{P}_{\mathrm{n}}=\mathrm{p}_{1} \mathrm{p}_{2} \ldots \mathrm{p}_{\mathrm{n}}$.
If $\mathrm{P}_{\mathrm{n}}$ tends to a finite limit (non-zero) p as n tends to infinity, we say that the infinite product (83) is convergent and write as

$$
\begin{equation*}
\prod_{n=1}^{\infty} p_{n}=p \tag{84}
\end{equation*}
$$

An infinite product which does not tend to a non-zero finite limit as n tends to infinity is said to be divergent.

To find the necessary condition for convergence for the infinite product $\prod_{n=1}^{\infty} p_{n}$, say (84) holds, then writing $\mathrm{p}_{\mathrm{n}}$ as

$$
\mathrm{p}_{\mathrm{n}}=\frac{\mathrm{P}_{\mathrm{n}}}{\mathrm{P}_{\mathrm{n}-1}}
$$

we conclude in view of (84) that $\lim _{n \rightarrow \infty} p_{n}=\lim _{n \rightarrow \infty} \frac{P_{n}}{P_{n-1}}=\frac{P}{P}=1$
Thus, $\quad \lim _{n \rightarrow \infty} p_{n}=1$
is a necessary condition for convergence of the infinite product (83). It is then better to write the product as

$$
\begin{equation*}
\prod_{n=1}^{\infty}\left(1+a_{n}\right) \tag{86}
\end{equation*}
$$

so that $\mathrm{a}_{\mathrm{n}} \rightarrow 0$ as $\mathrm{n} \rightarrow \infty$ is a necessary condition for convergence.
Theorem 6.1 : The infinite product (86) converges if and only if

$$
\begin{equation*}
\sum_{n=1}^{\infty} \log \left(1+a_{n}\right) \tag{87}
\end{equation*}
$$

converges. We use the principal branch of the log function and omit, as usual, the terms with $\mathrm{a}_{\mathrm{n}}=-1$.

Proof. Let $P_{n}=\prod_{k=1}^{n}\left(1+a_{k}\right)$ and $S_{n}=\sum_{k=1}^{n} \log \left(1+a_{k}\right)$.
Then $\log P_{n}=S_{n}$ and $P_{n}=e^{S n}$. Now if the given series is convergent i.e. $S_{n} \rightarrow S$ as $\mathrm{n} \rightarrow \infty, \mathrm{P}_{\mathrm{n}}$ tends to the limit $\mathrm{P}=\mathrm{e}^{\mathrm{S}}(\neq 0)$. This proves the sufficiency of the condition.

Conversely, assume that the product converges i.e. $P_{n} \rightarrow P(\neq 0)$ as $n \rightarrow \infty$. We shall show, by virtue of $P_{n}=e^{S n}$, that the series (87) converges to some value of $\log$ $P$, not necessarily the principal value of $\log \mathrm{P}$.

$$
\text { For } \mathrm{n} \rightarrow \infty, \frac{\mathrm{P}_{\mathrm{n}}}{\mathrm{P}} \rightarrow 1 \text { and } \log \left(\frac{\mathrm{P}_{\mathrm{n}}}{\mathrm{P}}\right) \rightarrow 0 \text {. }
$$

Now there exists an integer $\mathrm{K}_{\mathrm{n}}$ such that

$$
\begin{equation*}
\log \left(\frac{P_{n}}{P}\right)=S_{n}-\log P+2 k_{n} \pi i \tag{88}
\end{equation*}
$$

To establish the convergence of the sequence $\left\{\mathrm{k}_{\mathrm{n}}\right\}$, we form the difference

$$
\begin{aligned}
\left(\mathrm{k}_{\mathrm{n}+1}-\mathrm{k}_{\mathrm{n}}\right) 2 \pi \mathrm{i} & =\log \left(\frac{\mathrm{P}_{\mathrm{n}+1}}{\mathrm{P}}\right)-\log \left(\frac{\mathrm{P}_{\mathrm{n}}}{\mathrm{P}}\right)-\log \left(1+\mathrm{a}_{\mathrm{n}+1}\right) \\
& =\mathrm{i}\left\{\operatorname{Arg}\left(\frac{\mathrm{P}_{\mathrm{n}+1}}{\mathrm{P}}\right)-\operatorname{Arg}\left(\frac{\mathrm{P}_{\mathrm{n}}}{\mathrm{P}}\right)-\operatorname{Arg}\left(1+\mathrm{a}_{\mathrm{n}+1}\right)\right\}
\end{aligned}
$$

and that

$$
\mathrm{k}_{\mathrm{n}+1}-\mathrm{k}_{\mathrm{n}}=\frac{1}{2 \pi}\left\{\operatorname{Arg}\left(\frac{\mathrm{P}_{\mathrm{n}+1}}{\mathrm{P}}\right)-\operatorname{Arg}\left(\frac{\mathrm{P}_{\mathrm{n}}}{\mathrm{P}}\right)-\operatorname{Arg}\left(1+\mathrm{a}_{\mathrm{n}+1}\right)\right\}
$$

tends to zero as $\mathrm{n} \rightarrow \infty$, and let the limit of the sequence $\left\{\mathrm{k}_{\mathrm{n}}\right\}$ be k .
Taking limit in (88), we find that

$$
\mathrm{S}_{\mathrm{n}} \rightarrow \log \mathrm{P}-2 \mathrm{k} \pi \mathrm{i}
$$

and so the condition assumed is necessary.
Definition : An infinite product $\prod_{n=1}^{\infty}\left(1+a_{n}\right)$ is absolutely convergent if and only if $\sum_{n=1}^{\infty}\left|\log \left(1+a_{n}\right)\right|$ is convergent.

Theorem 6.2 : The infinite product (86) converges absolutely if and only if the series $\sum \mathrm{a}_{\mathrm{n}}$ converges absolutely.

Proof : If $\sum \mathrm{a}_{\mathrm{n}}$ converges absolutely, then in particular $\mathrm{a}_{\mathrm{n}} \rightarrow 0$ as $\mathrm{n} \rightarrow \infty$. Also, if $\sum_{n=1}^{\infty} \log \left(1+a_{n}\right)$ converges absolutely then $\log \left(1+a_{n}\right) \rightarrow 0$ and $a_{n} \rightarrow 0$. Thus in
either of the cases $\mathrm{a}_{\mathrm{n}} \rightarrow 0$ and we can take $\left|\mathrm{a}_{\mathrm{n}}\right| \leq \frac{1}{2}$ for sufficiently large n . Then by elementary calculation,

$$
\begin{gathered}
\left|1-\frac{\log \left(1+a_{n}\right)}{a_{n}}\right|=\left|\frac{a_{n}}{2}-\frac{a_{n}^{2}}{3}+\cdots\right| \\
\leq \frac{1}{2}\left\{\left|a_{n}\right|+\left|a_{n}\right|^{2}+\left|a_{n}\right|^{3}+\cdots\right\} \leq \frac{1}{2}, n=\text { large enough. It follows that } \\
\left.\frac{1}{2}\left|a_{n}\right| \leq \log \left(1+a_{n}\right)\left|\leq \frac{3}{2}\right| a_{n} \right\rvert\,
\end{gathered}
$$

confirming the occurrence of the absolute convergence simultaneously for the two series.

### 6.3 Infinite product of functions

So far we have considered infinite product of complex numbers. Now we shall study infinite products whose factors are functions of a complex variable. Some of the factors (finite in number) may vanish on a region considered. In that case we consider the infinite product omitting those factors. The theorems proved earlier hold good in this case too with some modifications.

Definition : (Uniform convergence of infinite products)
An infinite product

$$
\begin{equation*}
\prod_{\mathrm{n}=1}^{\infty}\left\{1+\mathrm{a}_{\mathrm{n}}(\mathrm{z})\right\} \tag{89}
\end{equation*}
$$

where the functions $a_{n}(z)$ are defined on a region $D$, is said to be uniformly convergent on D if the sequence of partial products

$$
\mathrm{P}_{\mathrm{n}}(\mathrm{z})=\prod_{\mathrm{k}=1}^{\mathrm{n}}\left\{1+\mathrm{a}_{\mathrm{k}}(\mathrm{z})\right\}
$$

converges uniformly to a non-zero limit on D .
Theorem 6.3 : An infinite product (89) is uniformly convergent on a domain D if the series $\sum_{\mathrm{n}=1}^{\infty}\left|\mathrm{a}_{\mathrm{n}}(\mathrm{z})\right|$ converges uniformly and has a bounded sum there.

Proof : Let M be the upper bound of the sum $\sum\left|a_{n}(z)\right|$ on D. Then

$$
\left\{1+\mathrm{a}_{1}(\mathrm{z}) \mid\right\}\left\{1+\left|\mathrm{a}_{2}(\mathrm{z})\right|\right\} \ldots\left\{1+\left|\mathrm{a}_{\mathrm{n}}(\mathrm{z})\right|\right\}<\mathrm{e}^{\left|\mathrm{a}_{1}(\mathrm{z})\right|+\left|\mathrm{a}_{2}(\mathrm{z})\right|+\ldots \mathrm{a}_{\mathrm{n}}(\mathrm{z}) \mid} \leq \mathrm{e}^{\mathrm{M}}
$$

Let us consider the sequence $\left\{\mathrm{Q}_{\mathrm{n}}\right\}$ with

$$
\mathrm{Q}_{\mathrm{n}}(\mathrm{z})=\prod_{\mathrm{k}=1}^{\mathrm{n}}\left\{1+\left|\mathrm{a}_{\mathrm{k}}(\mathrm{z})\right|\right\}
$$

We observe

$$
\begin{aligned}
\mathrm{Q}_{\mathrm{n}}(\mathrm{z})-\mathrm{Q}_{\mathrm{n}-1}(\mathrm{z})= & \left\{1+\left|\mathrm{a}_{1}(\mathrm{z})\right|\right\}\left\{1+\left|\mathrm{a}_{2}(\mathrm{z})\right|\right\} \ldots\left\{1+\mid \mathrm{a}_{\mathrm{n}-1}(\mathrm{z})\right\}\left|\mathrm{a}_{\mathrm{n}}(\mathrm{z})\right| \\
& <\mathrm{e}^{\mathrm{M}}\left|\mathrm{a}_{\mathrm{n}}(\mathrm{z})\right|
\end{aligned}
$$

Now since the series $\sum \mathrm{a}_{\mathrm{n}}(\mathrm{z}) \mid$ is uniformly convergent, the series $\sum\left\{\mathrm{Q}_{\mathrm{n}}(\mathrm{z})-\mathrm{Q}_{\mathrm{n}-1}(\mathrm{z})\right\}$ is uniformly convergent. Thus the sequence $\left\{\mathrm{Q}_{\mathrm{n}}\right\}$ tends to a limit. Again

$$
\left|P_{n}(z)-P_{n-1}(z)\right| \leq Q_{n}(z)-Q_{n-1}(z),
$$

so the result follows.
Theorem 6.4 : An infinite product $\prod_{n=1}^{\infty}\left\{1+a_{n}(z)\right\}$ converges uniformly and absolutely in a closed bounded domain $D$ if each function $a_{n}(z)$ satisfies $\left|a_{n}(z)\right| \leq M_{n}$ for all $z \varepsilon D$ and $M_{n}$ is independent of $z$ and moreover $\Sigma M_{n}$ is convergent.

Proof : Given $\Sigma M_{n}$ is convergent, so the infinite product $M=\prod_{n=1}^{\infty}\left(1+M_{n}\right)$ converges by theorem 6.2

Now, for $\mathrm{n}>\mathrm{m}$

$$
\begin{equation*}
\left|\mathrm{Q}_{\mathrm{n}}(\mathrm{z})-\mathrm{Q}_{\mathrm{m}}(\mathrm{z})\right|=\left|\mathrm{Q}_{\mathrm{m}}(\mathrm{z})\right| \prod_{\mathrm{m}+1}^{\mathrm{n}}\left\{1+\mathrm{a}_{\mathrm{k}}(\mathrm{z})\right\}-1 \mid \tag{90}
\end{equation*}
$$

Again,

$$
\begin{aligned}
\prod_{\mathrm{m}+1}^{\mathrm{n}}\left\{1+\mathrm{a}_{\mathrm{k}}(\mathrm{z})\right\}-1= & \sum_{\mathrm{k}=\mathrm{m}+\mathrm{l}}^{\mathrm{n}} \mathrm{a}_{\mathrm{k}}(\mathrm{z})+\sum_{\mathrm{i}, \mathrm{j}}^{\mathrm{n}} \mathrm{a}_{\mathrm{i}}(\mathrm{z}) \mathrm{a}_{\mathrm{j}}(\mathrm{z})+\sum_{\mathrm{i}, \mathrm{j}, l}^{\mathrm{n}} \mathrm{a}_{\mathrm{i}}(\mathrm{z}) \mathrm{a}_{\mathrm{j}}(\mathrm{z}) \mathrm{a}_{l}(\mathrm{z}) \\
& +\ldots+\mathrm{a}_{\mathrm{m}+1}(\mathrm{z}) \mathrm{a}_{\mathrm{m}+2}(\mathrm{z}) \ldots \mathrm{a}_{\mathrm{n}}(\mathrm{z}) .
\end{aligned}
$$

Taking moduli

$$
\begin{aligned}
\mid \prod_{\mathrm{m}+1}^{\mathrm{n}}\left\{1+\mathrm{a}_{\mathrm{k}}(\mathrm{z})\right. & \}-1 \mid \leq \sum_{\mathrm{k}=\mathrm{m}+1}^{\mathrm{n}} \mathrm{M}_{\mathrm{k}}+\sum_{\mathrm{i}, \mathrm{j}}^{\mathrm{n}} \mathrm{M}_{\mathrm{i}} \mathrm{M}_{\mathrm{j}}+\sum_{\mathrm{i}, \mathrm{j}, l}^{\mathrm{n}} \mathrm{M}_{\mathrm{i}} \mathrm{M}_{\mathrm{j}} \mathrm{M}_{l}+ \\
& +\ldots+\mathrm{M}_{\mathrm{m}+1} \mathrm{M}_{\mathrm{m}+2} \cdots \mathrm{M}_{\mathrm{n}} \\
= & \prod_{\mathrm{m}+1}^{\mathrm{n}}\left(1+\mathrm{M}_{\mathrm{k}}\right)-1
\end{aligned}
$$

Utilising this in (90) we obtain

$$
\begin{align*}
&\left|\mathrm{Q}_{\mathrm{n}}(\mathrm{z})-\mathrm{Q}_{\mathrm{m}}(\mathrm{z})\right| \leq \prod_{\mathrm{k}=1}^{\mathrm{m}}\left(1+\mathrm{M}_{\mathrm{k}}\right)\left\{\prod_{\mathrm{m}=1}^{\mathrm{n}}\left(1+\mathrm{M}_{\mathrm{k}}\right)-1\right\} \\
&=\prod_{\mathrm{k}=1}^{\mathrm{n}}\left(1+\mathrm{M}_{\mathrm{k}}\right)-\prod_{\mathrm{k}=1}^{\mathrm{m}}\left(1+\mathrm{M}_{\mathrm{k}}\right) \tag{91}
\end{align*}
$$

Now as the infinite product $\prod_{1}^{\infty}\left(1+\mathrm{M}_{\mathrm{k}}\right)$ is convergent, we choose m large enough so that r.h.s in (91) is less than $\varepsilon$ and hence

$$
\left|\mathrm{Q}_{\mathrm{n}}(\mathrm{z})-\mathrm{Q}_{\mathrm{m}}(\mathrm{z})\right|<\varepsilon \text {, when } \mathrm{n}>\mathrm{m}
$$

Thus the sequence $\left\{\mathrm{Q}_{\mathrm{n}}(\mathrm{z})\right\}$ converge uniformly, since m depends only on $\varepsilon$.
Finally, absolute convergence of the infinite product follows on utilising Th. 6.2
Example 1: Test for convergence of the infinite product

$$
\prod_{n=1}^{\infty}\left(1-\frac{z^{2}}{n^{2}}\right)
$$

Solution : The terms of the product vanish when $\mathrm{z}= \pm 1, \pm 2, \ldots$ etc.
Here $\mathrm{a}_{\mathrm{n}}(\mathrm{z})=-\frac{\mathrm{z}^{2}}{\mathrm{n}^{2}}$ and $\left|\mathrm{a}_{\mathrm{n}}(\mathrm{z})\right| \leq\left|\mathrm{z}^{2}\right| \frac{1}{\mathrm{n}^{2}}$
Now since the series $\sum \frac{1}{\mathrm{n}^{2}}$ is convergent, the given infinite product is uniformly and absolutely convergent in the entire plane excluding the points $\mathrm{z}= \pm 1, \pm 2$, etc.

Example 2: Discuss the convergence of the infinite product

$$
\left(1-\frac{z}{1}\right)\left(1+\frac{z}{1}\right)\left(1-\frac{z}{2}\right)\left(1+\frac{z}{2}\right) \cdots
$$

Solution : Let $\mathrm{P}_{\mathrm{n}}(\mathrm{z})=\prod_{\mathrm{k}=1}^{\mathrm{n}}\left(1-\frac{\mathrm{z}^{2}}{\mathrm{k}^{2}}\right)$ and we consider a bounded closed domain D which does not contain the points $\mathrm{z}= \pm 1, \pm 2, \ldots$. The sequence $\left\{\mathrm{P}_{\mathrm{n}}(\mathrm{z})\right\}$ converges uniformly in D (see example 1). Again let

$$
\begin{aligned}
& \mathrm{F}_{2 \mathrm{n}}(\mathrm{z})=\left(1-\frac{\mathrm{z}}{1}\right)\left(1+\frac{\mathrm{z}}{1}\right)\left(1-\frac{\mathrm{z}}{2}\right)\left(1+\frac{\mathrm{z}}{2}\right) \cdots\left(1-\frac{\mathrm{z}}{\mathrm{n}}\right)\left(1+\frac{\mathrm{z}}{\mathrm{n}}\right) \\
& \mathrm{F}_{2 \mathrm{n}+1}(\mathrm{z})=\mathrm{F}_{2 \mathrm{n}}(\mathrm{z})\left(1-\frac{\mathrm{z}}{\mathrm{n}+1}\right)
\end{aligned}
$$

then

$$
\mathrm{F}_{2 \mathrm{n}}(\mathrm{z})=\mathrm{P}_{\mathrm{n}}(\mathrm{z}) \text { and } \mathrm{F}_{2 \mathrm{n}+1}(\mathrm{z})=\left(1-\frac{\mathrm{z}}{\mathrm{n}+1}\right) \mathrm{P}_{\mathrm{n}}(\mathrm{z})
$$

and obviously the sequences $\mathrm{F}_{2}, \mathrm{~F}_{4}, \mathrm{~F}_{6}, \ldots$ and $\mathrm{F}_{1}, \mathrm{~F}_{3}, \mathrm{~F}_{5} \ldots$ converge uniformly in D . Hence the given infinite product converges uniformly in D .

To test for the absolute convergence of the given product we notice that

$$
\sum_{\mathrm{i}}^{\infty}\left|\mathrm{a}_{\mathrm{n}}\right|=\left\lvert\, \mathrm{z}\left\{1+1+\frac{1}{2}+\frac{1}{2}+\frac{1}{3}+\frac{1}{3}+\cdots\right\}\right.
$$

and it is divergent since the series on the right is divergent and $|z|$ is finite. Therefore the given product does not converge absolutely.

Considering the theorem 4.4 on uniformly convergent sequence of analytic functions [(14) Page-72] we get the following theorem :

Theorem 6.5: If an infinite product $\Pi\left\{1+\mathrm{f}_{\mathrm{n}}(\mathrm{z})\right\}$ converges uniformly to $\mathrm{f}(\mathrm{z})$ in a bounded closed domain $D$ and if each function $f_{n}(z)$ is analytic in $D$, then $f(z)$ is also analytic in $D$.

### 6.4 Weierstrass' Factorization

Theorem 6.6: If $f(z)$ is an entire function and never vanishes on $\mathscr{C}$, then $f(z)$ is of the form $f(z)=e^{g(z)}$, or, more generally, $f(z)=e^{g(z)}, c \neq 0$, constant.
where $g(z)$ is also an entire function.
Proof : Since $f$ is entire and never vanishes on $\mathbb{C}, \mathrm{f}^{1 / f}$ is also entire and is thus the derivative of an entire function $g(z)$. [follows from Result 1, PG(MT) 02-complex analysis [14, page-54]. Then

$$
\begin{array}{rlrl}
\frac{\mathrm{f}^{\prime}}{\mathrm{f}} & =\mathrm{g}^{\prime} \\
\text { i.e. } \quad & \mathrm{f}^{\prime} & =\mathrm{fg}^{\prime}
\end{array}
$$

Now, $\left(\mathrm{fe}^{-\mathrm{g}}\right)^{\prime}=\mathrm{f}^{\prime} \mathrm{e}^{-\mathrm{g}}-\mathrm{fg}^{\prime} \mathrm{e}^{-\mathrm{g}}=0$
Hence, $\quad f(z)=e^{g}(z)$ proving the result.
Assume now that f possesses finitely many zeros, a zero of order $\mathrm{m}>0$ at the origin, and the non-zero ones, possibly repeated are $a_{1}, \ldots a_{n}$. Then

$$
f(z)=z^{m} \prod_{k=1}^{n}\left(1-\frac{z}{a_{n}}\right) e^{g(z)}
$$

where g is entire.
This is clear, since if we divide f by the factors which produce zero at the points $\mathrm{z}=$ $0, a_{1}, \ldots, a_{n}$ we get an entire function with no zeros.

However we cannot expect, in general, such a simple formula to hold in the case of infinitely many zeros. Here we have to take care of convergence problems for an infinite product. In fact the obvious generalization.

$$
f(z)=z^{m} \prod_{k=1}^{n}\left(1-\frac{z}{a_{k}}\right) e^{g(z)}
$$

is valid in a bounded closed domain D if the infinite product converges uniformly in D.

## Theorem 6.7 (Weierstrass' Factorization Theorem) :-

Let $\left\{a_{n}\right\}$ be a sequence of complex numbers with the property $a_{n} \rightarrow \infty$ as $n \rightarrow \infty$. Then it is possible to construct an entire function $f(z)$ with zeros precisely at these points.

Proof : We need Weierstrass' primary factors to construct the desired function. The expressions $E(z, o)=1-z, E(z, p)=(1-z) e^{z+\frac{z^{2}}{2}+\cdots+\frac{z^{p}}{p}}, p=1,2 \ldots$, are called Weierstrass' primary factors. Each primary factor is an entire function having only one simple zero at $\mathrm{z}=1$.

Now, when $|z|<1$ we have, $\log E(z, p)=\log (1-z)+z+\frac{z^{2}}{2}+\cdots+\frac{z^{p}}{p}$

$$
=\left(-z-\frac{z^{2}}{2}-\ldots-\frac{z^{p}}{p}-\frac{z^{p+1}}{p+1}-\ldots\right)+\left(z+\frac{z^{2}}{2}+\ldots+\frac{z^{p}}{p}\right)=-\frac{z^{p+1}}{p+1}-\frac{z^{p+2}}{p+2}-\ldots
$$

Here we have taken the principal branch of $\log (1-z)$.
Hence if

$$
\begin{align*}
& |z| \leq \frac{1}{2},|\log \mathrm{E}(\mathrm{z}, \mathrm{p})| \leq|\mathrm{z}|^{p+1}+|z|^{p+2}+\ldots=|z|^{p+1}\left(1+|z|+\left|z^{2}\right|+\ldots\right) \\
& \leq|z|^{p+1}\left(1+\frac{1}{2}+\frac{1}{2^{2}}+\ldots\right)=2|z|^{p+1} \ldots \tag{92}
\end{align*}
$$

We may suppose that the origin is not a zero of the entire function $f(z)$ to be constructed so that $\mathrm{a}_{\mathrm{n}} \neq 0$ for all n .

For, if origin is a zero of $f(z)$ of order $m$ we need only multiply the constructed function by $\mathrm{z}^{\mathrm{m}}$. We also arrange the zeros in order of non-decreasing modulus (if several distinct points $\mathrm{a}_{\mathrm{n}}$ have the same modulus, we take them in any order) so that $\left|a_{1}\right| \leq\left|a_{2}\right| \leq \ldots$. Let $\left|a_{n}\right|=r_{n}$.

Since $r_{n} \rightarrow \infty$ we can always find a sequence of positive inegers
$m_{1}, m_{2}, \ldots m_{n}, \ldots$ such that the series $\sum_{\mathrm{n}=1}^{\infty}\left(\frac{\mathrm{r}}{\mathrm{r}_{\mathrm{n}}}\right)^{\mathrm{m}_{\mathrm{n}}}$ converges for all positive values of r .

In fact, we may take $m_{n}=n$ since for any given value of $r$, we have $\left(\frac{r}{r_{n}}\right)^{n}<\frac{1}{2^{n}}$ for all sufficiently large n and the series is therefore convergent. Next we take an arbitrary positive number R and choose the integer N such that $\mathrm{r}_{\mathrm{N}} \leq 2 \mathrm{R}<\mathrm{r}_{\mathrm{N}+1}$. Hence, when $|z| \leq R$ and $n>N$ we have,
$\left|\frac{\mathrm{z}}{\mathrm{a}_{\mathrm{n}}}\right| \leq \frac{\mathrm{R}}{\mathrm{r}_{\mathrm{n}}} \leq \frac{\mathrm{R}}{\mathrm{r}_{\mathrm{N}+1}}<\frac{1}{2}$ and so by (92), $\left|\log E\left(\frac{z}{a_{n}}, m_{n}\right)\right| \leq 2\left|\frac{R}{r_{n}}\right|^{m_{n}+1}$ By Weierstrass' M-test the series $\sum_{n=1}^{\infty} \log E\left(\frac{z}{a_{n}}, m_{n}\right)$ converges absolutely and uniformly when $|z| \leq R$ and so the infinite product $\prod_{n=1}^{\infty} E\left(\frac{z}{a_{n}}, m_{n}\right)$ converges absolutely and uniformly in the disc $|z| \leq R$, however large $R$ may be. Hence the above product represents an entire function, say $\mathrm{G}(\mathrm{z})$.

> Thus,

$$
\mathrm{G}(\mathrm{z})=\prod_{\mathrm{n}=1}^{\infty} \mathrm{E}\left(\frac{\mathrm{z}}{\mathrm{a}_{\mathrm{n}}}, \mathrm{~m}_{\mathrm{n}}\right)
$$

With the same value of $R$, we choose another integer $k$ such that $r_{k} \leq R<r_{K+1}$.
Then each of the functions of the sequence $\prod_{n=1}^{m} E\left(\frac{z}{a_{n}}, m_{n}\right), m=k+1, k+2, \ldots$, vanish at the points $a_{1} \ldots, a_{k}$ and nowhere else in $|z| \leq R$. Hence by Hurwitz's theoroem the only zeros of $G$ in $|z| \leq R$ are $a_{1}, \ldots a_{k}$. Since $R$ is arbitrary, this implies that the only zeros of $G$ are the points of the sequence $\left\{a_{n}\right\}$.

Now, if origin is a zero of order $m$ of the required entire function $f(z)$, then $f(z)$ is of the form $f(z)=z^{m} G(z)$. Again, for any entire function $g(z)$, $e^{g(z)}$ is also an entire function without any zero. Hence the general form of the required entire function $f(z)$ is

$$
\begin{align*}
f(z) & =z^{m} e^{g(z)} G(z) \\
& =z^{m} e^{g(z)} \prod_{n=1}^{\infty} E\left(\frac{z}{a_{n}}, m_{n}\right)  \tag{94}\\
& =z^{m} e^{g(z)} \prod_{n=1}^{\infty}\left(1-\frac{z}{a_{n}}\right) e^{\frac{z}{a_{n}}+\frac{1}{2}\left(\frac{z}{a_{n}}\right)^{2}+\ldots+\frac{1}{m_{n}}\left(\frac{z}{a_{n}}\right)^{m_{n}}} \tag{95}
\end{align*}
$$

Remark : As there are many possible sequences $\left\{\mathrm{m}_{\mathrm{n}}\right\}$ in the construction of the function $\mathrm{G}(\mathrm{z})$ and ultimately of $\mathrm{f}(\mathrm{z})$, the form of the function $\mathrm{f}(\mathrm{z})$ achieved is not unique.

### 6.5 Counting zeros of analytic functions

The rate of growth of an entire function is closely related to the density of zeros. We have a quite effective formula in this regard due to J.L.W.V. Jensen, a Danish mathematician who discovered it in the year 1899.

## Theorem 6.8 [Jensen's Formula] :-

Let $\mathrm{f}(\mathrm{z})$ be analytic on $|\mathrm{z}| \leq \mathrm{R}, \mathrm{f}(0) \neq 0$ and $\mathrm{f}(\mathrm{z}) \neq 0$ on $|\mathrm{z}|=R$. If $\mathrm{a}_{1}, \ldots, \mathrm{a}_{\mathrm{n}}$ be the zeros of $f(z)$ within the circle $|z|=R$, multiple zeros being repeated according to their multiplicities, then

$$
\begin{equation*}
\log |\mathrm{f}(0)|=\frac{1}{2 \pi} \int_{0}^{2 \pi} \log \left|\mathrm{f}\left(\operatorname{Re}^{\mathrm{i} \theta}\right)\right| \mathrm{d} \theta-\sum_{\mathrm{k}=1}^{\mathrm{n}} \log \left(\frac{\mathrm{R}}{\left|\mathrm{a}_{\mathrm{k}}\right|}\right) \cdots \tag{96}
\end{equation*}
$$

Proof : Let $\phi(\mathrm{z})=\mathrm{f}(\mathrm{z}) . \prod_{\mathrm{k}=1}^{\mathrm{n}} \frac{\mathrm{R}^{2}-\overline{\mathrm{a}}_{\mathrm{k}} \mathrm{z}}{\mathrm{R}\left(\mathrm{z}-\mathrm{a}_{\mathrm{k}}\right)} \cdots$
The zeros of the denominator of $\phi(z)$ are also the zeros of $f(z)$ of the same order. Hence the zeros of $f(z)$ cancels the poles $a_{n}$ in the product and so $\phi(z)$ is analytic on $|z| \leq R$. Also, $\phi(z) \neq 0$ on $|z| \leq R$. For, if $R^{2}-\bar{a}_{k} z=0$ then $z=\frac{R^{2}}{\bar{a}_{k}}$ is the inverse point of $\mathrm{a}_{\mathrm{k}}$ with respect to the circle $|\mathrm{z}|=\mathrm{R}$ and so lies outside the circle. Again,

$$
|\phi(\mathrm{z})|=|\mathrm{f}(\mathrm{z})|\left|\frac{\mathrm{R}^{2}-\overline{\mathrm{a}}_{1} \mathrm{z}}{\mathrm{R}\left(\mathrm{z}-\mathrm{a}_{1}\right)}\right| \cdots\left|\frac{\mathrm{R}^{2}-\overline{\mathrm{a}}_{\mathrm{n}} \mathrm{z}}{\mathrm{R}\left(\mathrm{z}-\mathrm{a}_{\mathrm{n}}\right)}\right| \text {. Now, when }|\mathrm{z}|=\mathrm{R}
$$

we have, $\left|\frac{R^{2}-\bar{a}_{k} z}{R\left(z-a_{k}\right)}\right|=\left|\frac{\bar{z}-\bar{a}_{k} z}{R\left(z-a_{k}\right)}\right|=\frac{|z|}{R}\left|\frac{\bar{z}-\bar{a}_{k}}{z-a_{k}}\right|=1$
Hence, $|\phi(\mathrm{z})|=|\mathrm{f}(\mathrm{z})|$ on $|\mathrm{z}|=\mathrm{R}$.
Since $\phi(z)$ is analytic and non-zero on $|z| \leq R, \log \phi(z)$ is also analytic on $|z| \leq R$ and consequently $\operatorname{Re} \log \phi(z)=\log |\phi(z)|$ is harmonic on $|z| \leq R$. Hence by Gauss' mean value theorem,

$$
\begin{equation*}
\log |\phi(0)|=\frac{1}{2 \pi} \int_{0}^{2 \pi} \log \left|\phi\left(\operatorname{Re}^{\mathrm{i} \theta}\right)\right| \mathrm{d} \theta \tag{98}
\end{equation*}
$$

From (97) we have, $|\phi(0)|=|f(0)| \frac{\mathrm{R}}{\left|\mathrm{a}_{1}\right|} \cdot \frac{\mathrm{R}}{\left|\mathrm{a}_{2}\right|} \cdots \cdots \cdot \frac{\mathrm{R}}{\left|\mathrm{a}_{\mathrm{n}}\right|}$.
Hence from (98) we get,
$\log |f(0)|+\sum_{\mathrm{k}=1}^{\mathrm{n}} \log \left(\frac{\mathrm{R}}{\left|\mathrm{a}_{\mathrm{k}}\right|}\right)=\frac{1}{2 \pi} \int_{0}^{2 \pi} \log \left|\phi\left(\mathrm{Re}^{\mathrm{i} \theta}\right)\right| \mathrm{d} \theta$
i.e. $\log |f(0)|=\frac{1}{2 \pi} \int_{0}^{2 \pi} \log \left|f\left(\operatorname{Re}^{i \theta}\right)\right| d \theta-\sum_{k=1}^{n} \log \left(\frac{R}{\left|a_{k}\right|}\right)$
(since $|\phi(\mathrm{z})|=|\mathrm{f}(\mathrm{z})|$ on $|\mathrm{z}|=\mathrm{R}$ )
Note : We observe that Jensen's formula can also be expressed as

$$
\begin{equation*}
\log \frac{\mathrm{R}^{\mathrm{n}}}{\left|\mathrm{a}_{1} \ldots \mathrm{a}_{\mathrm{n}}\right|}=\frac{1}{2 \pi} \int_{0}^{2 \pi} \log \left|\mathrm{f}\left(\mathrm{Re}^{\mathrm{i} \theta}\right)\right| \mathrm{d} \theta-\log |\mathrm{f}(0)| \ldots \ldots \tag{99}
\end{equation*}
$$

or as, $\left.\quad \log \frac{R^{n}}{r_{1} \ldots r_{n}}=\frac{1}{2 \pi} \int_{0}^{2 \pi} \log \left|f\left(\operatorname{Re}^{\mathrm{i} \theta}\right) \mathrm{d} \theta-\log \right| \mathrm{f}(0) \right\rvert\, \ldots \ldots$
where

$$
\begin{equation*}
\left|\mathrm{a}_{\mathrm{i}}\right|=\mathrm{r}_{\mathrm{i}}, \mathrm{i}=1, \ldots, \mathrm{n} . \tag{100}
\end{equation*}
$$

Theorem 6.9 (Jensen's inequality) :- Let $f(z)$ be analytic on $|z| \leq R, f(0) \neq 0$ and $f(z) \neq 0$ on $|z|=R$. If $a_{1}, \ldots, a_{n}$ be the zeros of $f(z)$ within $|z|=R$, multiple zeros being repeated according to their multiplicities, and $\left|a_{i}\right|=r_{i}, i=1, \ldots, n$, then

$$
\begin{equation*}
\frac{\mathrm{R}^{\mathrm{n}}|\mathrm{f}(0)|}{\mathrm{r}_{1} \ldots \mathrm{r}_{\mathrm{n}}} \leq \mathrm{M}(\mathrm{R}) \tag{101}
\end{equation*}
$$

where

$$
\mathrm{M}(\mathrm{R})=\max _{|z|=\mathrm{R}}|\mathrm{f}(\mathrm{z})| .
$$

Proof : As in Jensen's formula (theorem 6.8) we have, $|\phi(\mathrm{z})|=|\mathrm{f}(\mathrm{z})|$ on $|\mathrm{z}|=\mathrm{R}$ and so by the maximum modulus theorem, $|\phi(\mathrm{z})| \leq \mathrm{M}(\mathrm{R})$ for $|\mathrm{z}| \leq \mathrm{R}$. In particular,

$$
|\phi(0)| \leq \mathrm{M}(\mathrm{R})
$$

$$
\text { i.e. } \frac{R^{n}|f(0)|}{r_{1} \ldots r_{n}} \leq M(R) \text {. }
$$

Theorem 6.10 (Poisson-Jensen formula) :- Let $f(z)$ be analytic on $|z| \leq R, f(0) \neq 0$ and $f(z) \neq 0$ on $|z|=R$. If $a_{1} \ldots a_{n}$ be the zeros of $f(z)$ within the circle $|z|=R$, multiple zeros being repeated according to the their multiplicities, then for any $z=r e^{i \theta}, r<R$,

$$
\log \left|f\left(\mathrm{re}^{\mathrm{i} \theta}\right)\right|=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\mathrm{R}^{2}-\mathrm{r}^{2}}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \operatorname{Rr} \cos (\mathrm{t}-\theta)} \log \left|\mathrm{f}\left(\mathrm{Re}^{\mathrm{it}}\right)\right| \mathrm{dt}-\sum_{\mathrm{k}=1}^{\mathrm{n}} \log \left|\frac{\mathrm{R}^{2}-\overline{\mathrm{a}}_{\mathrm{k}} \mathrm{re}}{\mathrm{i} \mathrm{\theta}} \underset{\mathrm{R}\left(\mathrm{re}^{\mathrm{i} \theta}-\mathrm{a}_{\mathrm{k}}\right)}{ }\right| .
$$

Proof : Let $\phi(z)=f(z) . \prod_{k=1}^{n} \frac{R^{2}-\bar{a}_{k} z}{R\left(z-a_{k}\right)}$. Then, as in Jensen's formula we have, $|\phi(z)|$ $=|f(z)|$ on $|z|=R$. Since $\phi(z)$ is analytic and non-zero on $|z| \leq R, \log \phi(z)$ is also analytic on $|z| \leq R$ and consequently $\log |\phi(z)|$ is harmonic on $|z| \leq R$.

So, by Poisson's integral formula,

$$
\begin{equation*}
\log \left|\phi\left(\mathrm{re}^{\mathrm{i} \theta}\right)\right|=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\mathrm{R}^{2}-\mathrm{r}^{2}}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \mathrm{Rr} \cos (\mathrm{t}-\theta)} \log \left|\phi\left(\mathrm{Re}^{\mathrm{it}}\right)\right| \mathrm{dt} \tag{102}
\end{equation*}
$$

Now, $\log \left|\phi\left(\mathrm{re}^{\mathrm{i} \theta}\right)\right|=\log \left|\mathrm{f}\left(\mathrm{re}^{\mathrm{i} \theta}\right)\right|+\sum_{\mathrm{k}=1}^{\mathrm{n}} \log \left|\frac{\mathrm{R}^{2}-\overline{\mathrm{a}}_{\mathrm{k}} \mathrm{re}^{\mathrm{i} \theta}}{\mathrm{R}\left(\mathrm{re}^{\mathrm{i} \theta}-\mathrm{a}_{\mathrm{k}}\right)}\right|$
Since $\log |\phi(\mathrm{z})|=\log |\mathrm{f}(\mathrm{z})|$ on $|\mathrm{z}|=\mathrm{R}$ we get from (102)

$$
\begin{align*}
\log \left|f\left(\mathrm{re}^{\mathrm{i} \theta}\right)\right|= & \frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\mathrm{R}^{2}-\mathrm{r}^{2}}{\mathrm{R}^{2}+\mathrm{r}^{2}-2 \operatorname{Rr} \cos (\mathrm{t}-\theta)} \cdot \log \left|\mathrm{f}\left(\mathrm{Re}^{\mathrm{it}}\right)\right| \mathrm{dt} \\
& -\sum_{\mathrm{k}=1}^{\mathrm{n}} \log \left|\frac{\mathrm{R}^{2}-\overline{\mathrm{a}}_{\mathrm{k}} \mathrm{re}^{\mathrm{i} \theta}}{\mathrm{R}\left(\mathrm{re}^{\mathrm{i} \theta}-\mathrm{a}_{\mathrm{k}}\right)}\right| \tag{103}
\end{align*}
$$

### 6.6 Convex functions

The property of convexity plays an important role in function theory because in several cases some lead factors associated with entire, meromorphic and subharmonic functions appear to be convex functions.

A real-valued function $\phi$ defined on the interval $\mathrm{I}=[\mathrm{a}, \mathrm{b}]$ is said to be convex if for any two points s , u in $[\mathrm{a}, \mathrm{b}]$

$$
\begin{equation*}
\phi(\lambda u+(1-\lambda) s \leq \lambda \phi(u)+(1-\lambda) \phi(s) \text { for } 0 \leq \lambda \leq 1 \tag{104}
\end{equation*}
$$

Geometrically, the condition (104) is equivalent to the condition that if $s<x<u$, then the point ( $\mathrm{x}, \phi(\mathrm{x})$ ) should lie below or on the chord joining the points ( $\mathrm{s}, \phi(\mathrm{s})$ ) and ( $\mathrm{u}, \phi(\mathrm{u})$ ) in the plane.

Analytical condition for $\phi(\mathbf{x})$ to be convex in [a, b]:- Let the coordinates of the points $\mathrm{A}, \mathrm{B}, \mathrm{C}$ on the curve $\mathrm{y}=\phi(\mathrm{x})$ as shown in the adjoining figure be $(\mathrm{s}, \phi(\mathrm{s})),(\mathrm{u}, \phi(\mathrm{u}))$ and ( $\mathrm{x}, \phi(\mathrm{x})$ ) respectively where $\mathrm{s}<\mathrm{x}<\mathrm{u}$.

Equation of the chord $A B$ is $y-\phi(x)=\frac{\phi(u)-\phi(s)}{u-s}(x-s)$.

$$
\begin{equation*}
\text { or, } y=\phi(s)+\frac{\phi(u)-\phi(s)}{u-s}(x-s) \tag{105}
\end{equation*}
$$

Let the coordinates of any point D on the chord AB be ( $\mathrm{x}, \mathrm{y}$ ). According to definition $\phi(\mathrm{x})$ will be convex if and only if $\mathrm{CN} \leq$ DN. i.e., if and only if $\phi(x) \leq y$; i.e. if and only if
$\phi(x) \leq \phi(s)+\frac{\phi(u)-\phi(s)}{u-s}(x-s)$; i.e., if and only if
$\phi(x) \leq \frac{u-x}{u-s} \phi(s)+\frac{x-s}{u-s} \phi(u)$

for $\mathrm{s}<\mathrm{x}<\mathrm{u}$.
We now state two results on convex functions without proof.
Result 1. A differentiable function $f(x)$ on $[a, b]$ is convex if and only if $f^{\prime}(x)$ is increasing in $[\mathrm{a}, \mathrm{b}]$.

Result 2. A sufficient condition for $f(x)$ to be convex is that $f^{\prime \prime}(x)>0$.
The maximum modulus function : Let $f(z)$ be a non-constant analytic function in $|z|$ $<\mathrm{R}$. Then for $0 \leq \mathrm{r}<\mathrm{R}$ we define the maximum modulus function $\mathrm{M}(\mathrm{r}, \mathrm{f})$ or, simply $\mathrm{M}(\mathrm{r})$ by $\mathrm{M}(\mathrm{r})=\max _{|\mathrm{z}|=\mathrm{r}}|\mathrm{f}(\mathrm{z})|$. By maximum modulus theorem we can also write $M(r)=\max _{|z|=r}|f(z)|$.

Result : Let $\mathrm{f}(\mathrm{z})$ be a non-constant analytic function in $|\mathrm{z}|<\mathrm{R}$. Then $\mathrm{M}(\mathrm{r})$ is a strictly increasing function of $r$ in $0 \leq r \leq R$.

Proof : Let $0 \leq r_{1}<r_{2}<R$. Since $f(z)$ is analytic in $|z| \leq r_{2}$, the maximum value of $|f(z)|$ for $|z| \leq r_{2}$ is attained on $|z|=r_{2}$. Let $z_{2}$ be a point on $|z|=r_{2}$ such that $\left|f\left(z_{2}\right)\right|$ $=M\left(r_{2}\right)$. Similarly, the maximum value of $|f(z)|$ for $|z| \leq r_{1}$ is attained on $|z|=r_{1}$. Let $z_{1}$ be a point on $|z|=r_{1}$ such that $\left|f\left(z_{1}\right)\right|=M\left(r_{1}\right)$.

Since $r_{1}<r_{2}, z_{1}$ is an interior point of the closed region $|z| \leq r_{2}$. Hence by maximum modulus theorem,
$\left|f\left(\mathrm{z}_{1}\right)\right|<\mathrm{M}\left(\mathrm{r}_{2}\right)$; i.e. $\mathrm{M}\left(\mathrm{r}_{1}\right)<\mathrm{M}\left(\mathrm{r}_{2}\right)$.
This proves the result.

Corollary : Let $\mathrm{f}(\mathrm{z})$ be a non-constant entire function. Then its maximum modulus function $\mathrm{M}(\mathrm{r}) \rightarrow \infty$ as $|\mathrm{z}|=\mathrm{r} \rightarrow \infty$. For, if $\mathrm{M}(\mathrm{r})$ is bounded, then by Liouville's theorem $\mathrm{f}(\mathrm{z})$ would be a constant function.

Theorem 6.11 [Hadamard's three-circles theorem].
Let $0<r_{1}<r<r_{3}$ and suppose that $f(z)$ is analytic on the closed annulus $r_{1} \leq|z| \leq$ $r_{3}$. If $M(r)=\max _{|z|=r}|f(z)|$, then

$$
\begin{equation*}
M(r)^{\log \left(\frac{r_{3}}{r_{1}}\right)} \leq M\left(r_{1}\right)^{\log \left(\frac{r_{3}}{r_{1}}\right)} \cdot M\left(r_{3}\right)^{\log \left(\frac{r}{r_{1}}\right)} \tag{107}
\end{equation*}
$$

Proof : Let us consider the function $\phi(\mathrm{z})=\mathrm{z}^{\alpha} \mathrm{f}(\mathrm{z})$, where $\alpha$ is a real constant to be chosen later. If $\alpha \neq$ an integer, $\phi(\mathrm{z})$ is multi-valued in $\mathrm{r}_{1} \leq|\mathrm{z}| \leq \mathrm{r}_{3}$ and so we cut the annulus along the negative part of the real axis. Thus we obtain a simply connected region G in which the principal branch of $\phi(\mathrm{z})$ is analytic. Hence the maximum modulus of this branch of $\phi(\mathrm{z})$ in G is attained on the boundary of G . Since $\alpha$ is real, all the branches of $\phi(\mathrm{z})$ have the same modulus. If we consider another branch of $\phi(\mathrm{z})$ which is analytic in another cut annulus it is clear that the principal branch of $\phi(z)$ can not attain
 its maximum value on the cut. Hence maximum of $|\phi(z)|$ is attained on at least one of the bounding circles $|z|=r_{1}$ or, $|z|=r_{3}$. Thus,

$$
\begin{gather*}
\left|\mathrm{z}^{\alpha} \mathrm{f}(\mathrm{z})\right| \leq \max \left(\mathrm{r}_{1}^{\alpha} \mathrm{M}\left(\mathrm{r}_{1}\right), \mathrm{r}_{3}^{\alpha} \mathrm{M}\left(\mathrm{r}_{3}\right)\right) . \text { Hence on }|\mathrm{z}|=\mathrm{r} \text {, } \\
\mathrm{r}^{\alpha} \mathrm{M}(\mathrm{r}) \leq \max \left(\mathrm{r}_{1}^{\alpha} \mathrm{M}\left(\mathrm{r}_{1}\right), \mathrm{r}_{3}^{\alpha} \mathrm{M}\left(\mathrm{r}_{3}\right)\right) \tag{108}
\end{gather*}
$$

We now choose $\alpha$ such that $r_{1}^{\alpha} M\left(r_{1}\right)=r_{3}^{\alpha} M\left(r_{3}\right)$. Then
$\alpha=-\frac{\left.\log \left(\mathrm{M}\left(\mathrm{r}_{3}\right)\right) / \mathrm{M}\left(\mathrm{r}_{1}\right)\right)}{\log \left(\mathrm{r}_{3} / \mathrm{r}_{1}\right)}$. Substituting this value of $\alpha$ in (108) we get,

$$
\begin{aligned}
& M(r) \leq\left(\frac{r}{r_{1}}\right)^{-\alpha} M\left(r_{1}\right) \\
& =\left(\frac{r}{r_{1}}\right)^{\log \left(\frac{M\left(r_{3}\right)}{M\left(r_{1}\right)}\right)} / \log \left(\frac{r_{3}}{r_{1}}\right) \cdot M\left(r_{1}\right)
\end{aligned}
$$

and so

$$
M(r)^{\log \left(r_{3} / r_{1}\right)} \leq\left(\frac{r}{r_{1}}\right)^{\log \left(M\left(r_{3}\right) / M\left(r_{1}\right)\right)} \cdot M\left(r_{1}\right)^{\log \left(r_{3} / r_{1}\right)}
$$

That is, $\quad M(r)^{\log \left(r_{3} / r_{1}\right)} \leq\left(\frac{M\left(r_{3}\right)}{M\left(r_{1}\right)}\right)^{\log \left(r / r_{1}\right)} \cdot M\left(r_{1}\right)^{\log \left(r_{3} / r_{1}\right)}\left[\right.$ since $\left.a^{\log b}=b^{\log a}\right]$

$$
=\mathrm{M}\left(\mathrm{r}_{1}\right)^{\log \left(\mathrm{r}_{3} / \mathrm{r}\right)} \cdot \mathrm{M}\left(\mathrm{r}_{3}\right)^{\log \left(\mathrm{r} / \mathrm{r}_{1}\right)} .
$$

Note : Equality in (107) occurs when $\phi(\mathrm{z})$ is a constant, i.e. when $\mathrm{f}(\mathrm{z})$ is of the form $\mathrm{cz}^{\alpha}$ for some real $\alpha$ and c is a constant.

Corollary : $\log \mathrm{M}(\mathrm{r})$ is a convex function of $\log \mathrm{r}$.
Proof : Let $\mathrm{f}(\mathrm{z})$ be analytic in the closed annulus $0<\mathrm{r}_{1} \leq|\mathrm{z}| \leq \mathrm{r}_{2}$.
If $r_{1}<r<r_{2}$ we have, by Hadamard's three-circles theorem,
$\mathrm{M}(\mathrm{r})^{\log \left(r_{2} / r_{1}\right)} \leq \mathrm{M}\left(\mathrm{r}_{1}\right)^{\log \left(\mathrm{r}_{2} / \mathrm{r}\right)} \cdot \mathrm{M}\left(\mathrm{r}_{2}\right)^{\log \left(\mathrm{r} / \mathrm{r}_{1}\right)}$. Taking logarithms we get
$\left(\log r_{2}-\log r_{1}\right) \log M(r) \leq\left(\log r_{2}-\log r\right) \log M\left(r_{1}\right)+$
$\left(\log r-\log r_{1}\right) \log M\left(r_{2}\right)$. That is,
$\log M(r) \leq \frac{\log r_{2}-\log r}{\log r_{2}-\log r_{1}} \log M\left(r_{1}\right)+\frac{\log r-\log r_{1}}{\log r_{2}-\log r_{1}} \log M\left(r_{2}\right)$
The inequality (109) shows that $\log \mathrm{M}(\mathrm{r})$ is a convex function of $\log \mathrm{r}$.

### 6.7 Order of an entire function

An entire function $f(z)$ is said to be of finite order if there is a positive number A such that as $|z|=r \rightarrow \infty$, the inequality $\mathrm{M}(\mathrm{r})<\mathrm{e}^{\mathrm{A}}$ holds.

The lower bound $\rho$ of such numbers A is called the order of the function.
f is said to be of infinite order if it is not of finite order. From the definition it is clear that order of an entire function is non-negative.

Result : Let f be an entire function of order $\rho$ and $\mathrm{M}(\mathrm{r})=\max \{|\mathrm{f}(\mathrm{z})|:|\mathrm{z}|=\mathrm{r}\}$. Then

$$
\begin{equation*}
\rho=\lim _{r \rightarrow \infty} \sup \frac{\log \log M(r)}{\log r} \tag{110}
\end{equation*}
$$

Proof: By hypothesis, given $\varepsilon>0$ there exists $\mathrm{r}_{0}(\varepsilon)>0$ such that

$$
\mathrm{M}(\mathrm{r})<\mathrm{e}^{\mathrm{r}+\varepsilon} \text { for } \mathrm{r}>\mathrm{r}_{0}
$$

while $M(r)>e^{\mathrm{r}^{\mathrm{r}+\varepsilon}}$ for an increasing sequence $\left\{\mathrm{r}_{\mathrm{n}}\right\}$ of values of r , tending to infinity. In otherwords,

$$
\begin{align*}
& \frac{\log \log \mathrm{M}(\mathrm{r})}{\log r}<\rho+\varepsilon \quad \forall r>r_{0} \text { and }  \tag{111}\\
& \frac{\log \log \mathrm{M}(\mathrm{r})}{\log r}>\rho-\varepsilon \tag{112}
\end{align*}
$$

for a sequence of values of $\mathrm{r} \rightarrow+\infty$
(111) and (112) precisely means

$$
\rho=\limsup _{\mathrm{r} \rightarrow \infty} \frac{\log \log M(\mathrm{r})}{\log \mathrm{r}}
$$

Example 3 : Determine the order of the functions.
(i) $p(z)=a_{0}+a_{1} z+\ldots+a_{n} z^{n}, a_{n} \neq 0$. (ii) $e^{k z}, k \neq 0$.
(iii) $\sin \mathrm{z}$ (iv) $\cos \sqrt{\mathrm{Z}}$

Solution :
(i) $|p(z)|=\left|a_{0}+a_{1} z+\ldots+a_{n} z^{n}\right| \leq\left|a_{0}\right|+\left|a_{1}\right||z|+\ldots+\left|a_{n}\right||z|^{n}$

Hence, $M(r)=\max _{|z|=r}|p(z)| \leq\left|a_{0}\right|+\left|a_{1}\right| r+\ldots+\left|a_{n}\right| r^{n}$
$\leq r^{n}\left(\left|a_{0}\right|+\ldots+\left|a_{n}\right|\right)$ (choosing $r \geq 1$. Since ultimately $r \rightarrow \infty$, the choice is justified).
$=\mathrm{Br}^{\mathrm{n}}$, where $\mathrm{B}=\left|\mathrm{a}_{0}\right|+\ldots+\left|\mathrm{a}_{\mathrm{n}}\right|$. Hence
$\log \mathrm{M}(\mathrm{r}) \leq \log \mathrm{B}+\mathrm{n} \log \mathrm{r} \leq \log r+n \log r$ (Taking $r$ sufficiently large).
$=(n+1) \log r$. Now,
$\rho=\underset{r \rightarrow \infty}{\limsup } \frac{\log \log M(r)}{\log r} \leq \operatorname{limssup}_{r \rightarrow \infty} \frac{\log (n+1)+\log \log r}{\log r}=0$
i.e. $\rho \leq 0$. But by definition $\rho \geq 0$. Hence $\rho=0$
(ii) Here $\mathrm{M}(\mathrm{r})=\mathrm{e}^{\mathrm{k} \mid \mathrm{r}}$ and hence
$\rho=\underset{r \rightarrow \infty}{\limsup } \frac{\log \log M(r)}{\log r}=\underset{r \rightarrow \infty}{\limsup } \frac{\log (|k| r)}{\log r}=1$
(iii) We know that

$$
\sin \mathrm{z}=\mathrm{z}-\frac{\mathrm{z}^{3}}{3!}+\frac{\mathrm{z}^{5}}{5!}-\cdots
$$

and so

$$
\begin{aligned}
& \qquad|\sin \mathrm{z}| \leq|\mathrm{z}|+\frac{|\mathrm{z}|^{3}}{3!}+\frac{|\mathrm{z}|^{5}}{5!}+\cdots=\mathrm{r}+\frac{\mathrm{r}^{3}}{3!}+\frac{\mathrm{r}^{5}}{5!}+\cdots=\sinh \mathrm{r} \text { on }|\mathrm{z}| \leq \mathrm{r} . \\
& =\frac{\mathrm{e}^{\mathrm{r}}-\mathrm{e}^{-\mathrm{r}}}{2} \text {. Also at } \mathrm{z}=\operatorname{ir}, \sin \mathrm{z}=\frac{\mathrm{e}^{-r}-\mathrm{e}^{\mathrm{r}}}{2 \mathrm{i}} \text { and so }|\sin \mathrm{z}|=\frac{\mathrm{e}^{\mathrm{r}}-\mathrm{e}^{-r}}{2} . \\
& \text { Hence } \mathrm{M}(\mathrm{r})=\frac{\mathrm{e}^{\mathrm{r}}-\mathrm{e}^{-\mathrm{r}}}{2}=\frac{\mathrm{e}^{\mathrm{r}}\left(1-\mathrm{e}^{-2 \mathrm{r}}\right)}{2} \\
& \log \mathrm{M}(\mathrm{r})=\mathrm{r}+\log \left(\frac{1-\mathrm{e}^{-2 \mathrm{r}}}{2}\right)=\mathrm{r}\left\{1+\frac{1}{\mathrm{r}} \log \left(\frac{1-\mathrm{e}^{-2 \mathrm{r}}}{2}\right)\right\}
\end{aligned}
$$

Therefore,

$$
\lim _{r \rightarrow \infty} \frac{\log \log M(r)}{\log r}=\lim _{r \rightarrow \infty}\left[1+\log \left\{1+\frac{1}{r} \log \left(\frac{1-e^{-2 r}}{2}\right)\right\} / \log r\right]=1
$$

So order of $\sin \mathrm{z}$ is 1 .
(iv) Following as in (iii) we find that the order of $\cos \sqrt{z}=1 / 2$.

Let $\mathrm{f}(\mathrm{z})=\sum_{\mathrm{n}=0}^{\infty} \mathrm{a}_{\mathrm{n}} \mathrm{z}^{\mathrm{n}}$ be an entire function. We now state a theorem which will give us order of $f(z)$ in terms of the coefficients $a_{n}$ of the power series expansion of $f(z)$.

Theorem : Let $f(z)=\sum_{n=0}^{\infty} a_{n} z^{n}$ be an entire function of finite order $\rho$. Then,

$$
\rho=\limsup _{n \rightarrow \infty} \frac{-\log n}{\log \left|a_{n}\right|^{1 / n}}=\limsup _{n \rightarrow \infty} \frac{-n \log n}{\log \left|a_{n}\right|}
$$

### 6.8 The function $\mathbf{n}(\mathbf{r})$

Let $\mathrm{f}(\mathrm{z})$ be an entire function with zeros at the points $\mathrm{a}_{1}, \mathrm{a}_{2}, \ldots$, arranged in order of non-decreasing modulus, i.e. $\left|\mathrm{a}_{1}\right| \leq\left|\mathrm{a}_{2}\right| \leq \cdots$, multiple zeros being repeated according to
their multiplicities. We define the function $n(r)$ to be the number of zeros of $f(z)$ in $|z| \leq r$. Evidently $n(r)$ is a non-decreasing, non-negative function of $r$ which is constant in any interval which does not contain the modulus of a zero of $f(z)$. We observe that if $f(0) \neq$ 0. $n(r)=0$ for $r<\left|a_{1}\right|$. Also, $n(r)=n$ for $\left|a_{n}\right| \leq r<\left|a_{n+1}\right|$.

Jensen's inequality can also be written in the following form involving $n(r)$.
Theorem 6.12 (Jensen's inequality) : Let $\mathrm{f}(\mathrm{z})$ be an entire function with $\mathrm{f}(0) \neq$ 0 , and $a_{1}, a_{2}, \ldots$ be the zeros of $f(z)$ such that $\left|a_{1}\right| \leq\left|a_{2}\right| \leq \cdots$, multiple zeros being repeated according to their multiplicities. If $\left|a_{N}\right| \leq r<\left|a_{N+1}\right|$, then

$$
\begin{equation*}
\log \frac{r^{N}}{\left|a_{1} \cdots \cdots a_{N}\right|}=\int_{0}^{r} \frac{n(x)}{x} d x \leq \log M(r)-\log |f(0)| \tag{113}
\end{equation*}
$$

Proof : Let $\left|a_{i}\right|=r_{i}, i=1,2, \ldots$, and $r$ be a positive number such that $r_{N} \leq r<r_{N+1}$. Let $x_{1} \ldots, x_{m}$ be the distinct numbers of the set $A=\left\{r_{1}, \ldots, r_{N}\right\}$ where $x_{1}=r_{1}, \ldots, x_{m}$ $=r_{N}$. Suppose $x_{i}$ is repeated $p_{i}$ times in A. Then, $p_{1}+\ldots+p_{m}=N$. Also let $t_{i}=p_{1}$ $+\ldots+p_{i}, i=1, \ldots, m$.

We now consider two cases.
Case 1) Let $\mathrm{r}_{\mathrm{N}}<\mathrm{r}$. Then,

$$
\begin{aligned}
& \int_{0}^{r} \frac{n(x)}{x} d x=\lim _{\varepsilon \rightarrow 0}\left\{\int_{x_{1}}^{x_{2}-\varepsilon} \frac{n(x)}{x} d x+\int_{x_{2}}^{x_{3}-\varepsilon} \frac{n(x)}{x} d x+\ldots+\int_{x_{m-1}}^{x_{m}-\varepsilon} \frac{n(x)}{x} d x\right\}+\int_{x_{m}}^{r} \frac{n(x)}{x} d x \\
& \text { (since } \int_{0}^{x_{1}-\varepsilon} \frac{n(x)}{x} d x=0 \text { as } n(x)=0 \text { for } 0 \leq x<x_{1} \text { ). } \\
& =\lim _{\varepsilon \rightarrow 0}\left\{\int_{x_{1}}^{x_{2}-\varepsilon} \frac{t_{1}}{x} d x+\int_{x_{2}}^{x_{3}-\varepsilon} \frac{t_{2}}{x} d x+\cdots+\int_{x_{m-1}}^{x_{m}-\varepsilon} \frac{t_{m-1}}{x} d x\right\}+\int_{T_{N}}^{t} \frac{N}{x} d x \\
& =\lim _{\varepsilon \rightarrow 0}\left\{\left[t_{1} \log x\right\}_{x_{1}}^{x_{2}-\varepsilon}+\left[t_{2} \log x\right]_{x_{1}}^{x_{1}-\varepsilon}+\cdots+\left[t_{m-1} \log x\right]_{x_{m-1}}^{x_{m}-\varepsilon}+[N \log x]_{r_{N}}^{r}\right. \\
& =\lim _{\varepsilon \rightarrow 0}\left[\mathrm{t}_{1}\left\{\log \left(\mathrm{x}_{2}-\varepsilon\right)-\log \mathrm{X}_{1}\right\}+\mathrm{t}_{2}\left\{\log \left(\mathrm{x}_{3}-\varepsilon\right)-\log \mathrm{x}_{2}\right\}+\right. \\
& \left.\cdots+\mathrm{t}_{\mathrm{m}-1}\left\{\log \left(\mathrm{x}_{\mathrm{m}}-\varepsilon\right)-\log \mathrm{x}_{\mathrm{m}-1}\right\}\right]+\mathrm{N}\left(\log \mathrm{r}-\log \mathrm{r}_{\mathrm{N}}\right) \\
& =\mathrm{t}_{1}\left(\log \mathrm{x}_{2}-\log \mathrm{x}_{1}\right)+\mathrm{t}_{2}\left(\log \mathrm{x}_{3}-\log \mathrm{x}_{2}\right)+\ldots \\
& +\mathrm{t}_{\mathrm{m}-1}\left(\log \mathrm{x}_{\mathrm{m}}-\log \mathrm{x}_{\mathrm{m}-1}\right)+\mathrm{N}\left(\log \mathrm{r}-\log \mathrm{r}_{\mathrm{N}}\right) \\
& =\mathrm{p}_{1} \log \mathrm{x}_{2}-\mathrm{p}_{1} \log \mathrm{x}_{1}+\left(\mathrm{p}_{1}+\mathrm{p}_{2}\right) \log \mathrm{x}_{1}-\left(\mathrm{p}_{1}+\mathrm{p}_{2}\right) \log \mathrm{x}_{2}+\ldots+\left(\mathrm{p}_{1}+\ldots+\mathrm{p}_{\mathrm{m}-1}\right) \\
& \log \mathrm{x}_{\mathrm{m}}-\left(\mathrm{p}_{1}+\ldots+\mathrm{p}_{\mathrm{m}-1}\right) \log \mathrm{x}_{\mathrm{m}-1}+\mathrm{N} \log \mathrm{r}-\left(\mathrm{p}_{1}+\ldots+\mathrm{p}_{\mathrm{m}}\right) \log \mathrm{x}_{\mathrm{m}} \\
& =N \log r-\left(p_{1} \log x_{1}+p_{2} \log \mathrm{x}_{2}+\ldots+\mathrm{p}_{\mathrm{m}} \log \mathrm{x}_{\mathrm{m}}\right)
\end{aligned}
$$

$=\log r^{N}-\log x_{1}^{p_{1}} x_{2}^{p_{2}} \cdots x_{m}^{p_{m}}=\log \frac{r^{N}}{x_{1}^{p_{1}} x_{2}^{p_{2}} \cdots x_{m}^{p_{m}}}$
$=\log \frac{\mathrm{r}^{\mathrm{N}}}{\mathrm{r}_{1} \cdots \mathrm{r}_{\mathrm{N}}}$ Thus,
$\int_{0}^{r} \frac{n(x)}{x} d x=\log \frac{r^{N}}{\left|a_{1} \cdots a_{N}\right|}$
Case 2). Let $r_{N}=r$. As before,
$\int_{0}^{r} \frac{n(x)}{x} d x=\lim _{\varepsilon \rightarrow 0}\left\{\int_{x_{1}}^{x_{2}-\varepsilon} \frac{t_{1}}{x} d x+\cdots+\int_{x_{m-1}}^{x_{m}-\varepsilon} \frac{t_{m-1}}{x} d x\right\}$
$=\sum_{i=1}^{m-1} t_{i}\left(\log x_{i+1}-\log x_{i}\right)+t_{m}\left(\log r-\log r_{N}\right)$
$=\log \frac{\mathrm{r}^{\mathrm{N}}}{\left|\mathrm{a}_{1} \cdots \mathrm{a}_{\mathrm{N}}\right|}$ (Proceeding as in case 1).
Thus in any case,
$\int_{0}^{\mathrm{r}} \frac{\mathrm{n}(\mathrm{x})}{\mathrm{x}} \mathrm{dx}=\log \frac{\mathrm{r}^{\mathrm{N}}}{\left|\mathrm{a}_{1} \cdots \mathrm{a}_{\mathrm{N}}\right|}$. But Jensen's inequality gives us
$\frac{\mathrm{r}^{\mathrm{N}}}{\left|\mathrm{a}_{1} \cdots \mathrm{a}_{\mathrm{N}}\right|} \leq \frac{\mathrm{M}(\mathrm{r})}{|\mathrm{f}(0)|}$. Hence,
$\int_{0}^{r} \frac{n(x)}{x} d x=\log \frac{r^{N}}{\left|a_{1} \cdots a_{N}\right|} \leq \log M(r)-\log |f(0)|$.
Theorem 6.13: If $f(z)$ be an entire function with finite order $\rho$, then $n(r)=O\left(r^{\rho+\varepsilon}\right)$ for $\varepsilon>0$ and for sufficiently large values of $r$.

Proof : By Jensen's inequalilty,

$$
\begin{equation*}
\int_{0}^{\mathrm{r}} \frac{\mathrm{n}(\mathrm{x})}{\mathrm{x}} \mathrm{dx} \leq \log \mathrm{M}(\mathrm{r})-\log |\mathrm{f}(0)| \tag{115}
\end{equation*}
$$

We replace $r$ by $2 r$ in (115) and obtain

$$
\begin{equation*}
\int_{0}^{2 r} \frac{\mathrm{n}(\mathrm{x})}{\mathrm{x}} \mathrm{dx} \leq \log \mathrm{M}(2 \mathrm{r})-\log |\mathrm{f}(0)| \tag{116}
\end{equation*}
$$

Since order of $f(z)$ is $\rho$ we have for any $\varepsilon>0$, $\log \mathrm{M}(2 \mathrm{r})<(2 \mathrm{r})^{\rho+\varepsilon}=\mathrm{Kr}^{\rho+\varepsilon}$ for all large $\mathrm{r}, \mathrm{K}$ being a constant. Hence from (116).

$$
\int_{0}^{2 r} \frac{n(x)}{x} d x<\operatorname{Ar}^{\rho+\varepsilon} \text { for all large } r \text {, A being a constant independent of } r \text {. Since } n(x)
$$ is non-negative and non-decreaing function of $x, \int_{r}^{2 r} \frac{n(x)}{x} d x \leq \int_{0}^{2 r} \frac{n(x)}{x} d x<$

$A r^{\rho+\varepsilon}$ and also $\int_{r}^{2 r} \frac{n(x)}{x} d x \geq \int_{r}^{2 r} \frac{n(r)}{x} d x=n(r) \log 2$
Hence, $n(r) \log 2 \leq \int_{r}^{2 r} \frac{n(x)}{x} d x<A r^{\rho+\varepsilon}$,
i.e., $n(r)<\frac{A}{\log 2} r^{\rho+\varepsilon}$ for all large r. Hence, $n(r)=O\left(r^{\rho+\varepsilon}\right)$.

### 6.9 Convergence exponent (or, exponent of Convergence)

Let $f(z)$ be an entire function with zeros at the points $a_{1}, a_{2}, \ldots$, arranged in order of non-decreasing modulus, multiple zeros being repeated according to their multiplicities and $\left|a_{i}\right|=r_{i}, i=1,2, \ldots$, We define convergence exponent $\rho_{1}$ of the zeros of $f(z)$ by the equation

$$
\begin{equation*}
\rho_{1}=\limsup _{n \rightarrow \infty} \frac{\log n}{\log r_{n}} \tag{117}
\end{equation*}
$$

or, equivalently by $\rho_{1}=\limsup _{n \rightarrow \infty} \frac{\log n(r)}{\log r}$
The convergence exponent has the following property.
Theorem 6.14: Let $f(z)$ be an entire function with zeros at $a_{1} a_{2}, \ldots$, arranged in order of non-decreasing modulus, multiple zeros being repeated according to their multiplicities and $\left|a_{i}\right|=r_{i}$. If the convergence exponent $\rho_{1}$ of the zeros of $f(z)$ is finite, then the series $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ converges when $\alpha>\rho_{1}$ and diverges when $\alpha<\rho_{1}$.

If $\rho_{1}$ is infinite, the above series diverges for all positive values of $\alpha$.
Proof : Let $\rho_{1}$ be finite and $\alpha>\rho_{1}$. Then, $\rho_{1}<\frac{1}{2}\left(\rho_{1}+\alpha\right)$.
Hence, $\frac{\log n}{\log r_{n}}<\frac{1}{2}\left(\rho_{1}+\alpha\right)$ for all large $n$.
or, $\log \mathrm{n}<\log _{\mathrm{n}} \frac{1}{2}^{\frac{1}{2}\left(\rho_{1}+\alpha\right)}$, i.e.
$\mathrm{n}<\mathrm{r}_{\mathrm{n}}^{\frac{1}{2}\left(\rho_{1}+\alpha\right)} ;$ or, $\frac{2}{\mathrm{n}^{\rho_{1}+\alpha}}<\mathrm{r}_{\mathrm{n}}$ i.e.,
$\mathrm{r}_{\mathrm{n}}^{\alpha}>\frac{2 \alpha}{\mathrm{n}^{\rho_{1}+\alpha}}=\mathrm{n}^{1+\frac{\alpha-\rho_{1}}{\alpha+\rho_{1}}}=\mathrm{n}^{1+\mathrm{p}}$, where $\mathrm{p}=\frac{\alpha-\rho_{1}}{\alpha+\rho_{1}}>0$.
Hence, $\frac{1}{\mathrm{r}_{\mathrm{n}}^{\alpha}}<\frac{1}{\mathrm{n}^{1+\mathrm{p}}}$ for all large n . Hence,
$\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}^{\alpha}}$ converges.
Next, let $\alpha<\rho_{1}$. Then, $\frac{\log n}{\log r_{n}}>\alpha$ for a sequence of values of $n$, tending to infinity. That is, $\log n>\log r_{n}^{\alpha}$

$$
\begin{equation*}
\text { or, } \frac{1}{\mathrm{r}_{\mathrm{n}}^{\alpha}}>\frac{1}{\mathrm{n}} \tag{119}
\end{equation*}
$$

for a sequence of values of $n$ tending to infinity.
Let N be such a value of n for which (119) holds and m be the least integer $>\frac{\mathrm{N}}{2}$. Then, as $r_{n}$ is non-decreasing,

$$
\sum_{n=N-m}^{N} \frac{1}{r_{n}^{\alpha}}=\frac{1}{r_{N-m}^{\alpha}}+\frac{1}{r_{N-m+1}^{\alpha}}+\cdots+\frac{1}{r_{N}^{\alpha}} \geq \frac{1}{r_{N}^{\alpha}}+\cdots+\frac{1}{r_{N}^{\alpha}}
$$

$=\frac{\mathrm{m}+1}{\mathrm{r}_{\mathrm{N}}^{\alpha}}>\frac{\mathrm{m}}{\mathrm{r}_{\mathrm{N}}^{\alpha}}>\frac{\mathrm{m}}{\mathrm{N}}>\frac{1}{2}$. Since N may be as large as we please, by Cauchy's principle of convergence, the series $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ diverges.

If $\rho_{1}$ is infinite, then for any positive value of $\alpha, \frac{\log n}{\log r_{n}}>\alpha$ for a sequence of values
of n tending to infinity; i.e., $\mathrm{n}>\mathrm{r}_{\mathrm{n}}^{\alpha}$ for a sequence of values of n tending to infinity. Hence as before, the series $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ diverges for any positive $\alpha$.

Note 1. Observe that $\rho_{1}$ may also be defined as the lower bound of the positive numbers $\alpha$ for which the series $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ is convergent. If $f(z)$ has no zeros we define $\rho_{1}=0$ and if $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ diverges for all positive $\alpha$, then $\rho_{1}=\infty$.

Note 2. If $\rho_{1}$ is finite, the series $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\rho_{1}}}$ may be convergent or divergent. For example, if $r_{n}=n$, then $\rho_{1}=\underset{n \rightarrow \infty}{\limsup } \frac{\log n}{\log r_{n}}=1$
and $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}}=\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{n}}$ diverges. Again, if $\mathrm{r}_{\mathrm{n}}=\mathrm{n}(\log \mathrm{n})^{2}$,
then, $\rho_{1}=\limsup _{n \rightarrow \infty} \frac{\log n}{\log n+2 \log \log n}=1$, and
$\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\rho_{1}}}=\sum_{n=1}^{\infty} \frac{1}{n(\log n)^{2}}$ converges.
Theorem 6.15: If $f(z)$ is an entire function with finite order $\rho$ and $r_{1}, r_{2}, \ldots$, are the moduli of the zeros of $f(z)$,
then $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}^{\alpha}}$ converges if $\alpha>\rho$.
Proof : We choose $\beta$ such that $\rho<\beta<\alpha$. Since for any $\varepsilon>0$,
$\mathrm{n}(\mathrm{r})=0\left(\mathrm{r}^{\rho+\varepsilon}\right), \mathrm{n}(\mathrm{r})<\mathrm{Kr}^{\beta}$
for all large r , K being a constant.
Putting $r=r_{n}$, $n$ large, (120) gives $n<\operatorname{Kr}_{n}^{\beta}$, i.e.,
$r_{n}>\frac{n^{1 / \beta}}{k^{1 / \beta}}$ or, $\frac{1}{r_{n}^{\alpha}}<\frac{B}{n^{\alpha / \beta}}$ for all large $n$, B being a constant. Since $\frac{\alpha}{\beta}>1, \sum_{n=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ converges.

Corollary : Since convergence exponent $\rho_{1}$ is the lower bound of positive numbers $\alpha$ for which $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ is convergent, it follows that $\rho_{1} \leq \rho$.

Note : $\rho_{1}$ may be 0 or $\infty$. For example if $r_{n}=e^{n}, \rho_{1}=0$ and if $r_{n}=\log n$, then $\rho_{1}=\infty$. For the function $f(z)=e^{z}, \rho=1$ and $\rho_{1}=0$ so that $\rho_{1}<\rho$. But for $\sin z$ or $\cos \mathrm{z}, \rho=\rho_{1}=1$.

Result : If the convergence exponent $\rho_{1}$ of the zeros of an entire function $f(z)$ is greater than 0 , then $f(z)$ has infinite number of zeros.

Proof : If possible, suppose $f(z)$ has finite number of zeros with moduli $r_{1} \ldots, r_{N}$. The series $\sum_{n=1}^{N} \frac{1}{r_{n}^{\alpha}}$, being of finite number of terms, converges for every $\alpha>0$. Hence $\rho_{1}=0$, a contradiction. Hence $f(z)$ has infinite number of zeros.

Note : For an entire function with finite number of zeros, $\rho_{1}=0$.
Example : Find the convergence exponent of the zeros of $\cos z$.
Solution : First method : The zeros of $\cos \mathrm{z}$ are $\frac{\pi}{2},-\frac{\pi}{2}, \frac{3 \pi}{2},-\frac{3 \pi}{2}, \ldots$.
Now, $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}^{\alpha}}=\left(\frac{2}{\pi}\right)^{\alpha}+\left(\frac{2}{\pi}\right)^{\alpha}+\left(\frac{2}{\pi}\right)^{\alpha} \cdot \frac{1}{3^{\alpha}}+\cdots$
$=2\left(\frac{2}{\pi}\right)^{\alpha}\left(1+\frac{1}{3^{\alpha}}+\frac{1}{5^{\alpha}}+\cdots\right)$. The series $\frac{1}{1^{\alpha}}+\frac{1}{3^{\alpha}}+\frac{1}{5^{\alpha}}+\cdots$
converges when $\alpha>1$ and diverges when $\alpha<1$. Hence the lower bound of the positive numbers $\alpha$ for which $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\alpha}}$ converges is 1 i.e., $\rho_{1}=1$.

Second method : The zeros of $\cos \mathrm{z}$ are $(2 \mathrm{n}+1) \frac{\pi}{2}$,

$$
\mathrm{n}=0, \pm 1, \pm 2, \cdots \text {; i.e. } \frac{\pi}{2},-\frac{\pi}{2}, \frac{3 \pi}{2},-\frac{3 \pi}{2}, \cdots
$$

Let $a_{1}=\frac{\pi}{2}, a_{1}^{\prime}=-\frac{\pi}{2}, a_{2}=\frac{3 \pi}{2}, a_{2}^{\prime}=-\frac{3 \pi}{2}, \cdots$,
$\mathrm{a}_{\mathrm{n}}=(2 \mathrm{n}-1) \frac{\pi}{2}, \mathrm{a}_{\mathrm{n}}^{\prime}=-(2 \mathrm{n}-1) \frac{\pi}{2}, \cdots$, Hence,
$r_{1}=\left|a_{1}\right|=\left|a_{1}^{\prime}\right|=\frac{\pi}{2}, r_{2}=\left|a_{2}\right|=\left|a_{2}^{\prime}\right|=\frac{3 \pi}{2}, \cdots, r_{n}=\left|a_{n}\right|=\left|a_{n}^{\prime}\right|=$

$$
\begin{aligned}
& (2 n-1) \frac{\pi}{2}, \cdots \text { Hence, } \rho_{1}=\limsup _{n \rightarrow \infty} \frac{\log n}{\log r_{n}} \\
& =\underset{n \rightarrow \infty}{\limsup } \frac{\log n}{\log (2 n-1)+\log \frac{\pi}{2}}=\limsup _{n \rightarrow \infty} \frac{\log n}{\log \left\{n\left(2-\frac{1}{n}\right)\right\}+\log \frac{\pi}{2}} \\
& =\underset{n \rightarrow \infty}{\lim \sup } \frac{1}{1+\frac{\log \left(2-\frac{1}{n}\right)}{\log n}+\frac{\log \pi / 2}{\log n}}=1 .
\end{aligned}
$$

### 6.10 Canonical Product

Let $\mathrm{f}(\mathrm{z})$ be an entire function with infinite number of zeros at $\mathrm{a}_{\mathrm{n}}, \mathrm{n}=1,2, \ldots \mathrm{a}_{\mathrm{n}} \neq 0$. If there exists a least non-negative integer p such that the series $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}^{\mathrm{p}+1}}$ is convergent, where $r_{n}=\left|a_{n}\right|$, we form the infinite product $G(z)=\prod_{n=1}^{\infty} E\left(\frac{z}{a_{n}}, p\right)$. By Weirstrass' factor theorem $\mathrm{G}(\mathrm{z})$ represents an entire function having zeros precisely at the points $\mathrm{a}_{n}$. We call $G(z)$ as the Canonical product corresponding to the sequence $\left\{a_{n}\right\}$ and the integer $p$ is called its genus. If $\mathrm{z}=0$ is a zero of $\mathrm{f}(\mathrm{z})$ of order m , then the canonical product is $\mathrm{z}^{\mathrm{m}} \mathrm{G}(\mathrm{z})$.

Observe that if the convergence exponent $\rho_{1} \neq$ an integer, then $\mathrm{p}=\left[\rho_{1}\right]$ and if $\rho_{1}=$ an integer, then $p=\rho_{1}$ when $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\rho_{1}}}$ is divergent and $p=\rho_{1}-1$ if $\sum_{n=1}^{\infty} \frac{1}{r_{n}^{\rho_{1}}}$ is convergent.

In any case, $\rho_{1}-1 \leq \mathrm{p} \leq \rho_{1} \leq \rho$, where $\rho=$ order of $\mathrm{f}(\mathrm{z})$.
Examples: (i) Let $\mathrm{a}_{\mathrm{n}}=\mathrm{n}$. Then $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}^{2}}=\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{n}^{2}}$ is convergent while $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}}=\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{n}}$ is divergent. So, $\mathrm{p}=1$.
(ii) Let $a_{n}=e^{n}$. Then $p=0$.

We now state an important theorem without proof. The proof can be found in any standard book.

Borel's theorem : The order of a canonical product is equal to the convergence exponent of its zeros.

Example : Find the canonical product of $\mathrm{f}(\mathrm{z})=\sin \mathrm{z}$.
Solution : $f(z)$ is an entire function with infinite number of zeros at $z=n \pi, n$ being an integer. First we consider the zeros of $f(z)$ excluding the simple zero at $z=0$. Let $a_{n}=n \pi$, $\mathrm{n}= \pm 1, \pm 2, \ldots$
$\left|a_{n}\right|=r_{n}$. Then, $r_{n}=|n \pi|$. Now, $\sum_{n=1}^{\infty} \frac{1}{r_{n}}=\sum_{n=1}^{\infty} \frac{1}{|n \pi|}$
$=\frac{1}{\pi} \sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{n}}$ is divergent, but $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{r}_{\mathrm{n}}^{2}}=\frac{1}{\pi^{2}} \sum_{\mathrm{n}=1}^{\infty} \frac{1}{\mathrm{n}^{2}}$ is convergent. Hence genus of the required canonical product $\mathrm{p}=1$.

Hence the canonical product $\mathrm{G}(\mathrm{z})$ is given by
$\mathrm{G}(\mathrm{z})=\prod_{\mathrm{n}=-\infty}^{\infty} \mathrm{E}\left(\frac{\mathrm{z}}{\mathrm{a}_{\mathrm{n}}}, 1\right)$, where $\prod_{\mathrm{n}=-\infty}^{\infty}$ means $\mathrm{n}=0$ is excluded in the product.
$=\prod_{n=-\infty}^{\infty}\left(1-\frac{z}{n \pi}\right) e^{\frac{z}{n \pi}}=\prod_{n=1}^{\infty}\left\{\left(1-\frac{z}{n \pi}\right) e^{\frac{z}{n \pi}} \cdot\left(1-\frac{z}{n \pi}\right) e^{-\frac{z}{n \pi}}\right\}$
$=\prod_{n=-\infty}^{\infty}\left(1-\frac{z^{2}}{n^{2} \pi^{2}}\right)$. Since origin is a simple zero of $\sin \mathrm{z}$, the required canonical product of $\sin \mathrm{z}$ is given by
$\sin \mathrm{z}=\mathrm{z} \prod_{\mathrm{n}=1}^{\infty}\left(1-\frac{\mathrm{z}^{2}}{\mathrm{n}^{2} \pi^{2}}\right)$.

## Exercises

1. Find the order of the entire functions:
(a) $\sinh z(b) e^{z} \sin z$,
(c) $\mathrm{e}^{\mathrm{z}^{\mathrm{n}}}$,
(d) $\mathrm{e}^{\mathrm{e}^{z}}$,
(e) $\cos \mathrm{z}$, (f) $\mathrm{e}^{\mathrm{p}(\mathrm{z}) \text {, where } \mathrm{p}(\mathrm{z})=}$ $a_{0}+a_{1} z+\cdots+a_{n} z^{n}, a_{n} \neq 0,(g) \sum_{n=0}^{\infty} \frac{z^{n}}{(n!)^{\alpha}}, \alpha>0,(h) \sum_{n=0}^{\infty}\left(\frac{e \alpha}{n}\right)^{n / \alpha} z^{n}, \alpha>0$
2. Given $f_{1}(z)$ and $f_{2}(z)$ are two entire functions of orders $\rho_{1}$ and $\rho_{2}$ respectively, show that (i) order of $f_{1}(z) f_{2}(z)$ is $\leq \max \left(\rho_{1}, \rho_{2}\right)$ (ii) order of $f_{1}(z)+f_{2}(z)$ is $\leq \max$ ( $\rho_{1}, \rho_{2}$ ), and equality occurs if $\rho_{1} \neq \rho_{2}$.
3. Find the convergence exponent of the zeros of $\sin \mathrm{z}$.
4. Find the canonical product of $\cos \mathrm{z}$.
5. Show that if $\mathrm{a}>1$, the entire function $\prod_{\mathrm{n}=1}^{\infty}\left(1-\frac{\mathrm{z}}{\mathrm{n}^{\mathrm{a}}}\right)$ is of order $\frac{1}{\mathrm{a}}$.

### 6.11 Hadamard's Factorization Theorem

Before taking up Hadamard's factorization theorem we state a theorem due to Borel and Caratheodory.

Borel and Caratheodory's theorem : Let $\mathrm{f}(\mathrm{z})$ be analytic in

$$
|\mathrm{z}| \leq \mathrm{R}, \mathrm{M}(\mathrm{r})=\max _{|\mathrm{z}|=\mathrm{r}}|\mathrm{f}(\mathrm{z})|, \mathrm{A}(\mathrm{r})=\max _{|\mathrm{z}|=\mathrm{r}}\{\operatorname{Re} \mathrm{f}(\mathrm{z})\} .
$$

Then for $0<r<R$,

$$
\begin{equation*}
\mathrm{M}(\mathrm{r}) \leq \frac{2 \mathrm{r}}{\mathrm{R}-\mathrm{r}} \mathrm{~A}(\mathrm{R})+\frac{\mathrm{R}+\mathrm{r}}{\mathrm{R}-\mathrm{r}}|\mathrm{f}(0)|<\frac{\mathrm{R}+\mathrm{r}}{\mathrm{R}-\mathrm{r}}\{\mathrm{~A}(\mathrm{R})+|\mathrm{f}(0)|\} \tag{121}
\end{equation*}
$$

Proof : Omitted (cf. Theory of entire functions-A.S.B Holland- p. 53).
Corollary : $\max _{|k|=\mathrm{r}}\left|\mathrm{f}^{(\mathrm{n})}(\mathrm{z})\right| \leq \frac{2^{\mathrm{n}+2} \cdot \mathrm{n}!\mathrm{R}}{(\mathrm{R}-\mathrm{r})^{\mathrm{n}+1}}(\mathrm{~A}(\mathrm{R})+|\mathrm{f}(0)|)$

## Hadamard's Factorization Theorem 6.16 :

If $f(z)$ is an entire function of finite order $\rho$ with infinite number of zeros and $f(0) \neq 0$, then $f(z)=e^{Q(z)} G(z)$, where $G(z)$ is the canonical product formed with the zeros of $f(z)$ and $\mathrm{Q}(\mathrm{z})$ is a polynomial of degree not greater than $\rho$.

Proof : By Weierstrass' factor theorem we already have

$$
\begin{equation*}
\mathrm{f}(\mathrm{z})=\mathrm{e}^{\mathrm{Q}(\mathrm{z})} \mathrm{G}(\mathrm{z}) \tag{123}
\end{equation*}
$$

where $\mathrm{G}(\mathrm{z})$ is the canonical product with genus p formed with the zeros $\mathrm{a}_{1}, \mathrm{a}_{2}, \ldots$ of $\mathrm{f}(\mathrm{z})$ and $\mathrm{Q}(\mathrm{z})$ is an entire function. Since $\rho$ is finite we need to show that $\mathrm{Q}(\mathrm{z})$ is a polynomial of degree $\leq \rho$. Let $m=[\rho]$. Then, $p \leq m$. Taking logarithms on both sides of (123) we have,

$$
\begin{align*}
\log f(z) & =Q(z)+\log G(z) \\
& =Q(z)+\sum_{n=1}^{\infty} \log \left(1-\frac{z}{a_{n}}\right)+\sum_{n=1}^{\infty}\left\{\frac{z}{a_{n}}+\frac{1}{2}\left(\frac{z}{a_{n}}\right)^{2}+\ldots+\frac{1}{p}\left(\frac{z}{a_{n}}\right)^{p}\right\} \tag{124}
\end{align*}
$$

Differentiating both sides of (124) $\mathrm{m}+1$ times,
$\frac{d^{m}}{d z^{m}}\left(\frac{f^{1}(z)}{f(z)}\right)=Q^{(m+1)}(z)-m!\sum_{n=1}^{\infty} \frac{1}{\left(a_{n}-z\right)^{m+1}}$
[Since $p \leq m, \frac{d^{m+1}}{d z^{m+1}} \sum_{n=1}^{\infty}\left\{\frac{z}{a_{n}}+\frac{1}{2}\left(\frac{z}{a_{n}}\right)^{2}+\ldots+\frac{1}{p}\left(\frac{z}{a_{n}}\right)^{p}\right\}=0$
and $\left.\frac{d^{m+1}}{d z^{m+1}} \log \left(1-\frac{z}{a_{n}}\right)=\frac{d^{m+1}}{d z^{m+1}} \log \left(a_{n}-z\right)=-m!\frac{1}{\left(a_{n}-z\right)^{m+1}}\right]$
Now, $\mathrm{Q}(\mathrm{z})$ will be a polynomial of degree m at most if we can show that $\mathrm{Q}^{(\mathrm{m}+1)}(\mathrm{z})=0$.

Let $g_{R}(z)=\frac{f(z)}{f(0)} \prod_{\mid a_{n} \leq R}\left(1-\frac{z}{a_{n}}\right)^{-1}$. Then $g_{R}(z)$ is an entire function and $g_{R}(z) \neq 0$ in $|z| \leq R$. [Since $f(z)$ is entire, $f(0) \neq 0$ and $\prod_{\mid a_{n} \leq R}\left(1-\frac{z}{a_{n}}\right)^{-1}$ cancels with factors in $\left.f(z)\right]$.

For $|z|=2 R$ and $\left|a_{n}\right| \leq R$ we have, $\left|1-\frac{z}{a_{n}}\right| \geq 1$. Hence,
$\left|g_{R}(\mathrm{z})\right| \leq \frac{|\mathrm{f}(\mathrm{z})|}{|\mathrm{f}(0)|}<\mathrm{Ae}^{(2 \mathrm{R})^{\rho+\varepsilon}}$ for $|\mathrm{z}|=2 \mathrm{R}$
By maximum modulus theorem, $\left|g_{R}(z)\right|<A e^{(2 R)^{\rho+\varepsilon}}$
for $|\mathrm{z}|<2 R$. Let $h_{R}(z)=\log g_{R}(z)$ such that $h_{R}(0)=0$.
Then $h_{R}(z)$ is analytic in $|z| \leq R$. Hence from (127)
$\operatorname{Re} h_{R}(z)=\log \left|g_{R}(z)\right|<K^{\rho+\varepsilon}, K=$ Constant
Hence from the corollary of the theorem of Borel and Caratheodory we have

$$
\left|h_{R}^{(m+1)}(z)\right| \leq \frac{2^{m+3}(m+1) \mid!R}{(R-r)^{m+2}} \cdot K^{p+\varepsilon} \text { for }|z|=r<R
$$

Hence for $|z|=r=\frac{R}{2}$,

$$
\begin{equation*}
\left|h_{R}^{(m+1)}(z)\right|=0\left(R^{p+\varepsilon-m-1}\right) \tag{129}
\end{equation*}
$$

$$
\text { But } h_{R}(z)=\log g_{R}(z)=\log f(z)-\log f(0)-\sum_{\left|a_{n}\right| \leq R} \log \left(1-\frac{z}{a_{n}}\right)
$$

Hence $h_{R}^{(m+1)}(z)=\frac{d^{m}}{d z^{m}}\left(\frac{f^{\prime}(z)}{f(z)}\right)+m!\sum_{\left|a_{n}\right| \leq R} \frac{1}{\left(a_{n}-z\right)^{m+1}}$
$=0\left(R^{\rho+\varepsilon-m-1}\right)+0\left(\sum_{\left|a_{n}\right\rangle R} \frac{1}{\left|a_{n}\right|^{m+1}}\right)$
for $|z|=\frac{R}{2}$ and so also for $|z|<\frac{R}{2}$ by maximum modulus theorem. The first term on the right of (130) tends to 0 as $R \rightarrow \infty$ if $\varepsilon>0$ is small enough since $m+1>\rho$. Also the second term tends to 0 since $\sum_{\mathrm{n}=1}^{\infty} \frac{1}{\left|\mathrm{a}_{\mathrm{n}}\right|^{\mathrm{m}+1}}$ is convergent.

In fact, $\sum_{\left|\mathrm{a}_{\mathrm{n}}\right\rangle \mathrm{R}} \frac{1}{\left|\mathrm{a}_{\mathrm{n}}\right|^{\mathrm{m+1}}}$ becomes the remainder term for large R .
Hence $\mathrm{Q}^{(\mathrm{m}+1)}(\mathrm{z})=0$ since $\mathrm{Q}^{(\mathrm{m}+1)}(\mathrm{z})$ is independent of R .
Thus, $\mathrm{Q}(\mathrm{z})$ is a polynomial of degree not greater than $\rho$.

### 6.12 Consequences of Hadamard's Theorem

Theorem 6.17 : An entire function of finite order admits any finite complex number except, perhaps, one number.

Proof. Let us suppose that $f$ does not admit two finite values a and $b$. Then $\mathrm{f}(\mathrm{z})-\mathrm{a} \neq 0$ for all z in $\mathbb{C}$ and hence there exists an entire function $\mathrm{g}(\mathrm{z})$ such that

$$
f(z)-a=e^{g(z)}
$$

The function $f(z)$ - a is of finite order since $f(z)$ has finite order. Following Hadamard's factorization theorem $g(z)$ must be a polynomial. Now $e^{g(z)}$ does not assume the value $b$ - a i.e. $\mathrm{g}(\mathrm{z}) \neq \log (\mathrm{b}-\mathrm{a})$ for any z in $\mathbb{C}$. As because $\mathrm{g}(\mathrm{z})$ is a polynomial it contradicts the essence of the Fundamental Theorem of Algebra [(14), Th. 3.11, page-65].

Theorem 6.18: An entire function of fractional order possesses infinitely many zeros.

Proof. Let f be an entire function of fractional order $\rho$. If possible, suppose the zeros of $f(z)$ are $\left\{a_{1}, a_{2}, \ldots a_{n}\right\}$, finite in number, counted according to their multiplicity. Then $f(z)$ can be expressed as

$$
f(z)=e^{g(z)}\left(z-a_{1}\right)\left(z-a_{2}\right) \ldots\left(z-a_{n}\right)
$$

where $g(z)$ is an entire function. Applying Hadamard's factorization theorem, the degree of the polynomial $g(z) \leq \rho$. It is easy to check that $f(z)$ and $e^{g(z)}$ are of same order. But we have already seen that the order of $e^{g(z)}$ is exactly the degree of $g(z)$, which is an integer. This implies $\rho$ is an integer. This contradiction completes the proof.

### 6.13 Meromorphic Functions

The term meromorphic comes from the Ancient Greek "meros" meaning part, as opposed to "holos" meaning whole. This function is analytic on a domain D except a set of isolated points, which are poles for the function.

Definition : A function $\mathrm{f}(\mathrm{z})$ analytic in a domain D except for poles is said to be meromorphic.

Theorem 6.19 : A rational function is meromorphic.
Proof : Let $\mathrm{f}(\mathrm{z})=\mathrm{p}(\mathrm{z}) / \mathrm{q}(\mathrm{z})$ where p and q are polynomials with no common zeros. If the degree of $p$ is less than or equal to the degree of $q$, then $f$ has only a finite number of poles and the point at infinity is not a pole. On the otherhand, if the degree of pis greater than the degree of $q$, then (taking degree of $p(z)=m$ and degree of $q(z)=n)$.

$$
\begin{aligned}
f(z) & =\frac{a_{m} z^{m}+a_{m-1} z^{m-1}+\ldots+a_{1} z+a_{0}}{b_{n} z^{n}+b_{n-1} z^{n-1}+\ldots+b_{1} z+b_{0}} \\
& =c_{m-n} z^{m-n}+c_{m-n-1} z^{m-n-1}+\ldots+c_{1} z+c_{0}+\frac{r(z)}{q(z)}
\end{aligned}
$$

where degree of $r(z) \leq n-1$. This shows that the point at infinity is a pole of order $(\mathrm{m}-\mathrm{n})$ and there lie a finite number of poles in the unextended plane. These establish that $\mathrm{f}(\mathrm{z})$ is meromorphic.

Theorem 6.20 : [Partial fraction decomposition]. Let $\mathrm{p}(\mathrm{z}), \mathrm{q}(\mathrm{z})$ be two polynomials with no common zeros and that $0 \leq \operatorname{deg}(\mathrm{p})<\operatorname{deg}(\mathrm{q})$. Let $\mathrm{a}_{1}, \ldots \mathrm{a}_{\mathrm{k}}$ be the zeros of $\mathrm{q}(\mathrm{z})$ with multiplicities $\alpha_{1}, \ldots, \alpha_{\mathrm{k}}$. Then $\mathrm{p}(\mathrm{z}) / \mathrm{q}(\mathrm{z})$ can be expressed uniquely as

$$
\begin{equation*}
\frac{p(z)}{q(z)}=\sum_{i=1}^{k} \sum_{j=1}^{\alpha_{i}} \frac{c_{i j}}{\left(z-a_{i}\right)^{j}} \tag{131}
\end{equation*}
$$

Proof. The decomposition is unique. We assume that the relation (131) exists. Let r $>0$ be small enough. Then for $\mathrm{z} \varepsilon N\left(\mathrm{a}_{\mathrm{i}}, \mathrm{r}\right),(131)$ can be rewritten as

$$
\begin{equation*}
\frac{p(z)}{q(z)}=g(z)+\sum_{j=1}^{\alpha_{i}} \frac{c_{i j}}{\left(z-a_{i}\right)^{j}} \tag{132}
\end{equation*}
$$

since $N\left(\mathrm{a}_{\mathrm{i}}, \mathrm{r}\right)$ does not contain any zero of $\mathrm{q}(\mathrm{z})$ other than $\mathrm{a}_{\mathrm{i}}, \mathrm{g}(\mathrm{z})$ is analytic at $\mathrm{z}=\mathrm{a}_{\mathrm{i}}$.

Multiplying both sides of (132) by $\left(\mathrm{z}-\mathrm{a}_{\mathrm{i}}\right)_{\mathrm{i}}$, we obtain

$$
\begin{equation*}
\frac{p(z)}{q(z)}\left(z-a_{i}\right)^{\alpha_{i}}=g(z)\left(z-a_{i}\right)^{\alpha_{i}}+\sum_{j=1}^{\alpha_{i}} c_{i j}\left(z-a_{i}\right)^{\alpha_{i}-j} \tag{133}
\end{equation*}
$$

Now the function $\frac{p(z)}{q(z)}\left(z-a_{i}\right)^{\alpha_{i}}$ is analytic for all $z$ belonging to $N\left(a_{i}, r\right)$ and hence can be expanded in a Taylor series in a neighbourhood of $\mathrm{a}_{\mathrm{i}}$ in $N\left(\mathrm{a}_{\mathrm{i}}, \mathrm{r}\right)$

$$
\begin{equation*}
\frac{\mathrm{p}(\mathrm{z})}{\mathrm{q}(\mathrm{z})}\left(\mathrm{z}-\mathrm{a}_{\mathrm{i}}\right)^{\alpha_{\mathrm{i}}}=\sum_{\mathrm{n}=0}^{\infty} \mathrm{c}_{\mathrm{n}}\left(\mathrm{z}-\mathrm{a}_{\mathrm{i}}\right)^{\mathrm{n}} \tag{134}
\end{equation*}
$$

Combining (133) and (134), we write

$$
\begin{aligned}
\sum_{n=0}^{\alpha} c_{n}\left(z-a_{i}\right)^{n}=g(z)\left(z-a_{i}\right)^{\alpha_{i}}+c_{i \alpha_{i}} & +c_{i \alpha_{i-1}}\left(z-a_{i}\right)+\ldots+ \\
& +c_{i 1}\left(z-a_{i}\right)^{\alpha_{i}-1}
\end{aligned}
$$

Comparing the coefficients we find
$c_{i} \alpha_{i}=c_{0}, c_{\alpha_{i}-1}=c_{1}, \ldots, c_{i 1}=c_{\alpha_{i}-1}$ uniquely

## Existence of the decomposition.

The principal part associated to each pole $\mathrm{a}_{\mathrm{i}}$ is

$$
\sum_{\mathrm{j}=1}^{\alpha_{\mathrm{i}}} \frac{\mathrm{c}_{\mathrm{ij}}}{\left(\mathrm{z}-\mathrm{a}_{\mathrm{i}}\right)^{\mathrm{j}}}
$$

Now if we subtract all the principal parts we find the function

$$
f(z)=\frac{p(z)}{q(z)}-\sum_{i=1}^{k} \sum_{j=1}^{\alpha_{i}} \frac{c_{i j}}{\left(z-a_{i}\right)^{j}}
$$

is analytic in the extended plane. Now each of the terms

$$
\frac{c_{i j}}{\left(z-a_{i}\right)^{j}}
$$

converges to zero for $\mathrm{z} \rightarrow \infty$, and also $\mathrm{p}(\mathrm{z}) / \mathrm{q}(\mathrm{z})$ converges to zero for $\mathrm{z} \rightarrow \infty$ since $\operatorname{deg}(q)>\operatorname{deg}(p)$. This shows that $f(z) \rightarrow 0$ for $z \rightarrow \infty$. But then $f$ is necessarily
bounded and hence constant by Liouville's theorem. A constant function tending to zero as $\mathrm{z} \rightarrow \infty$ must be identically zero.

Example 4 : Consider the rational function

$$
\frac{p(z)}{q(z)}=\frac{2 z^{3}+(5 i+3) z^{2}+(3-5 i)}{z^{4}-1}
$$

We can write this as

$$
\begin{align*}
\frac{\mathrm{p}(\mathrm{z})}{\mathrm{q}(\mathrm{z})} & =\frac{\alpha}{\mathrm{z}-1}+\frac{\beta}{\mathrm{z}+1}+\frac{\gamma}{\mathrm{z}-\mathrm{i}}+\frac{\delta}{\mathrm{z}+\mathrm{i}}  \tag{135}\\
& =\mathrm{g}_{1}(\mathrm{z})+\frac{\alpha}{\mathrm{z}-1}
\end{align*}
$$

considering z belonging to $|\mathrm{z}-1|<1$. Then

$$
\frac{\mathrm{p}(\mathrm{z})}{\mathrm{q}(\mathrm{z})}(\mathrm{z}-1)=\mathrm{g}_{1}(\mathrm{z})(\mathrm{z}-1)+\alpha \Rightarrow \alpha=2
$$

### 6.14 Partial Fraction Expansion of Meromorphic Functions

Let $f(z)$ be a meromorphic function and $z_{0}$ be a pole of order $m$ with the principal part

$$
\mathrm{p}(\mathrm{z})=\frac{\mathrm{c}_{-\mathrm{m}}}{\left(\mathrm{z}-\mathrm{z}_{0}\right)}+\frac{\mathrm{c}_{-\mathrm{m}+1}}{\left(\mathrm{z}-\mathrm{z}_{0}\right)^{\mathrm{m}+1}}+\ldots+\frac{\mathrm{c}_{-1}}{\mathrm{z}-\mathrm{z}_{0}}
$$

Then $f(z)$ can be written as [see § 6.2, (14)]

$$
\mathrm{f}(\mathrm{z})=\mathrm{p}(\mathrm{z})+\mathrm{g}(\mathrm{z})
$$

where $g(z)$ is an entire function. Now if, in general, $z_{1}, z_{2}, \ldots, z_{n}$ are the poles of a meromorphic function $f$ with the corresponding principl parts $P_{1}, P_{2} \ldots, P_{n}$ then $f$ can be expressed as

$$
\begin{equation*}
\mathrm{f}(\mathrm{z})=\sum_{\mathrm{j}=1}^{\mathrm{n}} \mathrm{P}_{\mathrm{j}}(\mathrm{z})+\psi(\mathrm{z}) \tag{136}
\end{equation*}
$$

where $\psi(\mathrm{z})$ is an entire function.
But the question arises whether it is possible to construct a meromorphic function possessing poles at the sequence of points $\left\{\mathrm{z}_{\mathrm{n}}\right\}$ with corresponding principal parts $\mathrm{P}_{1}, \mathrm{P}_{2} \ldots$ Because in this case the series $\Sigma P_{j}(z)$ in (136) turns out to be an infinite series $\sum_{j=1}^{n} P_{j}(z)$, which needs to be convergent.

Gösta Mittag Leffler (1846-1927), German in origin but his several generations lived in Sweden, overcame this difficulty by introducing a polynomial $\mathrm{p}_{\mathrm{n}}(\mathrm{z})$ dependent on $\mathrm{Z}_{\mathrm{n}}$ and $\mathrm{P}_{\mathrm{n}}(\mathrm{z})$ so that the series $\sum_{\mathrm{n}=1}^{\infty}\left\{\mathrm{P}_{\mathrm{n}}(\mathrm{z})-\mathrm{p}_{\mathrm{n}}(\mathrm{z})\right\}$ is uniformly convergent in any compact set K not containing any points of the sequence $\left\{\mathrm{z}_{\mathrm{n}}\right\}$.

Theorem 6.21 [The Mittag Leffler Theorem] : Given a sequence of distinct complex numbers $\left\{\mathrm{z}_{\mathrm{n}}\right\}$,

$$
\left|z_{1}\right| \leq\left|z_{2}\right| \leq \ldots, \lim _{n \rightarrow \infty} z_{n}=\infty
$$

and a sequence of rational functions $\left\{\mathrm{P}_{\mathrm{n}}(\mathrm{z})\right\}$,

$$
\begin{equation*}
\mathrm{P}_{\mathrm{n}}(\mathrm{z})=\sum_{\mathrm{k}=1}^{\ln } \frac{\mathrm{c}_{\mathrm{nk}}}{\left(\mathrm{z}-\mathrm{z}_{\mathrm{n}}\right)^{\mathrm{k}}}, 1_{\mathrm{n}} \geq 1, \mathrm{n}=1,2, \ldots \tag{137}
\end{equation*}
$$

there exists a meromorphic function $f(z)$ having poles at the points $\mathrm{Z}_{\mathrm{n}}$ and only there with $P_{n}(z)$ as its principal part at $\mathrm{z}_{\mathrm{n}}$ and can be represented in the form of an expansion

$$
\mathrm{f}(\mathrm{z})=\sum_{\mathrm{n}=1}^{\infty}\left[\mathrm{P}_{\mathrm{n}}(\mathrm{z})-\mathrm{p}_{\mathrm{n}}(\mathrm{z})\right]+\mathrm{h}(\mathrm{z})
$$

where $\mathrm{h}(\mathrm{z})$ is an arbitrary entire function and $\mathrm{p}_{\mathrm{n}}(\mathrm{z})$ is suitable partial sum of Taylor's expansion of the singular part which is analytic in the open disc $|z|<\left|z_{n}\right|$.

Proof. Without loss of generality we assume that $\mathrm{z}=0$ is not a pole of $\mathrm{f}(\mathrm{z})$. Now $\mathrm{P}_{\mathrm{k}}(\mathrm{z})$ is analytic for $|\mathrm{z}|<\left|\mathrm{z}_{\mathrm{k}}\right|$ and can be expanded in this neighbourhood of z :

$$
P_{k}(z)=\sum_{j=0}^{\infty} c_{j}^{(k)} z^{j}
$$

and hence this series converges uniformly in the disk $|z| \leq\left|z_{k}\right| / 2$. Let $p_{k}(z)=\sum_{j=0}^{\alpha_{k}} c_{j}^{(k)} z^{j}$ be a partial sum of this expansion such that

$$
\left|\mathrm{P}_{\mathrm{k}}(\mathrm{z})-\mathrm{p}_{\mathrm{k}}(\mathrm{z})\right|<\frac{1}{\mathrm{k}^{2}} \text { for }|\mathrm{z}| \leq\left|\mathrm{z}_{\mathrm{k}}\right| / 2 \text {. }
$$

Let R be an arbitrary large positive number and since $\mathrm{z}_{\mathrm{n}} \rightarrow \infty$ as $\mathrm{n} \rightarrow \infty$ we can find an $N(R)$ so large that $\left|z_{n}\right|>2 R$ when $n \geq N(R)$. Therefore in the circle $|z|<R<\frac{\left|z_{N}\right|}{2}$

$$
\sum_{n=1}^{\infty}\left[P_{n}(z)-p_{n}(z)\right]=\sum_{n=1}^{N(R)-1}\left[P_{n}(z)-p_{n}(z)\right]+\sum_{n=N(R)}^{\infty}\left[P_{n}(z)-p_{n}(z)\right]
$$

the first sum in the r.h.s is finite and the second sum $\sum_{\mathrm{N}(\mathrm{R})}^{\infty}$ is absolutely and uniformly convergent by comparison with the convergent series $\sum_{n=N(R)}^{\infty} 2^{-n}$. Therefore $\sum_{n=1}^{\infty}\left[P_{n}(z)-p_{n}(z)\right]$ is analytic in $|z|<R$ except at the poles belonging to the sequence $\left\{z_{n}\right\}$. It is thus a meromorphic function with the poles at $\mathrm{z}_{1}, \mathrm{z}_{2}, \ldots$ and with the principal parts $P_{1}(z), P_{2}(z), \ldots$ at each point $z_{n}$ respectively. Now if $f(z)$ possesses the same poles only with the same principal parts then

$$
\mathrm{f}(\mathrm{z})-\sum_{\mathrm{n}=1}^{\infty}\left[\mathrm{P}_{\mathrm{n}}(\mathrm{z})-\mathrm{p}_{\mathrm{n}}(\mathrm{z})\right]
$$

is an entire function $\mathrm{h}(\mathrm{z})$, say. This completes the proof.
Example 5 : Prove that

$$
\pi \cot \pi z=\frac{1}{z}+\sum_{n=-\infty}^{\infty},\left\{\frac{1}{z-n}+\frac{1}{n}\right\}
$$

Solution : The given function $\pi$ cot $\pi z$ has simple poles at $\mathrm{z}=0, \pm 1, \pm 2, \ldots$ with residue 1.

Here,

$$
\begin{equation*}
\frac{1}{\mathrm{z}-\mathrm{n}}=-\frac{1}{\mathrm{n}} \frac{1}{\left(1-\frac{\mathrm{z}}{\mathrm{n}}\right)}=-\frac{1}{\mathrm{n}}\left(1+\frac{\mathrm{z}}{\mathrm{n}}+\frac{\mathrm{z}^{2}}{\mathrm{n}^{2}}+\ldots\right),|\mathrm{z}|<\mathrm{n} \tag{138}
\end{equation*}
$$

Let $|z|<R$ and $N(R)$ be so large that $R<\frac{n}{2}$ when $n \geq N(R)$. Then from (138), we find

$$
\left|\frac{1}{\mathrm{z}-\mathrm{n}}+\frac{1}{\mathrm{n}}\right| \leq \frac{2 \mathrm{R}}{\mathrm{~N}^{2}}, \mathrm{n} \geq \mathrm{N}
$$

Now, since $\Sigma 1 / \mathrm{N}^{2}$ is convergent, the series

$$
\sum_{n=-\infty}^{\infty} \cdot\left\{\frac{1}{z-n}+\frac{1}{n}\right\}
$$

converges uniformly on any compact set (lying in $|z|<R$ ) not containing any of the points $\mathrm{z}= \pm 1, \pm 2, \ldots$ Therefore applying the Mittag-Leffler theorem we can express

$$
\begin{equation*}
\pi \cot \pi z=\frac{1}{z}+\sum_{n=-\infty}^{\infty},\left\{\frac{1}{z-n}+\frac{1}{n}\right\}+h(z) \tag{139}
\end{equation*}
$$

where $\mathrm{h}(\mathrm{z})$ is an entire function. Differentiating term-wise, we obtain

$$
\begin{aligned}
\pi^{2} \operatorname{cosec}^{2} \pi z & =\frac{1}{z^{2}}+\sum_{n=-\infty}^{\infty} \frac{1}{(\mathrm{z}-\mathrm{n})^{2}}-\mathrm{h}^{\prime}(\mathrm{z}) \\
& =\sum_{\mathrm{n}=-\infty}^{\infty} \frac{1}{(\mathrm{z}-\mathrm{n})^{2}}-\mathrm{h}^{\prime}(\mathrm{z})
\end{aligned}
$$

and

$$
\begin{equation*}
h^{\prime}(\mathrm{z})=\sum_{\mathrm{n}=-\infty}^{\infty} \frac{1}{(\mathrm{z}-\mathrm{n})^{2}}-\pi^{2} \operatorname{cosec}^{2} \pi \mathrm{z}=\mathrm{f}(\mathrm{z})-\psi(\mathrm{z}) \text {, say } \tag{140}
\end{equation*}
$$

We notice that the functions $\mathrm{f}(\mathrm{z})$ and $\psi(\mathrm{z})$ are both periodic with period 1 and consequently $\mathrm{h}^{\prime}(\mathrm{z})$ is also periodic with the same period.

Let $\mathrm{z}=\mathrm{x}+\mathrm{iy}$. Consider the strip $0 \leq \mathrm{x} \leq 1$. In fact, the convergence of the series in (140) is uniform for $\mathrm{y} \geq 1$, say and the limit tends to 0 as $\mathrm{y} \rightarrow \infty$ (this can be seen on taking the limit in each term of the series).

$$
\text { Again, } \begin{aligned}
\sin (x+i y) & =\sin x \cos (i y)+\cos x \sin (i y) \\
& =\sin x \cosh y+i \cos x \sinh y
\end{aligned}
$$

and so

$$
\begin{aligned}
|\sin \pi \mathrm{z}|^{2} & =|\sin \pi(\mathrm{x}+\mathrm{iy})|^{2} \\
& =\sin ^{2} \pi \mathrm{x} \cosh ^{2} \pi \mathrm{y}+\cos ^{2} \pi \mathrm{x} \sinh ^{2} \pi \mathrm{y} \\
& =\cosh ^{2} \pi \mathrm{y}-\cos ^{2} \pi \mathrm{x}
\end{aligned}
$$

which establishes that $\pi^{2} \operatorname{cosec}^{2} \pi z$ tends uniformly to zero as $y \rightarrow \infty$. From these we conclude that $\mathrm{h}^{\prime}(\mathrm{z})$ is bounded in the period strip $0 \leq \mathrm{x} \leq 1$ and due to its periodicity it is bounded in the entire plane. By Liouville's theorem it then reduces to a constant. Now since

$$
\lim _{y \rightarrow \infty} h^{\prime}(z)=\lim _{y \rightarrow \infty} f(z)-\lim _{y \rightarrow \infty} \psi(z)=0-0=0
$$

$\mathrm{h}^{\prime}(\mathrm{z})$ is indeed zero and $\mathrm{h}(\mathrm{z})=\mathrm{c}$, a constant. Then from (139),

$$
\pi \cot \pi z=\frac{1}{z}+\sum_{n=-\infty}^{\infty} \cdot\left(\frac{1}{z-n}+\frac{1}{n}\right)+c
$$

For, $\mathrm{z}=\frac{1}{2}$

$$
0=2+\sum_{1}^{\infty}\left(\frac{2}{1-2 \mathrm{k}}+\frac{2}{1+2 \mathrm{k}}\right)+\mathrm{c}
$$

$$
\begin{aligned}
& =2+2\left\{\left(\frac{1}{-1}+\frac{1}{3}\right)+\left(-\frac{1}{3}+\frac{1}{5}\right)+\left(-\frac{1}{5}+\frac{1}{7}\right)+\ldots\right\}+\mathrm{c} \\
& =2-2+\mathrm{c}
\end{aligned}
$$

$\Rightarrow \mathrm{c}=0$ i.e. $\mathrm{h}(\mathrm{z}) \equiv 0$. Finally we obtain

$$
\pi \cot \pi \mathrm{z}=\frac{1}{\mathrm{z}}+\sum_{\mathrm{z}=-\infty}^{\infty},\left\{\frac{1}{\mathrm{z}-\mathrm{n}}+\frac{1}{\mathrm{n}}\right\}
$$

Now since the series on the r.h.s is uniformly convergent on any compact set not containing the points $\mathrm{Z}=0, \pm 1, \pm 2 \ldots$, rearrangement of the terms are permissible and hence

$$
\begin{equation*}
\pi \cot \pi \mathrm{z}=\frac{1}{\mathrm{z}}+\sum_{\mathrm{n}=1}^{\infty} \frac{2 \mathrm{z}}{\mathrm{z}^{2}-\mathrm{n}^{2}} \tag{141}
\end{equation*}
$$

Remark : Here it is proved incidentally that

$$
\begin{equation*}
\pi^{2} \operatorname{cosec}^{2} \pi z=\sum_{\mathrm{n}=-\infty}^{\infty} \frac{1}{(\mathrm{z}-\mathrm{n})^{2}} \tag{142}
\end{equation*}
$$

[see equation (140)]
We can now utilize the identity (141) to calculate easily some familiar sums. Here the 1.h.s of (141) has the Laurent series expansion in the neighbourhood of $z=0$.

$$
\pi \cot \pi z=\frac{1}{\mathrm{z}}-\frac{\pi^{2} \mathrm{z}}{3}-\frac{\pi^{4} \mathrm{z}^{3}}{45}-\frac{2 \pi^{6} \mathrm{z}^{5}}{945}-\ldots
$$

Note that the series on the r.h.s of (141) converges uniformly near $\mathrm{z}=0$. By Th. 4.14 [14] it converges uniformly together with all derivatives. Again

$$
\frac{2 z}{z^{2}-n^{2}}=-2\left(\frac{z}{n^{2}}+\frac{z^{3}}{n^{4}}+\frac{z^{5}}{n^{6}}+\ldots\right)
$$

and we obtain easily,

$$
\begin{equation*}
\sum_{n \geq 1} \frac{1}{n^{2}}=\frac{\pi^{2}}{6}, \sum_{n=1}^{\infty} \frac{1}{n^{4}}=\frac{\pi^{4}}{90}, \sum_{n=1}^{\infty} \frac{1}{n^{6}}=\frac{\pi^{6}}{945} \tag{143}
\end{equation*}
$$

Example 6. Prove that

$$
\pi \tan \pi z=-\sum_{n=-\infty}^{\infty}\left[\frac{1}{z-\left(n+\frac{1}{2}\right)}+\frac{1}{n+\frac{1}{2}}\right]
$$

[or, equivalently, $\left.\pi \tan \pi z=2 z \sum_{n=0}^{\infty}\left[\left(n+\frac{1}{2}\right)^{2}-z^{2}\right]^{-1}\right]$
Solution : Here the given function $\pi$ tan $\pi z$ possesses simple poles at $\mathrm{z}= \pm \frac{1}{2}, \pm \frac{3}{2}, \ldots$ with residue -1 .

Then, $\frac{-1}{z-\left(n+\frac{1}{2}\right)}=\frac{1}{\left(n+\frac{1}{2}\right)\left(1-\frac{\mathrm{z}}{\mathrm{n}+\frac{1}{2}}\right)}=\frac{1}{\mathrm{n}+\frac{1}{2}}\left[1+\frac{\mathrm{z}}{\mathrm{n}+\frac{1}{2}}+\left(\frac{\mathrm{z}}{\mathrm{n}+\frac{1}{2}}\right)^{2}+\cdots\right]$
and the series

$$
\sum_{n=-\infty}^{\infty}\left[\frac{-1}{z-\left(n+\frac{1}{2}\right)}-\frac{1}{n+\frac{1}{2}}\right]
$$

converges uniformly on any compact set not containing any of the poles of the given function. By Mittag-Leffler theorem,

$$
\pi \tan \pi z=-\sum_{n=-\infty}^{\infty}\left[\frac{1}{z-\left(n+\frac{1}{2}\right)}+\frac{1}{n+\frac{1}{2}}\right]+h(z)
$$

where $\mathrm{h}(\mathrm{z})$ is an arbitray entire function. Now proceeding as in example 5, we can have the desired result.

Example 7 : Establish that

$$
\frac{1}{e^{z}-1}=-\frac{1}{2}+\frac{1}{z}+\sum_{n=1}^{\infty} \frac{2 z}{z^{2}+4 n^{2} \pi^{2}}
$$

Solution : We rewrite $1 / \mathrm{e}^{\mathrm{z}}-1$ as

$$
\frac{1}{\mathrm{e}^{\mathrm{z}}-1}=\frac{\mathrm{e}^{-\mathrm{z} / 2}}{\mathrm{e}^{\mathrm{z} / 2}-\mathrm{e}^{-\mathrm{z} / 2}}=\frac{1}{2} \frac{\mathrm{e}^{-\mathrm{z} / 2}-\mathrm{e}^{\mathrm{z} / 2}+\mathrm{e}^{\mathrm{z} / 2}+\mathrm{e}^{-\mathrm{z} / 2}}{\mathrm{e}^{\mathrm{z} / 2}-\mathrm{e}^{-\mathrm{z} / 2}}=-\frac{1}{2}+\frac{1}{2} \operatorname{coth} \frac{\mathrm{z}}{2}
$$

But $\operatorname{coth} \frac{Z}{2}=\frac{\cosh \frac{z}{2}}{\sinh \frac{z}{2}}=\frac{i \cos \left(i \frac{z}{2}\right)}{\sin \left(i \frac{z}{2}\right)}=i \cot \left(i \frac{z}{2}\right)$
Now utilising (141) we get the result.

### 6.15 Partial Fraction Expansion of Meromorphic Functions Using Residue theorem

Let us suppose f to be a meromorphic function whose only singularities are simple poles $\mathrm{z}_{1}, \mathrm{z}_{2}, \ldots$ with increasing moduli $0<\left|\mathrm{z}_{1}\right| \leq\left|\mathrm{z}_{2}\right| \leq \ldots$,
$\lim _{n \rightarrow \infty} Z_{n}=\infty$ and $\operatorname{Res}\left(f(z) ; Z_{n}\right)=A_{n}$. Suppose there exists a sequence $\left\{C_{n}\right\}$ of simple closed contours such that
(i) $\mathrm{C}_{\mathrm{n}}$ does not contain any of the poles $\mathrm{z}_{\mathrm{k}}$
(ii) each $\mathrm{C}_{\mathrm{n}}$ lies inside $\mathrm{C}_{\mathrm{n}+1}$
(iii) $\min _{z \in \mathrm{C}_{\mathrm{n}}}|\mathrm{z}|=\mathrm{R}_{\mathrm{n}} \rightarrow+\infty$ as $\mathrm{n} \rightarrow+\infty$
(iv) length of $\mathrm{C}_{\mathrm{n}}$ is $0\left(\mathrm{R}_{\mathrm{n}}\right)$
(v) $\max _{z \in C_{n}}|f(z)|=0\left(R_{n}\right)$

Then $\quad f(z)=f(0)+\sum_{k=1}^{\infty} A_{k}\left(\frac{1}{z-z_{k}}+\frac{1}{z_{k}}\right)$
The series (144) converges uniformly in any bounded domain not containing the poles of $f(z)$.

To prove the above result we consider the integral

$$
\begin{equation*}
\mathrm{I}_{\mathrm{n}}(\mathrm{z})=\frac{1}{2 \pi \mathrm{i}} \int_{\mathrm{C}_{\mathrm{n}}} \frac{\mathrm{zf}(\varsigma)}{\zeta(\varsigma-\mathrm{z})} \mathrm{d} \zeta \tag{145}
\end{equation*}
$$

where $\mathrm{z} \in \operatorname{Int} \mathrm{C}_{\mathrm{n}}$ and $\mathrm{z} \neq \mathrm{z}_{\mathrm{k}}(\mathrm{k}=1,2, \ldots)$
Here the integrand in (145) possesses simple poles at $\varsigma=0, \varsigma=z$ and $\varsigma=\mathrm{z}_{\mathrm{k}} \in \operatorname{Int} \mathrm{C}_{\mathrm{n}}$. Then using the Residue theorem, we find from (145) that

$$
\mathrm{I}_{\mathrm{n}}(\mathrm{z})=\left[\frac{\mathrm{zf}(\varsigma)}{\varsigma-\mathrm{z}}\right]_{\varsigma=0}+\left[\frac{\mathrm{zf}(\varsigma)}{\varsigma}\right]_{\varsigma=\mathrm{z}}+\left[\frac{1}{\varsigma(\varsigma-\mathrm{z})}\right]_{\varsigma=z_{\mathrm{k}}} \operatorname{Res}\left(\mathrm{f}(\varsigma) ; \mathrm{z}_{\mathrm{k}}\right)
$$

$$
=-\mathrm{f}(0)+\mathrm{f}(\mathrm{z})+\sum_{\mathrm{z}_{\mathrm{k}} \in \operatorname{lntC}_{\mathrm{n}}} \frac{\mathrm{zA}_{\mathrm{k}}}{\mathrm{z}_{\mathrm{k}}\left(\mathrm{z}_{\mathrm{k}}-\mathrm{z}\right)}
$$

Thus,

$$
\begin{equation*}
\mathrm{f}(\mathrm{z})=\mathrm{f}(0)+\sum_{\mathrm{z}_{\mathrm{k}} \in \operatorname{lntC}_{\mathrm{n}}} A_{\mathrm{k}}\left(\frac{1}{\mathrm{z}-\mathrm{z}_{\mathrm{k}}}+\frac{1}{\mathrm{z}_{\mathrm{k}}}\right)+\frac{1}{2 \pi \mathrm{i}} \int_{\mathrm{C}_{\mathrm{n}}} \frac{\mathrm{zf}(\varsigma)}{\zeta(\varsigma-\mathrm{z})} \mathrm{d} \varsigma \tag{146}
\end{equation*}
$$

We now show that $\lim _{n \rightarrow \infty}\left|I_{n}(z)\right|=0$ for $|z|<R$.

$$
\left|I_{\mathrm{n}}(\mathrm{z})\right| \leq \frac{|\mathrm{z}|}{2 \pi} \int_{\mathrm{C}_{\mathrm{n}}} \frac{|\mathrm{f}(\varsigma)|}{|\zeta||\varsigma-\mathrm{z}|}|\mathrm{d} \varsigma|<\frac{\mathrm{R}}{2 \pi} \int_{\mathrm{C}_{\mathrm{n}}} \frac{|\mathrm{f}(\varsigma)|}{|\varsigma||\varsigma-\mathrm{R}|}|\mathrm{d} \zeta| \rightarrow 0
$$

as $\mathrm{n} \rightarrow \infty$ by the given conditions (iii), (iv) and (v).
Then (144) follows from (146) considering all the contours $\mathrm{C}_{1}, \mathrm{C}_{2}, \ldots$ etc.
Example 8 : If $\alpha_{n}$ are positive roots of the equation $\tan z=z$, show that

$$
\frac{z \sin z}{\sin z-z \cos z}=\frac{3}{z}+\sum_{n=1}^{\infty} \frac{2 z}{z^{2}-\alpha_{n}^{2}}
$$

where $\left(\mathrm{n}-\frac{1}{2}\right) \pi<\alpha_{\mathrm{n}}<\left(\mathrm{n}+\frac{1}{2}\right) \pi$.
Solution : Given $\alpha_{n}$ are positive roots of $\tan z=z$, so $\pm \alpha_{n}$ are roots of $\sin z-z$ $\cos z=0$. To check whether the function $f(z) / g(z)$, where $f(z)=z \sin z$ and $g(z)=\sin$ $z-z \cos z$, has any pole at $z=0$ we notice that
$f^{\prime}(z)=\sin z+z \cos z$
$f^{\prime \prime}(z)=2 \cos z-z \sin z$
$\mathrm{f}^{\prime}(0)=0$ but $\mathrm{f}^{\prime \prime}(0) \neq 0$

$$
\begin{aligned}
& \mathrm{g}^{\prime}(\mathrm{z})=\mathrm{z} \sin \mathrm{z}=\mathrm{f}(\mathrm{z}) \\
& \mathrm{g}^{\prime \prime}(\mathrm{z})=\mathrm{f}^{\prime}(\mathrm{z}) \\
& \mathrm{f}^{\prime \prime}(\mathrm{z})=\mathrm{g}^{\prime \prime \prime}(\mathrm{z}) \\
& \text { so, } \mathrm{g}^{\prime}(0)=\mathrm{g}^{\prime \prime}(0)=0 \text { but } \mathrm{g}^{\prime \prime \prime}(0) \neq 0
\end{aligned}
$$

Thus origin is the double zero of $f(z)$ and triple zero of $g(z)$. As a result the given function $\mathrm{f} / \mathrm{g}$ possesses a simple pole at $\mathrm{z}=0$. To find its residue at $\mathrm{z}=0$ we note that

$$
\frac{\mathrm{f}^{\prime \prime}(\mathrm{z})}{\left(\mathrm{z}^{2}\right)^{\prime \prime}}=1 \text { and } \frac{\mathrm{g}^{\prime \prime \prime}(\mathrm{z})}{\left(\mathrm{z}^{3}\right)^{\prime \prime \prime}}=\frac{1}{3}
$$

and so residue there is 3 . Thus the function $F(z)=\frac{z \sin z}{\sin z-z \cos z}-\frac{3}{z}$ has the
simple poles at $\mathrm{z}= \pm \alpha_{\mathrm{n}}$ as its only singularities and $\operatorname{Res}\left(\mathrm{F}(\mathrm{z}) ; \pm \alpha_{\mathrm{n}}\right)=1$ and $\mathrm{F}(0)=0$ since $F(z)=-F(-z)$.

Since $\left(\mathrm{n}-\frac{1}{2}\right) \pi<\alpha_{\mathrm{n}}<\left(\mathrm{n}+\frac{1}{2}\right) \pi$, we consider the sequence of contours $\left\{\mathrm{C}_{\mathrm{n}}\right\}$, formed by the straight lines $x= \pm b_{n}, y= \pm b_{n}$ with $b_{n}=\left(n+\frac{1}{2}\right) \pi, n=1,2 \ldots$,


$$
\mathrm{A}_{\mathrm{n}} \mathrm{~B}_{\mathrm{n}} \mathrm{P}_{\mathrm{n}} \mathrm{Q}_{\mathrm{n}} \text { shown below : }
$$

We find that when $\mathrm{z} \in \mathrm{B}_{\mathrm{n}} \mathrm{P}_{\mathrm{n}}, \mathrm{z}=\mathrm{b}_{\mathrm{n}}+\mathrm{iy}$, where $-b_{n} \leq y \leq b_{n}$.

Hence,

$$
\begin{align*}
|\cot \mathrm{z}| & =\left|\frac{\cos \left\{\left(\mathrm{n}+\frac{1}{2}\right) \pi+\mathrm{iy}\right\}}{\sin \left\{\left(\mathrm{n}+\frac{1}{2}\right) \pi+\mathrm{iy}\right\}}\right| \\
& =\left|\frac{\sin (\mathrm{iy})}{\cos (\mathrm{iy})}\right|=\left|\frac{\mathrm{e}^{\mathrm{y}}-\mathrm{e}^{-\mathrm{y}}}{\mathrm{e}^{\mathrm{y}}+\mathrm{e}^{-\mathrm{y}}}\right| \tag{147}
\end{align*}
$$

Same result holds when $z \in A_{n} Q_{n}$. Now when $z$ lies on either of the lines $A_{n} B_{n}$ or $\mathrm{Q}_{\mathrm{n}} \mathrm{P}_{\mathrm{n}}, \mathrm{z}=\mathrm{x} \pm \mathrm{i}\left(\mathrm{n}+\frac{1}{2}\right) \pi$

$$
\begin{align*}
|\cot \mathrm{z}| & =\left|\frac{\cos \left\{\mathrm{x} \pm \mathrm{i}\left(\mathrm{n}+\frac{1}{2}\right) \pi\right\}}{\sin \left\{\mathrm{x} \pm \mathrm{i}\left(\mathrm{n}+\frac{1}{2}\right) \pi\right\}}\right| \geq \frac{\sinh \left(\mathrm{n}+\frac{1}{2}\right) \pi}{\cosh \left(\mathrm{n}+\frac{1}{2}\right) \pi} \\
& =\frac{1-\mathrm{e}^{-2(\mathrm{n}+1)^{\pi}}}{1+\mathrm{e}^{-(2 \mathrm{n}+1)^{\pi}}} \geq \frac{\mathrm{e}^{\pi}-1}{\mathrm{e}^{\pi}+1} \tag{148}
\end{align*}
$$

The given function can be rewritten as

$$
\frac{\mathrm{z} \sin \mathrm{z}}{\sin \mathrm{z}-\mathrm{z} \cos \mathrm{z}}=\frac{1}{\frac{1}{\mathrm{z}}-\cot \mathrm{z}}
$$

I. Bound on the sides $A_{n} Q_{n} \& B_{n} P_{n}$ of the square $C_{n}$ : Using (147), we obtain

$$
\left|\frac{1}{\frac{1}{\mathrm{z}}-\cot \mathrm{z}}\right| \leq \frac{1}{|\cot \mathrm{z}|-\frac{1}{|\mathrm{z}|}}=\frac{1}{\left|\frac{\mathrm{e}^{\mathrm{y}}-\mathrm{e}^{-y}}{\mathrm{e}^{\mathrm{y}}+\mathrm{e}^{-\mathrm{y}} \mid}\right|-\frac{1}{\sqrt{\mathrm{~b}_{\mathrm{n}}^{2}+\mathrm{y}^{2}}}} \rightarrow 1 \text { as } \mathrm{n} \rightarrow \infty .
$$

II. Bound on the sides $A_{n} B_{n} \& Q_{n} P_{n}$ of $C_{n}$ : Here we apply (148) to achieve

$$
\left|\frac{1}{\frac{1}{\mathrm{z}}-\cot \mathrm{z}}\right| \leq \frac{1}{|\cot \mathrm{z}|-\frac{1}{|\mathrm{z}|}} \leq \frac{1}{\frac{\mathrm{e}^{\pi}-1}{\mathrm{e}^{\pi}+1}-\frac{1}{\sqrt{\mathrm{~b}_{\mathrm{n}}^{2}+\mathrm{y}^{2}}}} \rightarrow \frac{\mathrm{e}^{\pi}+1}{\mathrm{e}^{\pi}-1} \text { as } \mathrm{n} \rightarrow \infty .
$$

Thus,

$$
\left|\frac{\mathrm{z} \sin \mathrm{z}}{\sin \mathrm{z}-\mathrm{z} \cos \mathrm{z}}\right| \leq \frac{\mathrm{e}^{\pi}+1}{\mathrm{e}^{\pi}-1}, \mathrm{z} \in \mathrm{C}_{\mathrm{n}}, \mathrm{n}=1,2, \ldots
$$

This shows that the function $\mathrm{F}(\mathrm{z})$ is bounded on the sequence of contours $\left\{\mathrm{C}_{\mathrm{n}}\right\}$ and we can apply (144) to prove

$$
\begin{aligned}
\frac{z \sin z}{\sin z-z \cos z} & =\frac{3}{2}+\sum_{n=1}^{\infty}\left[\frac{1}{z-\alpha_{n}}+\frac{1}{\alpha_{n}}+\frac{1}{z+\alpha_{n}}-\frac{1}{\alpha_{n}}\right] \\
& =\frac{3}{2}+\sum_{n=1}^{\infty} \frac{2 z}{z^{2}-\alpha_{n}^{2}}
\end{aligned}
$$

## Exercises

1. Obtain partial fraction expansion of $\operatorname{cosec} \mathrm{z}$.
2. Prove that

$$
\sec z=\sum_{n=1}^{\infty}(-1)^{n} \frac{(2 n-1) \pi}{z^{2}-\left(n-\frac{1}{2}\right)^{2} \pi^{2}}
$$

3. Show that

$$
\tan \mathrm{z}=-\sum_{\mathrm{n}=1}^{\infty} \frac{2 \mathrm{z}}{\mathrm{z}^{2}-\left(\mathrm{n}-\frac{1}{2}\right)^{2} \pi^{2}}
$$

and hence deduce

$$
1+\frac{1}{3^{2}}+\frac{1}{5^{2}}+\cdots=\frac{\pi^{2}}{8}
$$

### 6.16 The Gamma Function

The gamma function $\Gamma(\mathrm{z})$ was introduced by Swedish Mathematician L. Euler (17071783), in 1729 while he was seeking for a function of a real variable x which is continuous for positive x and reduces to x ! when x is a positive integer. Gamma function is widely used in the fields of probability and statistics, as well as combinatorics.

Gamma function $\Gamma(\mathrm{z})$ can be introduced in either of the ways :
(i) in terms of infinite product
(ii) in the form of infinite integral
(iii) in limit formula

We establish the form (i) first considering the fact that it possesses simple poles at z $=0,-1,-2, \ldots$ and nowhere vanishes in the entire plane and satisfies

$$
\begin{equation*}
\mathrm{z} \Gamma(\mathrm{z})=\Gamma(\mathrm{z}+1), \Gamma(1)=1 \tag{149}
\end{equation*}
$$

To construct $\Gamma(\mathrm{z})$ we claim that $\mathrm{f}(\mathrm{z})=1 / \Gamma(\mathrm{z})$ is entire with simple zeros at $\mathrm{z}=-\mathrm{n}$ ( $\mathrm{n}=0,1,2, \ldots$ ).

Again we know that $\mathrm{k}=1$ is the largest non-negative integer for which

$$
\sum_{n=1}^{\infty} \frac{1}{n^{k}}
$$

diverges. Then utilizing the Weierstrass Factorization theorem $\mathrm{f}(\mathrm{z})$ can be represented as

$$
f(z)=z e^{g(z)} \prod_{n=1}^{\infty}\left(1+\frac{z}{n}\right) e^{\frac{-z}{n}}
$$

where $\mathrm{g}(\mathrm{z})$ is an entire function, so that gamma function will be of the form

$$
\begin{equation*}
\Gamma(\mathrm{z})=\mathrm{e}^{-\mathrm{g}(\mathrm{z})} \frac{1}{\mathrm{z} \prod_{1}^{\infty}\left(1+\frac{\mathrm{z}}{\mathrm{n}}\right) \mathrm{e}^{-\mathrm{z} / \mathrm{n}}} \tag{150}
\end{equation*}
$$

Now we find $g(z)$ so that (149) hold. We write (150) in the form

$$
\begin{array}{r}
\Gamma(z)=\lim _{n \rightarrow \infty} \frac{e^{-g(z)}}{z \prod_{1}^{n}\left(1+\frac{z}{m}\right) e^{\frac{-z}{m}}} \\
=\lim _{n \rightarrow \infty} \frac{n!\exp \left[-g(z)+\sum_{1}^{n} \frac{z}{m}\right]}{z(z+1) \ldots \ldots(z+n)}=\lim _{n \rightarrow \infty} \Gamma_{n}(z), \text { say }  \tag{151}\\
\frac{z \Gamma_{n}(z)}{\Gamma_{n}(z+1)}=\frac{n!z \exp \left[-g(z)+\sum_{1}^{n} \frac{z}{m}\right]}{z(z+1) \ldots \ldots .(z+n)} \frac{(z+1)(z+2) \ldots \ldots .(z+n+1)}{n!\exp \left[-g(z+1)+\sum_{1}^{n} \frac{z+1}{m}\right]} \\
=(z+n+1) \exp \left[g(z+1)-g(z)-\sum_{1}^{n} \frac{1}{m}\right] \\
=\left(1+\frac{z+1}{n}\right) n \exp \left[g(z+1)-g(z)-\sum_{1}^{n} \frac{1}{m}\right] \\
=\left(1+\frac{z+1}{n}\right) \exp \left[g(z+1)-g(z)-\sum_{1}^{n} \frac{1}{m}+\log n\right]
\end{array}
$$

Now from the relation $\frac{\mathrm{z} \Gamma(\mathrm{z})}{\Gamma(\mathrm{z}+1)}=\lim _{\mathrm{n} \rightarrow \infty} \frac{\mathrm{z} \Gamma_{\mathrm{n}}(\mathrm{z})}{\Gamma_{\mathrm{n}}(\mathrm{z}+1)}$, we find that

$$
\begin{align*}
& \frac{z \Gamma(z)}{\Gamma(z+1)}=\lim _{n \rightarrow \infty}\left(1+\frac{z+1}{n}\right) \exp \left[g(z+1)-g(z)-\sum_{1}^{n} \frac{1}{m}+\log n\right] \\
& =\exp [g(z+1)-g(z)-\gamma] \\
& \quad \gamma=\lim _{n \rightarrow \infty}\left(\sum_{1}^{n} \frac{1}{m}-\log n\right)=0 \cdot 57722 \tag{152}
\end{align*}
$$

where
is known as the Euler's constant.
Thus in order that the conditions in (149) to hold, we should have

$$
\begin{equation*}
\mathrm{g}(\mathrm{z}+1)-\mathrm{g}(\mathrm{z})=\gamma+2 \mathrm{k} \pi \mathrm{i}(\mathrm{k} \equiv \text { integer }) \tag{153}
\end{equation*}
$$

and

$$
1=\Gamma(1)=\lim _{n \rightarrow \infty} \Gamma_{n}(1)=\lim _{n \rightarrow \infty} \frac{e^{-g(1)+\sum_{1} \frac{z}{m}-\log n}}{1+\frac{1}{n}}=e^{-g(1)+\gamma}
$$

so that

$$
\begin{equation*}
\mathrm{g}(1)=\gamma+2 \mathrm{j} \pi \mathrm{i}(\mathrm{j} \equiv \text { integer }) \tag{154}
\end{equation*}
$$

The simplest entire function satisfying (154) is given by

$$
\mathrm{g}(\mathrm{z})=\gamma_{\mathrm{z}}
$$

Finally from (150),

$$
\begin{equation*}
\Gamma(\mathrm{z})=\frac{\mathrm{e}^{-\gamma \mathrm{z}}}{\mathrm{z}} \prod_{1}^{\infty}\left(1+\frac{\mathrm{z}}{\mathrm{n}}\right)^{-1} \mathrm{e}^{\mathrm{z} / \mathrm{n}} \tag{155}
\end{equation*}
$$

## Gauss's Formula

From (151) we have the representation

$$
\begin{align*}
& \Gamma(z)=\lim _{n \rightarrow \infty} \frac{n!\exp \left[\left(\sum_{1}^{n} \frac{1}{m}-\gamma\right) z\right]}{z(z+1) \ldots \ldots(z+n)} \\
& =\lim _{n \rightarrow \infty} \frac{n!\exp \left[\left\{\left(\sum_{1}^{n} \frac{1}{m}-\gamma-\log n\right)+\log n\right\} z\right]}{z(z+1) \ldots \ldots(z+n)} \\
& =\lim _{n \rightarrow \infty} \frac{n!n^{z}}{z(z+1) \ldots .(z+n)}, \text { since } \lim _{n \rightarrow \infty}\left(\sum_{1}^{n} \frac{1}{m}-\log n-\gamma\right)=0 \tag{156}
\end{align*}
$$

The above expression for $\Gamma(\mathrm{z}), \mathrm{z} \neq 0,-1,-2, \ldots$ is termed as Gauss's formula, though it was first derived by Euler.

In many places it is known as Euler's limit formula.
Example 9 : Let

$$
\Gamma(\mathrm{z}, \mathrm{n})=\frac{\mathrm{n}!\mathrm{n}^{\mathrm{z}}}{\mathrm{z}(\mathrm{z}+1) \ldots \ldots(\mathrm{z}+\mathrm{n})}
$$

Prove that

$$
\Gamma(\mathrm{z}, \mathrm{n})=\frac{\mathrm{n}^{\mathrm{z}} \Gamma(\mathrm{n}+1) \Gamma(\mathrm{z})}{\Gamma(\mathrm{n}+\mathrm{z}+1)}
$$

and hence deduce that

$$
\frac{\mathrm{n}^{\mathrm{z}} \Gamma(\mathrm{n})}{\Gamma(\mathrm{n}+\mathrm{z})} \rightarrow 1 \text { as } \mathrm{n} \rightarrow \infty
$$

## Solution :

$$
\Gamma(\mathrm{n}+\mathrm{z}+1)=\mathrm{z}(\mathrm{z}+1)(\mathrm{z}+2) \ldots \ldots(\mathrm{z}+\mathrm{n}) \Gamma(\mathrm{z})
$$

so, $\frac{\mathrm{n}^{\mathrm{z}} \Gamma(\mathrm{n}+1) \Gamma(\mathrm{z})}{\Gamma(\mathrm{n}+\mathrm{z}+1)}=\frac{\mathrm{n}^{\mathrm{z}} \Gamma(\mathrm{n}+1)}{\mathrm{z}(\mathrm{z}+1)(\mathrm{z}+2) \ldots \ldots(\mathrm{z}+\mathrm{n})}=\frac{\mathrm{n}!\mathrm{n}^{\mathrm{z}}}{\mathrm{z}(\mathrm{z}+1)(\mathrm{z}+2) \ldots \ldots(\mathrm{z}+\mathrm{n})}=\Gamma(\mathrm{z}, \mathrm{n})$
Now,

$$
\begin{aligned}
& \frac{n^{z} \Gamma(n)}{\Gamma(n+z)}=\frac{(n+z) \Gamma(z, n)}{n \Gamma(z)} \\
& \lim _{n \rightarrow \infty} \frac{n^{z} \Gamma(n)}{\Gamma(n+z)}=\lim _{n \rightarrow \infty}\left(1+\frac{z}{n}\right) \frac{\lim _{n \rightarrow \infty} \Gamma(z, n)}{\Gamma(z)}=1 \text { by Gauss's formula. }
\end{aligned}
$$

In the expression (155) for $\Gamma(z)$ the infinite product is uniformly convergent on every compact subset of $\mathbb{C}-\{0,-1, \ldots . .$.$\} . So calculating \Gamma^{\prime}(\mathrm{z}) / \Gamma(\mathrm{z})$ we find that

$$
\frac{\Gamma^{\prime}(\mathrm{z})}{\Gamma(\mathrm{z})}=-\gamma-\frac{1}{\mathrm{z}}+\sum_{\mathrm{n}=1}^{\infty}\left(-\frac{1}{\mathrm{n}+\mathrm{z}}+\frac{1}{\mathrm{n}}\right)
$$

This function $\frac{\Gamma^{\prime}(\mathrm{z})}{\Gamma(\mathrm{z})}$ is denoted by $\psi(\mathrm{z})$ and named as Gaussian psi function and it is seen from its expression that $\psi$ is meromorphic in $\mathbb{C}$ with simple poles at $\mathrm{z}=0,-1,-$ $2, \ldots$ and $\operatorname{Res}(\psi ;-\mathrm{n})=-1$ for $\mathrm{n}=0,1,2, \ldots$

Example 10 : Show that
(i) $\psi(1)=-\gamma$
(ii) $\psi(\mathrm{z}+1)-\psi(\mathrm{z})=\frac{1}{\mathrm{z}}$
(iii) $\psi(\mathrm{z})-\psi(1-\mathrm{z})=-\pi \cot \pi \mathrm{z}$.

## Solution :

$$
\begin{equation*}
\psi(\mathrm{z})=-\gamma-\frac{1}{\mathrm{z}}+\sum_{\mathrm{n}=1}^{\infty}\left(-\frac{1}{\mathrm{n}+\mathrm{z}}+\frac{1}{\mathrm{n}}\right) \tag{i}
\end{equation*}
$$

so,

$$
\begin{aligned}
\psi(1) & =-\gamma-1+\sum_{\mathrm{n}=1}^{\infty}\left(-\frac{1}{\mathrm{n}+1}+\frac{1}{\mathrm{n}}\right) \\
& =-\gamma-1+\left(1-\frac{1}{2}+\frac{1}{2}-\frac{1}{3}+\frac{1}{3}-\cdots\right) \\
& =-\gamma
\end{aligned}
$$

(ii) $\psi(z+1)-\psi(z)=-\gamma-\frac{1}{z+1}+\sum_{n=1}^{\infty}\left(-\frac{1}{n+z+1}+\frac{1}{n}\right)-\sum_{n=1}^{\infty}\left(-\frac{1}{n+z}+\frac{1}{n}\right)+\gamma+\frac{1}{z}$

$$
\begin{aligned}
& =\frac{1}{\mathrm{Z}}-\frac{1}{\mathrm{z}+1}+\sum_{\mathrm{n}=-1}^{\infty}\left(\frac{1}{\mathrm{n}+\mathrm{z}}-\frac{1}{\mathrm{n}+\mathrm{z}+1}\right) \\
& =\frac{1}{\mathrm{z}}-\frac{1}{\mathrm{z}+1}+\left(\frac{1}{\mathrm{z}+1}-\frac{1}{\mathrm{z}+2}+\frac{1}{\mathrm{z}+2}-\frac{1}{\mathrm{z}+3}+\cdots\right) \\
& =\frac{1}{\mathrm{z}}
\end{aligned}
$$

(iii) $\psi(z)-\psi(1-z)=-\frac{1}{z}+\frac{1}{1-z}+\sum_{1}^{\infty}\left(\frac{1}{\mathrm{n}}-\frac{1}{\mathrm{n}+\mathrm{z}}\right)-\sum_{1}^{\infty}\left(\frac{1}{\mathrm{n}}-\frac{1}{\mathrm{n}+1-\mathrm{z}}\right)$

$$
\begin{aligned}
& =-\frac{1}{\mathrm{z}}-\frac{1}{\mathrm{z}-1}+\sum_{1}^{\infty}\left(\frac{1}{\mathrm{n}+1-\mathrm{z}}-\frac{1}{\mathrm{n}+\mathrm{z}}\right) \\
& =-\frac{1}{\mathrm{z}}-\frac{1}{\mathrm{z}-1}-\frac{1}{\mathrm{z}+1}-\frac{1}{\mathrm{z}-2}-\frac{1}{\mathrm{z}+2}-\cdots \\
& =-\frac{1}{\mathrm{z}}-\left(\frac{1}{\mathrm{z}-1}+\frac{1}{\mathrm{z}+1}\right)-\left(\frac{1}{\mathrm{z}-2}+\frac{1}{\mathrm{z}+2}\right)-\cdots \\
& =-\frac{1}{\mathrm{z}}-\sum_{1}^{\infty} \frac{2 \mathrm{z}}{\mathrm{z}^{2}-\mathrm{n}^{2}}=-\pi \cot \pi \mathrm{z}, \text { by }(141)
\end{aligned}
$$

### 6.17 A Few Properties of $\Gamma(z)$

We have

$$
\frac{1}{\Gamma(\mathrm{z})}=\mathrm{e}^{\gamma \mathrm{z}} \prod_{1}^{\infty}\left(1+\frac{\mathrm{z}}{\mathrm{n}}\right) \mathrm{e}^{-\mathrm{z} / \mathrm{n}}
$$

Hence,

$$
\frac{1}{\Gamma(\mathrm{z}) \Gamma(-\mathrm{z})}=-\mathrm{z}^{2} \prod_{1}^{\infty}\left(1-\frac{\mathrm{z}^{2}}{\mathrm{n}^{2}}\right)
$$

$$
\begin{aligned}
& =-\frac{\mathrm{z}}{\pi} \pi \mathrm{z} \prod_{1}^{\infty}\left(1-\frac{\mathrm{z}^{2}}{\mathrm{n}^{2}}\right) \\
& =-\frac{\mathrm{z}}{\pi} \sin \pi \mathrm{z}
\end{aligned}
$$

or, $\quad \frac{1}{\Gamma(\mathrm{z})[-\mathrm{z} \Gamma(-\mathrm{z})]}=\frac{\sin \pi \mathrm{z}}{\pi}$
i.e. $\quad \frac{1}{\Gamma(z) \Gamma(1-z)}=\frac{\sin \pi z}{\pi}, \quad[$ using $\quad z \Gamma(z)=\Gamma(z+1)$ i.e., $-z \Gamma(-z)$
$=\Gamma(1-\mathrm{z})]$
In particular, $\left[\Gamma\left(\frac{1}{2}\right)\right]^{2}=\pi$ and $\Gamma\left(\frac{1}{2}\right)=\sqrt{\pi}$ (minus sign is excluded since $\Gamma\left(\frac{1}{2}\right)$ is positive by (155)). Likewise using

$$
\Gamma(\mathrm{z}+1)=\mathrm{z} \Gamma(\mathrm{z})
$$

we find

$$
\begin{aligned}
& \Gamma\left(\frac{3}{2}\right)=\frac{1}{2} \Gamma\left(\frac{1}{2}\right)=\frac{1}{2} \sqrt{\pi} \\
& \Gamma\left(\frac{5}{2}\right)=\frac{3}{2} \Gamma\left(\frac{3}{2}\right)=\frac{3}{2} \cdot \frac{1}{2} \sqrt{\pi} \\
& \Gamma\left(\frac{7}{2}\right)=\frac{5}{2} \Gamma\left(\frac{5}{2}\right)=\frac{5}{2} \cdot \frac{3}{2} \cdot \frac{1}{2} \sqrt{\pi}
\end{aligned}
$$

and in general

$$
\Gamma\left(\mathrm{n}+\frac{1}{2}\right)=\frac{1.3 \ldots \ldots(2 \mathrm{n}-1)}{2^{\mathrm{n}}} \sqrt{\pi},(\mathrm{n}=1,2, \cdots
$$

i.e. $\quad \Gamma\left(n+\frac{1}{2}\right) / \sqrt{\pi}=\frac{(2 n)!}{n!(2)^{2 n}}$

If n is a positive integer repeated use of (149) produce

$$
\Gamma(\mathrm{n}+1)=\mathrm{n}!
$$

The $\Gamma$-function can therefore be considered as an extension of the factorial function to the complex plane.

## Legendre's Duplication Formula

Let us consider the Gauss's formula

$$
\Gamma(z)=\lim _{n \rightarrow \infty} \frac{n!n^{z}}{z(z+1) \ldots .(z+n)}=\lim _{n \rightarrow \infty} \Gamma(z, n) \text {, say }
$$

Then,

$$
\begin{aligned}
\Gamma(2 \mathrm{z}, 2 \mathrm{n}) & =\frac{(2 \mathrm{n})!(2 \mathrm{n})^{2 \mathrm{z}}}{2 \mathrm{z}(2 \mathrm{z}+1) \ldots(2 \mathrm{z}+\mathrm{n}) \ldots \cdot(2 \mathrm{n}+2 \mathrm{z})} \\
& =\frac{2^{2 \mathrm{n}} \mathrm{n}!\Gamma\left(\mathrm{n}+\frac{1}{2}\right)(\sqrt{\pi})^{-1}(2 \mathrm{n})^{2 z}}{2 \mathrm{z}(2 \mathrm{z}+1)(2 \mathrm{z}+2) \ldots \ldots \cdot(2 \mathrm{z}+2 \mathrm{n})}[\text { Replacing }(2 \mathrm{n})!\text { by (158)] }
\end{aligned}
$$

$$
=\frac{2^{2 z-1} n!(n)^{2 z} \Gamma\left(n+\frac{1}{2}\right)}{\sqrt{\pi} z(z+1)(z+2) \ldots(z+n)\left(z+\frac{1}{2}\right)\left(z+\frac{3}{2}\right) \ldots\left(z+n-\frac{1}{2}\right)}
$$

$$
=\frac{2^{2 \mathrm{z}-1}}{\sqrt{\pi}} \Gamma(\mathrm{z}, \mathrm{n}) \Gamma\left(\mathrm{n}+\frac{1}{2}\right) \frac{1}{\left(\mathrm{z}+\frac{1}{2}\right)\left(\mathrm{z}+\frac{3}{2}\right) \ldots\left(\mathrm{z}+\mathrm{n}-\frac{1}{2}\right)}
$$

$$
=\frac{2^{2 z-1}}{\sqrt{\pi}} \Gamma(\mathrm{z}, \mathrm{n}) \Gamma\left(\mathrm{n}+\frac{1}{2}\right) \frac{\Gamma\left(\mathrm{z}+\frac{1}{2}, \mathrm{n}\right)}{\mathrm{n}^{1 / 2} \Gamma(\mathrm{n})} \frac{\mathrm{z}+\frac{1}{2}+\mathrm{n}}{\mathrm{n}}
$$

and $\Gamma(2 z)=\lim _{n \rightarrow \infty} \Gamma(2 z, 2 n)=\frac{2^{2 z-1}}{\sqrt{\pi}} \Gamma(z) \Gamma\left(z+\frac{1}{2}\right) \lim _{n \rightarrow \infty}\left[\frac{\Gamma\left(n+\frac{1}{2}\right)}{n^{1 / 2} \Gamma(n)} \frac{z+\frac{1}{2}+n}{n}\right]$

$$
\left.=\frac{2^{2 \mathrm{z}-1}}{\sqrt{\pi}} \Gamma(\mathrm{z}) \Gamma\left(\mathrm{z}+\frac{1}{2}\right) \text { [using example } 9\right]
$$

So that

$$
\begin{equation*}
\sqrt{\pi} \Gamma(2 \mathrm{z})=2^{2 \mathrm{z}-1} \Gamma(\mathrm{z}) \Gamma\left(\mathrm{z}+\frac{1}{2}\right) \tag{1599}
\end{equation*}
$$

## This is known as Legendre's duplication formula.

## Residue of $\Gamma(z)$ at its poles

$\Gamma(\mathrm{z})$ is analytic throughout the complex plane except at its only singularities which are simple poles situated at $\mathrm{z}=0,-1,-2, \ldots$. That is $\Gamma(\mathrm{z})$ is analytic in the right half of the complex plane $\operatorname{Re} z>0$. Using the fact that $\mathrm{z} \Gamma(\mathrm{z})=\Gamma(\mathrm{z}+1)$, we have
$\Gamma(\mathrm{z}+\mathrm{n}+1)=(\mathrm{z}+\mathrm{n})(\mathrm{z}+\mathrm{n}-1)(\mathrm{z}+\mathrm{n}-2) \ldots(\mathrm{z}+1) \mathrm{z} \Gamma(\mathrm{z}), \mathrm{n} \equiv$ positive integer and

$$
\Gamma(\mathrm{z})=\frac{\Gamma(\mathrm{z}+\mathrm{n}+1)}{\mathrm{z}(\mathrm{z}+1) \ldots \ldots(\mathrm{z}+\mathrm{n}-1)(\mathrm{z}+\mathrm{n})}
$$

$\operatorname{Res}(\Gamma(\mathrm{z}) ;-\mathrm{n})=\lim _{\mathrm{z} \rightarrow-\mathrm{n}}(\mathrm{z}+\mathrm{n}) \Gamma(\mathrm{z})$

$$
\begin{aligned}
& =\lim _{\mathrm{z} \rightarrow-\mathrm{n}} \frac{\Gamma(\mathrm{z}+\mathrm{n}+1)}{\mathrm{Z}(\mathrm{z}+1) \ldots(\mathrm{z}+\mathrm{n}-1)} \\
& =\frac{(-1)^{\mathrm{n}}}{\mathrm{n}!}, \mathrm{n}=0,1,2, \ldots
\end{aligned}
$$

## Integral representation of $\Gamma(z)$

Theorem : Prove that

$$
\Gamma(\mathrm{z})=\int_{0}^{\infty} \mathrm{e}^{-\mathrm{t}} \mathrm{t}^{\mathrm{z}-1} \mathrm{dt} \text { for } \operatorname{Re} \mathrm{z}>0
$$

Proof. Let

$$
\mathrm{F}_{\mathrm{n}}(\mathrm{z})=\frac{\mathrm{n}!\mathrm{n}^{2}}{\mathrm{z}(\mathrm{z}+1) \ldots(\mathrm{z}+\mathrm{n})}
$$

We prove the theorem in the following two steps :
(i) $\mathrm{F}_{\mathrm{n}}(\mathrm{z})=\int_{0}^{\mathrm{n}}\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}} \mathrm{t}^{\mathrm{z}-1} \mathrm{dt}$
(ii) $\lim _{\mathrm{n} \rightarrow \infty} \int_{0}^{\mathrm{n}}\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}} \mathrm{t}^{\mathrm{z-1}} \mathrm{dt}=\int_{0}^{\infty} \mathrm{e}^{-\mathrm{t}} \mathrm{t}^{\mathrm{z-1}} \mathrm{dt}$

To establish (i) we change the variable $t$ to $n s$ in

$$
\int_{0}^{\mathrm{n}}\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}} \mathrm{t}^{\mathrm{z}-1} \mathrm{dt}
$$

to obtain

$$
\int_{0}^{\mathrm{n}}\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}} \mathrm{t}^{\mathrm{z}-1} \mathrm{dt}=\mathrm{n}^{\mathrm{z}} \int_{0}^{1}(1-\mathrm{s})^{\mathrm{n}} \mathrm{~s}^{\mathrm{z-1}} \mathrm{ds}
$$

Now integrating by parts we find the right hand side is equal to
$\mathrm{n}^{\mathrm{z}}\left[\left.\frac{1}{\mathrm{z}} \mathrm{s}^{\mathrm{z}}(1-\mathrm{s})^{\mathrm{n}}\right|_{0} ^{1}+\frac{\mathrm{n}}{\mathrm{z}} \int_{0}^{1}(1-\mathrm{s})^{\mathrm{n}-1} \mathrm{~s}^{\mathrm{z}} \mathrm{ds}\right]$
$=\mathrm{n}^{\mathrm{z}} \frac{\mathrm{n}}{\mathrm{z}} \int_{0}^{1}(1-\mathrm{s})^{\mathrm{n}-1} \mathrm{~s}^{\mathrm{z}} \mathrm{ds}$
$=\mathrm{n}^{\mathrm{z}} \frac{\mathrm{n} .(\mathrm{n}-1) \ldots .1}{\mathrm{z}(\mathrm{z}+1) \ldots .(\mathrm{z}+\mathrm{n}-1)} \int_{0}^{1} \mathrm{~s}^{\mathrm{z}+\mathrm{n}-1} \mathrm{ds}$ [Integrating by parts $(\mathrm{n}-1)$ times]
$=\frac{\mathrm{n}!\mathrm{n}^{2}}{\mathrm{z}(\mathrm{z}+1) \ldots(\mathrm{z}+\mathrm{n})}=\mathrm{F}_{\mathrm{n}}(\mathrm{z})$
Now to prove (ii)we show that

$$
\begin{equation*}
\lim _{n \rightarrow \infty} \int_{0}^{n}\left[e^{-t}-\left(1-\frac{t}{n}\right)^{n}\right]^{t-1} d t=0, \operatorname{Re} z>0 \tag{161}
\end{equation*}
$$

For this, note that

$$
\begin{equation*}
1+\frac{t}{n} \leq e^{\frac{t}{n}} \leq \frac{1}{1-\frac{t}{n}} \text { for }|t|<n \tag{162}
\end{equation*}
$$

Then, $\quad\left(1+\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}} \leq \mathrm{e}^{\mathrm{t}}$ and $\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}} \leq \mathrm{e}^{-\mathrm{t}}$;
Consequently,

$$
\begin{aligned}
& 0 \leq \mathrm{e}^{-\mathrm{t}}-\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}}=\mathrm{e}^{-\mathrm{t}}\left[1-\mathrm{e}^{\mathrm{t}}\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}}\right] \leq \mathrm{e}^{-\mathrm{t}}\left[1-\left(1-\frac{\mathrm{t}^{2}}{\mathrm{n}^{2}}\right)^{\mathrm{n}}\right] \\
& =\mathrm{e}^{-\mathrm{t}} \frac{\mathrm{t}^{2}}{\mathrm{n}^{2}}\left[1+\left(1-\frac{\mathrm{t}^{2}}{\mathrm{n}^{2}}\right)+\ldots+\left(1-\frac{\mathrm{t}^{2}}{\mathrm{n}^{2}}\right)^{\mathrm{n}-1}\right] \leq \mathrm{e}^{-\mathrm{t}} \frac{\mathrm{t}^{2}}{\mathrm{n}} .
\end{aligned}
$$

Therefore,

$$
\left|\int_{0}^{\mathrm{n}}\left[\mathrm{e}^{-\mathrm{t}}-\left(1-\frac{\mathrm{t}}{\mathrm{n}}\right)^{\mathrm{n}}\right] \mathrm{t}^{\mathrm{z}-1} \mathrm{dt}\right|<\frac{1}{\mathrm{n}} \int_{0}^{\mathrm{n}} \mathrm{e}^{-\mathrm{t}} \mathrm{t}^{\mathrm{Re} \mathrm{z+1}} \mathrm{dt}
$$

which approaches zero as $\mathrm{n} \rightarrow \infty$ because the integral on the right converges. This completes the proof of (ii). Finally combining the results (i) and (ii) with the Gauss's formula (156) we get

$$
\Gamma(z)=\lim _{n \rightarrow \infty} F_{n}(z)=\lim _{n \rightarrow \infty} \int_{0}^{n}\left(1-\frac{t}{n}\right)^{n} t^{z-1} d t=\int_{0}^{\infty} e^{-t} t^{z-1} d t
$$

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## NETAJI SUBHAS OPEN UNIVERSITY

## PREFACE

In the curricular structure introduced by this University for students of Post-Graduate degree programme, the opportunity to pursue Post-Graduate course in Subjects introduced by this University is equally available to all learners. Instead of being guided by any presumption about ability level, it would perhaps stand to reason if receptivity of a learner is judged in the course of the learning process. That would be entirely in keeping with the objectives of open education which does not believe in artificial differentiation.

Keeping this in view, study materials of the Post-Graduate level in different subjects are being prepared on the basis of a well laid-out syllabus. The course structure combines the best elements in the approved syllabi of Central and State Universities in respective subjects. It has been so designed as to be upgradable with the addition. of new information as well as results of fresh thinking and analysis.

The accepted methodology of distance education has been followed in the preparation of these study materials. Co-operation in every form of experienced scholars is indispensable for a work of this kind. We, therefore, owe an enormous debt of gratitude to everyone whose tireless efforts went into the writing, editing and devising of a proper lay-out of the materials. Practically speaking, their role amounts to an involvement in invisible teaching. For, whoever makes use of these study materials would virtually derive the benefit of lcarning under their collective care without each being seen by the other.

The more a learner would seriously pursue these study materials the easier it will be for him or her to reach out to larger horizons of a subject. Care has also been taken to make the language lucid and presentation attractive so that it may be rated as quality self-learning materials. If anything remains still obscure or difficult to follow, arrangements are there to come to terms with them through the counselling sessions regularly available at the network of study centres set up by the University.

Needless to add, a great part of these efforts is still experimental-in fact, pioneering in certain areas. Naturally, there is every possibility of some lapse or deficiency here and there. However, these do admit of rectification and further improvement in due course. On the whole, therefore, these study materials are expected to evoke wider appreciation the more they receive serious attention of all concerned.

Professor (Dr:) Subha Sankar Sarkar<br>Vice-Chancellor

Fifth Reprint : July, 2017

Printed in accrodance with the regulations and financial assistance of the Distance Education Bureau of the Uiversity Grants Commission.

## Subject : Mathematics

## Post Graduate

## Paper : PG (MT) : IX A (II)

Writer
Prof. T. K. Pal

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## Unit 1 Classical Optimization Techniques

## Structure

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1.2 Multivariable optimization with no constraints
1.1 Introduction
1.2 Multivariable optimization with no constraints

## 1.1 (Introduction)

The methods of determining relative extrema of functions of several variables using differential calculus are so old and well-known that they are referred to as classical. The classical methods of optimization are used in finding the optimum of continuous and differentiable functions. Since practical problems involve objective functions that are not continuous and/or differentiable, the classical optimization techniques have limited people of applications. But these classical techniques forms a basis for developing most of the numerical techniques of optimization.

In this unit we consider three types of problems viz
(i) Multivariable optimization with no constraints.
(ii) Multivariable optimization with equality constraints and
(iii) Multivariable optimization with inequality constraints

### 1.2 Multivariable optimization with no constraints

We develop the necessary and sufficient conditions for an $n$-variable functions $\mathrm{f}(\mathrm{x})$ to have extremt. It is assumed that the first and second partial derivatives of $f(\mathrm{x})$ are continuous at every $x$.

Theorem 1.2.1 A necessary condition for $x_{0}$ to be an extreme point of $f(x)$ is that $\nabla \mathrm{f}\left(\mathrm{x}_{0}\right)=0$ i.e. $\left[\frac{\partial t}{\partial x_{i}}\right]_{\mathrm{v}_{0}}=0$ for $\mathrm{i}=1,2, \ldots . . \mathrm{n}$.

Proof : By Taylor's theorem we have

$$
\begin{equation*}
f\left(\mathrm{X}_{0}+h\right)=f\left(\mathrm{X}_{0}\right)+\sum_{i=1}^{n} h_{i}\left[\frac{\partial t}{\partial x_{i}}\right]_{x_{0}}+\frac{1}{2} \sum_{i=1}^{n} \sum_{j=1}^{n} h_{i} h_{j}\left[\frac{\partial^{2} f}{\partial x_{i} \partial x_{j}}\right]_{x_{0} o h} . \tag{1}
\end{equation*}
$$

where $0<\theta<1$.
Since the last term is of order $h_{j}^{2}$, the terms of order $h$ will dominate the higher order terms for small $h$. Thus the sign of $f\left(\mathrm{X}_{0}+h\right)-f\left(\mathrm{X}_{0}\right)$ is decided by the sign of $\sum_{i=1}^{n} h_{i}\left[\frac{\partial t}{\partial x_{i}}\right]_{x_{0}}$. Let $\mathrm{X}_{0}$ be are extreme point, say miximum point. Then $f\left(\mathrm{X}_{0}+h\right)$ $-f\left(\mathrm{X}_{0}\right)>0$ for all sufficiently small h . We are to show that $\left[\frac{\partial f}{\partial x_{i}}\right]_{r_{0}}=0 \forall \mathrm{i}=1$. $2, \ldots . . ., n$. If possible, let $\left[\frac{\partial f}{\partial x_{k}}\right]_{x_{0}} \neq 0$.

Let us choose $h_{\mathrm{i}}=0$ for all $\mathrm{i} \neq k$, and $\mathrm{h}_{\mathrm{k}}$ sufficiently small. Then the sign of $f\left(X_{0}+h\right)-f\left(X_{0}\right)$ is decided by the sign of $h_{k}\left[\frac{\partial f}{\partial x_{k}}\right]_{x_{0}}$. Since $\left[\frac{\partial f}{\partial x_{k}}\right]_{x_{0}} \neq=0$, let $\left[\frac{\partial f}{\partial x_{k}}\right]_{x_{0}}>0$. Then $f\left(\mathrm{X}_{0}+h\right)-f\left(\mathrm{X}_{0}\right)$ will be positive for $h_{k}>0$ and negative for $h_{k}<0$. This is a contradiction as $x_{0}$ is a minimum point. Similar contradiction occurs for $\left[\frac{\partial f}{\partial x_{k}}\right]_{x_{0}}<0$. Hence $\left[\frac{\partial f}{\partial x_{k}}\right]_{x_{0}} \neq 0$ is not possible. $\therefore\left[\frac{\partial f}{\partial x_{k}}\right]_{x_{0}}=0$. This is true for any $k=1,2, \ldots n$. Hence the theorem.

Theorem 1.2.2 A sufficient condition for a stationery point $x_{0}$ to be an extremum is that
(i) $\nabla f\left(\mathrm{X}_{0}\right)=0$ and the Hessian matrix $[\mathrm{H}]_{x_{0}}$ is positive definite when $x_{0}$ is a minumum point.
(ii) $\nabla f\left(\mathrm{X}_{0}\right)=0$ and the Hessian matrix $[\mathrm{H}]_{v_{0}}$ is negative definite when $x_{0}$ is a maximum point.

Prob: By Tayloris theorem we have
$f\left(\mathrm{X}_{0}+\cdot h\right)=f\left(\mathrm{X}_{0}\right)+\sum_{i=1}^{n} h_{i}\left[\frac{\partial f}{\partial x_{i}}\right]_{x_{0}}+\frac{1}{[2} \sum_{j=1}^{n} \sum_{j=1}^{n} h_{i} h_{j}\left[\frac{\partial^{2} f}{\partial x_{i} \partial x_{j}}\right]_{x_{0}+o b h}$
Where $0<\theta<1$.
$f\left(X_{0}+h\right)-f\left(X_{0}\right)=Q\left(x_{0}+o h\right)$
Where Q $\left(x_{0}+o h\right)=\frac{1}{2} \sum_{i=1}^{n} \sum_{j=1}^{n} h_{i} h_{j}\left[\frac{\partial^{2} f}{\partial x_{i} \partial x_{j}}\right]_{x_{0}+o h}$
Now we have assumed that the second order partial derivative $\frac{\partial^{2} f}{\partial x_{i} \partial x_{j}}$ is continuous in the neighbourhord of $x_{0}$. So far sufficiently small $h$, the signs of $\mathrm{Q}\left(x_{0}\right)$ + oh) and $\mathrm{Q}\left(x_{0}\right)$ are some. Hence $f\left(\mathrm{X}_{0}+h\right)-f\left(\mathrm{X}_{0}\right)$ and $\mathrm{Q}\left(x_{0}\right)$ have the same sign.

Let $\mathrm{J}\left(\mathrm{X}_{0}\right)$ be the Hessian matrix $\left[\left.\frac{\partial f}{\partial x_{i} \partial x_{j}}\right|_{x_{0}}\right]$. From matrix algebra we know that $\mathrm{Q}\left(\mathrm{X}_{0}\right)=\frac{1}{2} \sum_{j=1}^{n} \sum_{j=1}^{n} h_{i} h_{j}\left[\frac{\partial^{2} f}{\partial x_{i} \partial x_{j}}\right]_{x_{0}}$ will be positive (negative) for al $h$ if and only if the Hessian matrix $J\left(X_{0}\right)$ is positive definite (negative definite) at $X=X_{0}$.

Thus for sufficiently small $h$, the sign of $f\left(X_{0}+h\right)-f\left(X_{0}\right)$ is positive (negative) if $J\left(X_{0}\right)$ is positive definite (negative definite) i.e., $X_{0}$ is a relative minimum (maximum) ifJ $\left(X_{0}\right)$ is positive definite (negative definite). Hence the theorem,

Result : Let $A=\left[a_{i j}\right]_{1 \times n}$ and
$\mathrm{A}_{1}=\mathrm{a}_{11}, \quad \mathrm{~A}_{2}=\left|\begin{array}{ll}a_{11} & a_{12} \\ a_{13} & a_{14}\end{array}\right|, \mathrm{A}_{3}=\left|\begin{array}{lll}a_{11} & a_{12} & a_{13} \\ a_{21} & a_{22} & a_{23} \\ a_{31} & a_{32} & a_{33}\end{array}\right|$

Then the matrix A is
(i) positive definite iff $A_{i}>0$ for all $i=1,2, \ldots \ldots . n$
(ii) negative definite iff the sign of Ai is $(-\mathrm{i}) \mathrm{i}$ for $\mathrm{i}=1,2, \ldots . . \mathrm{n}$.
(iii) positive semidefinite iff $\mathrm{A}_{\mathrm{i}} \geq 0$ for all $\mathrm{i}=1,2, \ldots \ldots . \mathrm{n}$ wih equality holding for at least one i
(iv) negative semidefinite iff $\mathrm{Ai} \leq \mathrm{o}$ for all $\mathrm{i}=1,2, \ldots . . \mathrm{n}$ with equality holding for at least one i
(v) indefinite if it is neither definite nor semidefinite,

Example 1.2.1 Determine the extreme points of the function

$$
f\left(x_{1}, x_{2}\right)=x_{1}^{3}+x_{2}^{3}+4 x_{x}^{2}+2 x_{2}^{2}+12
$$

## Solution :

$$
\text { Here } \frac{\partial f}{\partial x_{1}}=3 x_{1}^{2}+8 x_{1}, \frac{\partial f}{\partial x_{2}}=3 x_{2}^{2}+4 x_{2}
$$

The necessary condition for the existence of an extreme points gives

$$
x_{1}\left(3 x_{1}+8\right)=0 \text { and } x_{2}\left(3 x_{2}+4\right)=0
$$

The solutions are $(0,0)(0,-4 / 3),(-8 / 3,0<(-8 / 3,-4 / 3)$. The Hessian matrix of $f\left(x_{1}, x_{2}\right)$ is given by

$$
\begin{aligned}
& \mathrm{J}\left(x_{1}, x_{2}\right)=\left|\begin{array}{cc}
\frac{\partial^{2} f}{d x_{1}^{2}} & \frac{\partial^{2} f}{d x_{1} \partial x_{2}} \\
\frac{\partial^{2} f}{\partial x_{2} \partial x_{1}} & \frac{\partial^{2} f}{\partial x_{2}^{2}}
\end{array}\right|=\left|\begin{array}{cc}
6 x_{1}+8 & 0 \\
0 & 6 x_{2}+4
\end{array}\right| \\
& \therefore \mathrm{J}_{1}=6 x_{1}+8
\end{aligned}
$$

$$
\text { and } J_{2}=\left|\begin{array}{cc}
6 x_{1}+8 & 0 \\
0 & 6 x_{2}+4
\end{array}\right|=\left(6 x_{1}+8\right)\left(6 x_{2}+4\right)
$$

For the point, $(0,0)$ we have
$\mathrm{J}_{1}=6.0+8=8>0$ and $\mathrm{J}_{2}=(6.0+8)(6.0+4)=32>0$
$\therefore \mathrm{J}$ is positive definite. Hence $(0,0)$ is a relative minimum point of $\mathrm{f}\left(x_{1}, x_{2}\right)$
For the point $(0,-4 / 3)$ we have
$\mathrm{J}_{1}=6.0+8=8>0$ and $\mathrm{J}_{2}=(6.0+8)(-6.4 / 3+4)=-32<0$
$\therefore \mathrm{J}$ is indefinite. Hence $(0,-4.3)$ is a saddle point of $\mathrm{f}\left(x_{1}, x_{2}\right)$.
For the point $(-8 / 3,0)$ we have
$\mathrm{J}_{1}=-6.8 / 3+8=-8<0$ and $\mathrm{J}_{2}=(-6.8 / 3+8)(6.0+4)=-32<0$.
$\therefore \mathrm{J}$ is indefinite. Hence $(-8 / 3,0)$ is a paddle point of $\mathrm{f}\left(x_{1}, x_{2}\right)$.
For the point $(-8 / 3,-4 / 3)$ we have
$\mathrm{J}_{1}=-6.8 / 3+8=-8<0$ and $\mathrm{J}_{2}=(-6.8 / 3+8)(-6.4 / 3+4)=32>0$
$\therefore \mathrm{J}$ is negative definite. Hence $(-8 / 3,-4 / 3)$ is a relative maximum point of $f\left(x_{1}\right.$, $x_{2}$ ).

### 1.3 Multivariable optimization with equality constraints

We shall consider two methods viz
(i) Method of constrained variation and
(ii) Method of Lagrange multipliers.

The general multivariable optimization problem with equality constraints is Minimize $f=f(X)$
subject to $\mathrm{g}_{\mathrm{i}}(\mathrm{X})=0, \mathrm{i}=1,2, \ldots . ., m$
Where $\mathrm{X}=\left[x_{1}, x_{2}, \ldots . . \mathrm{xn}\right]^{\top},(m<n)$

### 1.3.1 Method of constrained variation

To understand the salient features of the method we consider the simple problem
Minimize $\mathrm{f}\left(x_{1}, x_{2}\right)$
subject to $g\left(x_{1}, x_{2}\right)=0$
Let us assume that $\mathrm{g}\left(x_{1}, x_{2}\right)=0$ can be solved to obtain $x_{2}$ as $x_{2}=\mathrm{h}\left(x_{1}\right)$. Then the problem reduces to the unconstrained minimization problem Minimize $\mathrm{f}\left(x_{1}, \mathrm{~h}\left(x_{1}\right)\right)$
The necessary condition gives

$$
\frac{d f}{d x_{1}}=0
$$

or, $\quad \frac{\partial f}{\partial x_{1}}+\frac{\partial f}{\partial h} \cdot \frac{d h}{d x_{1}}=0$
or, $\quad \frac{\partial f}{\partial x_{1}}+\frac{\partial f}{\partial x_{2}} \frac{d x_{2}}{d x_{1}}=0$
or, $\quad \frac{\partial f}{\partial x_{1}} d x_{1}+\frac{\partial f}{\partial x_{2}} d x_{2}=0$.
Let $\left(x_{1}^{*}, x_{2}^{*}\right)$ be the minimum point. Then $\left(x_{1}^{*}, x_{2}^{*}\right)$ must satisfy the given constraint.

$$
\begin{equation*}
\therefore\left(x_{1}^{*}, x_{2}^{*}\right)=0 \tag{2}
\end{equation*}
$$

For admissible variations $d x_{1}, d x_{2}$ we have $g\left(x_{1}^{*}+d x_{1}, x_{2}^{*}+d x_{2}\right)=0$
Using Taylor's theorem we get

$$
g\left(x_{1}^{*}, v_{2}^{*}\right)+\left[\frac{\partial g}{\partial x_{1}}\right]_{\left(x_{1}^{*}, x_{2}^{*}\right)} d x_{1}+\left[\frac{\partial g}{\partial x_{2}}\right]_{\left(x_{1}^{*}, x_{2}^{*}\right)} d x_{2}=0
$$

or, $\quad\left[\frac{\partial g}{\partial x_{1}}\right]_{\left(x_{1}^{*}, x_{2}^{*}\right)} d x_{1}+\left[\frac{\partial g}{\partial x_{2}}\right]_{\left(x_{1}^{*}, x_{2}^{*}\right)} d x_{2}=0 \quad[$ by (2) $]$
Assuming $\left[\frac{\partial g}{\partial x_{2}}\right]_{(x, \ldots 2)} \neq 0$ we get,

$$
\begin{equation*}
d x_{2}=-\frac{\left[\frac{\partial g}{\partial x_{2}}\right]_{\left(x_{1}^{*}, x_{2}^{*}\right)}}{\left[\frac{\partial g}{\partial x_{2}}\right]_{\left(x_{1}^{*}, x_{2}^{*}\right)}} d x_{1} \tag{3}
\end{equation*}
$$

Thus the admissible variation $d x_{2}$ depends on dx land $d x_{1}$ can be chosen arlitravity. Using (3) in (1) we have for admissible unviations

$$
\begin{equation*}
\left[\frac{\partial f}{\partial x_{1}}-\frac{\frac{\partial g}{\partial x_{1}}}{\frac{\partial g}{\partial x_{2}}} \cdot \frac{\partial f}{\partial x_{2}}\right]_{\left(x_{1}^{*}, \ldots 2\right)} d x_{1}=0 \tag{12.}
\end{equation*}
$$

Since $d x_{1}$ is albitrary we have

$$
\left[\frac{\partial f}{\partial x_{1}} \frac{\partial g}{\partial x_{2}}-\frac{\partial g}{\partial x_{1}} \cdot \frac{\partial f}{\partial x_{2}}\right]_{\left(x_{1}^{*} \cdot x_{2}^{*}\right)}=0
$$

This is the necessary condition for $\left(x_{1}^{*}, x_{2}^{*}\right)$ to be an extreme point.
Result : The solution of the problem
Minimize $\mathrm{f}\left(x_{1}, x_{2}\right)$
subject to $g\left(x_{1}, x_{2}\right)=0$
is obtained by sloving

$$
\frac{\partial f}{\partial x_{1}} \frac{\partial g}{\partial x_{2}}-\frac{\partial g}{\partial x_{1}} \frac{\partial f}{\partial x_{2}}=0
$$

and $g\left(x_{1}, x_{2}\right)=0$
The above result can be generalized for general problem in the following theorem.
Theorem 1.3.1. Necessary conditions for $\left(x_{1}^{*}, x_{2}^{*}, \ldots, x_{n}^{*}\right)$ to be an extreme point of the function $\mathrm{f}\left(x_{1}, x_{2}, \ldots ., x_{n}\right)$ to exist under the $m$ equality constraints $\mathrm{g}_{\mathrm{j}}\left(x_{1}, x_{2}\right.$. $\left.\ldots . . x_{n}\right)=0, \mathrm{j}=1,2 \ldots . ., \mathrm{m}(m<n)$ are the following $(n-m)$ equations ar satisfied at $\left(x_{1}^{*}, x_{2}^{*}, \ldots \ldots x_{n}^{*}\right)$.
$\mathrm{J}\left(\frac{f, g_{1}, g_{2}, \ldots \ldots, g_{m}}{x_{1}, x_{2}, \ldots \ldots, x_{m}}\right)=\left|\begin{array}{ccccc}\frac{\partial f}{\partial x_{k}} & \frac{\partial f}{\partial x_{1}} & \frac{\partial f}{\partial x_{2}} & \ldots & \frac{\partial f}{\partial x_{m}} \\ \frac{\partial g_{1}}{\partial x_{k}} & \frac{\partial g_{1}}{\partial x_{1}} & \frac{\partial g_{1}}{\partial x_{2}} & \ldots & \frac{\partial g_{1}}{\partial x_{m}} \\ \vdots & \vdots & \vdots & \ldots & \vdots \\ \frac{\partial g_{m}}{\partial x_{k}} & \frac{\partial g_{m}}{\partial x_{1}} & \frac{\partial g_{m}}{\partial x_{2}} & \cdots & \frac{\partial g_{m}}{\partial x_{m}}\end{array}\right|=0$
$\mathrm{k}=\mathrm{m}+1, \mathrm{~m}+2, \ldots \ldots \mathrm{n}$
Where J $\left(\frac{g_{1}, g_{2}, \ldots \ldots g_{m}}{x_{1}, x_{2}, \ldots, x_{m}}\right)=\left|\begin{array}{cccc}\frac{\partial g_{1}}{\partial x_{1}} & \frac{\partial g_{1}}{\partial x_{2}} & \ldots & \frac{\partial g_{1}}{\partial x_{m}} \\ \frac{\partial g_{2}}{\partial x_{1}} & \frac{\partial g_{2}}{\partial x_{2}} & \ldots & \frac{\partial g_{2}}{\partial x_{m}} \\ \cdots & \cdots & \cdots & \cdots \\ \cdots & \cdots & \cdots & \cdots \\ \frac{\partial g_{m}}{\partial x_{1}} & \frac{\partial g_{m}}{\partial x_{2}} & \cdots & \frac{\partial}{g_{m}} \\ x_{m}\end{array}\right| \neq 0$
13

Note : In the above theorem $x_{\mathrm{m}+1}, x_{\mathrm{m}+2}, \ldots . . ., x_{\mathrm{n}}$ are independent variable. Also we note that the dependent variable, $x_{1}, x_{2}, \ldots \ldots . x_{\mathrm{m}}$ must satisfy I $\left(\frac{g_{1}, g_{2}, \ldots \ldots g_{m}}{x_{1}, x_{2}, \ldots, x_{m}}\right) \neq 0$

Example 1.3.1 Using method of constrained variation
Minimize $\mathrm{f}\left(x_{1}, x_{2}, x_{3}\right)=x_{1}^{2}+2 x_{2}^{2}+x_{3}^{2}$
subject to $2 x_{1}+4 x_{2}+3 x_{3}+9$

$$
4 x_{1}+8 x_{2}+5 x_{3}+17
$$

Solution. .
We are b minimize

$$
\begin{equation*}
\mathrm{f}=x_{1}^{2}+2 x_{2}^{2}+x_{3}^{2} \tag{1}
\end{equation*}
$$

subject to $g_{1}=2 x_{1}+4 x_{2}+3 x_{3}-9=0$

$$
\begin{equation*}
g_{2}=4 x_{1}+8 x_{2}+5 x_{3}-17=0 \tag{2}
\end{equation*}
$$

We are first to select independent and dependent variable.
Let us consider

$$
\mathrm{J}\left(\frac{g_{1}, g_{2}}{x_{1}, x_{2}}\right)=\left|\begin{array}{ll}
\frac{\partial g_{1}}{\partial x_{1}} & \frac{\partial g_{1}}{\partial x_{2}} \\
\frac{\partial g_{2}}{\partial x_{1}} & \frac{\partial g_{2}}{\partial x_{2}}
\end{array}\right|=\left|\begin{array}{ll}
2 & 4 \\
4 & 8
\end{array}\right|=0
$$

Thus $x_{3}$ cannot be chosen as independent variables.
Let us now conisder
$\mathrm{J}\left(\frac{g_{1}, g_{2}}{x_{1}, x_{2}}\right)=\left|\begin{array}{ll}\frac{\partial g_{1}}{\partial x_{1}} & \frac{\partial g_{1}}{\partial x_{2}} \\ \frac{\partial g_{2}}{\partial x_{1}} & \frac{\partial g_{2}}{\partial x_{2}}\end{array}\right|=\left|\begin{array}{ll}2 & 3 \\ 4 & 5\end{array}\right|=10-12=-2 \neq 0$
Thus $x_{2}$ cannot be chosen as independent variables.
The necessary condition is

$$
\mathrm{J}\left(\frac{f, g_{1}, g_{2}}{x_{2}, x_{1}, x_{3}}\right)=0
$$

or, $\quad\left|\begin{array}{lll}\frac{\partial f}{\partial x_{2}} & \frac{\partial f}{\partial x_{1}} & \frac{\partial f}{\partial x_{3}} \\ \frac{\partial g_{1}}{\partial x_{2}} & \frac{\partial g_{1}}{\partial x_{1}} & \frac{\partial g_{1}}{\partial x_{3}} \\ \frac{\partial g_{2}}{\partial x_{2}} & \frac{\partial g_{2}}{\partial x_{1}} & \frac{\partial g_{2}}{\partial x_{3}}\end{array}\right|=0$
or, $\quad\left|\begin{array}{ccc}4 x_{2} & 2 x_{1} & 2 x_{3} \\ 4 & 2 & 3 \\ 8 & 4 & 5\end{array}\right|=0$
or, $\quad 4 x_{2}(10-12)+2 x_{1}(24-20)+2 x_{3}(16-16)=0$
or, $-8 x_{2}+8 x_{1}+0=0$
or, $\quad x_{2}=x_{1} \ldots \ldots \ldots . .$. (3)
Using (3) in (1) \& (2) we get respectively

$$
\begin{array}{ll} 
& 6 x_{1}+3 x_{3}-9=0 \\
\text { and } & 12 x_{1}+5 x_{3}-17=0 \\
\therefore \quad & x_{1}=\frac{-15+45}{30-36}=1 \\
& x_{3}=\frac{-108+102}{30-36}=1
\end{array}
$$

From (3) we have $x_{2}=1$
Hence the required solution is $x_{1}=1, x_{2}=1, x_{3}=1$.

### 1.3.2 Method of Lagrange multipliers

In the Lagrange miltiplier method are additional variable is introduced to the problem for each constraint. If the original problem has $n$ variables and $m$ equality constraints then we are to add $m$ additional variables to the problem so that the final number of unknowns becomes $n+m$.

We now state the famous theorems of Lagrange.
Theorem 1.3.2 A necessary condition for a function $\mathrm{f}\left(x_{1}, x_{2}, \ldots ., x_{n}\right)$ subject to the constraints $\mathrm{g}_{\mathrm{j}}\left(x_{1}, x_{2}, \ldots . ., x_{\mathrm{n}}\right)=0, \mathrm{j}=1,2, \ldots \ldots, \mathrm{~m}$ to have a relative minimum at a point $\left(x_{1}^{*}, x_{2}^{*}, \ldots \ldots x_{n}^{*}\right)$ is that the first partial derivatives of the Lagrange function
$\mathrm{L}=\left(x_{1}, x_{2}, \ldots . ., x_{11}, \lambda_{1}, \lambda_{2}, \ldots . . . \lambda_{n}\right)=\mathrm{f}+\sum_{j=1} \lambda_{j} g_{j}$ with respect to each of its arguments must be zero.

The sufficient condition for a function subject to equality constranits is given in the following theorem.

Theorem 1.3.3 A sufficient condition for a function $\mathrm{f}\left(x_{1}, x_{2}, \ldots ., x_{\mathrm{n}}\right)$ subject to the constraints $\operatorname{gj}\left(x, x_{2}, \ldots ., x_{\mathrm{n}}\right)=0, \mathrm{j}=1,2, \ldots ., m$ to have a relative minimum (maximum) at a point $\left(x_{1}^{*}, x_{2}^{*}, \ldots \ldots x_{n}^{*}\right)$ is that the quadratic Q , defined by

$$
\mathrm{Q}=\sum_{i=1}^{n} \sum_{j=1}^{n} \frac{\partial^{2} L}{\partial x_{i} \partial x_{j}} d x_{i} d x_{j} \ldots \ldots \ldots \ldots . . \text { ( ) evalualed at }\left(x_{1}^{*}, x_{2}^{*}, \ldots \ldots x_{n}^{*}\right) \text { must be positive }
$$ (negative) definite for all choice of admissible variations $d x_{\mathrm{i}}$.

Theorem (Hanock) 1.3.4 A necessary condition for the quadratic form $\mathrm{Q}=$ $\sum_{i=1}^{n} \sum_{j=1}^{n} \frac{\partial^{2} L}{\partial x_{i} \partial x_{j}} d x_{i} d x_{j}$, evaluated at ( $x_{1}^{*}, x_{2}^{*}, \ldots \ldots x_{n}^{*}$ ) to be positive (negative) definite for all admissible variations $d x_{\text {}}$ is that each root of the polynomial defined by the following determinantal equation, be positive (negative) :

$$
\begin{aligned}
& \qquad\left|\begin{array}{cccccccccc}
\left(\mathrm{L}_{11}-\mathrm{Z}\right) & \mathrm{L}_{12} & \ldots & \ldots & \mathrm{~L}_{11} & g_{11} & g_{21} & \ldots & \ldots & g_{m 1} \\
\mathrm{~L}_{21} & \left(\mathrm{~L}_{22}-\mathrm{Z}\right) & \ldots & \ldots & \mathrm{L}_{2 \mathrm{n}} & g_{12} & g_{22} & \ldots & \ldots & g_{m 2} \\
\ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots \\
\ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots \\
\mathrm{~L}_{n 1} & \mathrm{~L}_{n 2} & \ldots & \ldots & \left(\mathrm{~L}_{n \mathrm{n}}-\mathrm{Z}\right) & g_{1 m} & g_{2 m} & \ldots & \ldots & g_{m m} \\
g_{11} & g_{12} & \ldots & \ldots & g_{1 m} & 0 & 0 & \ldots & \ldots & 0 \\
g_{21} & g_{22} & \ldots & \ldots & g_{2 m} & 0 & 0 & \ldots & \ldots & 0 \\
\ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots \\
\ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots & \ldots \\
g_{m 1} & g_{m 2} & \ldots & \ldots & g_{n m} & 0 & 0 & \ldots & \ldots & 0
\end{array}\right| \\
& \text { Where } \mathrm{L}_{\mathrm{ij}}=\left[\frac{\partial^{2} L}{\partial x_{j} \partial x_{j}}\right]_{x^{*}} \\
& \text { and } \mathrm{g}_{\mathrm{ij}}=\left[\frac{\partial g_{i}}{\partial x_{j}}\right]_{x^{*}}, \mathrm{X}^{*}=\left(x_{1}^{*}, x_{2}^{*}, \ldots \ldots x_{n}^{*}\right)
\end{aligned}
$$

Result : If some of the roots of the above determinantal equation are positive and some are negative then the point $x^{*}$ is not an extreme point.

Example 1.3.2 : Using Lagrange multiplier method minimize the function
$\mathrm{f}\left(x_{1}, x_{2}, x_{3}\right)=9-8 x_{1}-6 x_{2}-4 x_{3}+2 x_{1}^{2}+2 x_{2}^{2}+x_{3}^{2}+2 x_{1} x_{2}+2 x_{1} x_{3}$ subjecto the constrain $x_{1}+x_{2}+2 x_{3}=3$

Solution. Hope $f=9-8 x_{1}-6 x_{2}+4 x_{3}+2 x_{1}^{2}+2 x_{2}^{2}+x_{3}^{2}+2 x_{1} x_{2}+2 x_{1} x_{3}$

$$
\mathrm{g}=x_{1}-x_{2}+2 x_{3}-3=0
$$

The Lagrange functionis given by
$\mathrm{L}\left(x_{1}, x_{2}, x_{3}, \lambda\right)=\mathrm{f}+\lambda \mathrm{g}$
$=\left(9-8 x_{1}-6 x_{2}+4 x_{3}+2 x_{1}^{2}+2 x_{2}^{2}+x_{3}^{2}+2 x_{1} x_{2}+2 x_{1} x_{3}\right)+\lambda\left(x_{1}, x_{2}, 2 x_{3}-3\right)$
The necessary condition are

$$
\begin{array}{ll}
\frac{\partial L}{\partial x_{1}}=0 & \text { or. }
\end{array}-8+4 x_{1}+2 x_{2}+2 x_{3}+\lambda=0, ~ \begin{array}{ll}
\frac{\partial L}{\partial x_{2}}=0 & \text { or, }-6+4 x_{2}+2 x_{1}+\lambda=0 \\
\frac{\partial L}{\partial x_{3}}=0 & \text { or, }-4+2 x_{3}+2 x_{1}+2 \lambda=0 \\
\frac{\partial L}{\partial \lambda}=0 & \text { or, } x_{1}+x_{2}+2 x_{3}-3=0
\end{array}
$$

Solving these four equations we have

$$
x_{1}^{*}=4 / 3, x_{2}^{*}=7 / 9, x_{3}^{*}=4 / 9 \text { and } \lambda^{*}=2 / 9
$$

We now use sufficient condition to identify this extreme point.
We evaluate $\mathrm{L}_{\mathrm{ij}}$ and $\mathrm{g}_{\mathrm{ij}}$ at the point $(4 / 3,7 / 9,4 / 9)=\mathrm{X}^{*}$

$$
\begin{align*}
& \mathrm{L}_{11}=\left[\frac{\partial^{2} \mathrm{~L}}{\partial x_{1}^{2}}\right]_{x^{*}}=4 \\
& \mathrm{~L}_{12}=\mathrm{L}_{21}=\left[\frac{\partial^{2} \mathrm{~L}}{\partial x_{1} x_{2}}\right]_{x^{*}}=2 \tag{17}
\end{align*}
$$

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$$
\begin{aligned}
& \mathrm{L}_{13}=\mathrm{L}_{31}=\left[\frac{\partial^{2} \mathrm{~L}}{\partial x_{1} x_{3}}\right]_{\mathrm{x}^{*}}=2 \\
& \mathrm{~L}_{22}=\left[\frac{\partial^{2} \mathrm{~L}}{\partial x_{2}^{2}}\right]_{x^{*}}=4 \\
& \mathrm{~L}_{23}=\mathrm{L}_{32}=\left[\frac{\partial^{2} \mathrm{~L}}{\partial x_{2} x_{3}}\right]_{\mathrm{x}^{*}}=0 \\
& \mathrm{~L}_{33}=\left[\frac{\partial^{2} \mathrm{~L}}{\partial x_{3}^{2}}\right]_{\mathrm{x}^{*}}=2 \\
& \mathrm{~g}_{11}=\left[\frac{\partial \mathrm{g}}{\partial x_{1}}\right]_{\mathrm{x}^{*}}=1 \\
& \mathrm{~g}_{12}=\left[\frac{\partial \mathrm{g}}{\partial x_{2}}\right]_{\mathrm{x}^{*}}=1 \\
& \mathrm{~g}_{13}=\left[\frac{\partial \mathrm{g}}{\partial x_{3}}\right]_{\mathrm{x}^{*}}=2
\end{aligned}
$$

We now consider the determinautor equation

$$
\begin{aligned}
& \quad\left|\begin{array}{cccc}
\mathrm{L}_{11}-z & \mathrm{~L}_{12} & \mathrm{~L}_{13} & g_{11} \\
\mathrm{~L}_{21} & \mathrm{~L}_{22}-z & \mathrm{~L}_{23} & g_{12} \\
\mathrm{~L}_{31} & \mathrm{~L}_{32} & \mathrm{~L}_{33}-z & g_{13} \\
g_{11} & g_{12} & g_{13} & 0
\end{array}\right|=0 \\
& \text { or, } \quad\left|\begin{array}{cccc}
4-z & 2 & 2 & 1 \\
2 & 4-z & 0 & 1 \\
2 & 0 & 2-z & 2 \\
1 & 1 & 2 & 0
\end{array}\right|=0 \\
& \text { or, } \quad-1\left|\begin{array}{ccc}
2 & 2 & 1 \\
4-z & 0 & 1 \\
0 & 2-z & 2
\end{array}\right|+1\left|\begin{array}{ccc}
4-z & 2 & 2 \\
2 & 0 & 2-z \\
1 & 1 & 2
\end{array}\right|-2\left|\begin{array}{ccc}
4-z & 2 & 2 \\
1 & 4-z & 0 \\
1
\end{array}\right|=0
\end{aligned}
$$

or, $\quad z^{2}-6 z+9=0$
or, $\quad z=3,3$
Since the roots are all positive, ( $4 / 3,7 / 9,4 / 9$ ) is a relative. minimum of the function.

### 1.4 Multivariable optimization with inequality constraints

The general multivariable optimization problem with inequality constraints is
Minimize $f=f(X)$
subject to $\mathrm{g}_{\mathrm{i}}(\mathrm{x}) \leq \mathrm{b}_{\mathrm{j}, \mathrm{j}} \mathrm{j}=1,2, \ldots, \mathrm{~m}$
where $\mathrm{X}=\left[x_{1}, x_{2}, \ldots . . . . . . x_{\mathrm{n}}\right]^{\top}$.
This section is concerned with developing the necessary and sufficient conditions for identifying the stationery points of the above problem. These conditions are called Kuhn-Tucker conditions and the development is mainly based on Lagrangian method.

Theorem 1.4.1 (Kuhn-Tucker Necessary Conditions)
Given the problem to minimize

$$
\mathrm{f}=\mathrm{f}(\mathrm{x})=\mathrm{f}\left(x_{1}, x_{2}, \ldots \ldots, x_{\mathrm{n}}\right)
$$

subject to $g(\dot{X})=g_{\mathrm{j}}\left(x_{1}, x_{2}, \ldots \ldots, x_{\mathrm{n}}\right) \leq \mathrm{b}_{\mathrm{j}} \quad \mathrm{i}=1,2, \ldots \ldots ., \mathrm{m}$ the necessary conditions for $\mathrm{X}_{0}$ to be a local minimum are that
(i) $\frac{\partial f}{\partial x_{i}}+\sum_{j=1}^{m} \lambda_{j} \frac{\partial g_{j}}{\partial x_{i}}=0, \quad i=1,2, \ldots \ldots . n$
(ii) $\lambda_{\mathrm{j}}\left[\mathrm{g}_{\mathrm{j}}(\mathrm{X})-\mathrm{b}_{\mathrm{j}}\right]=0, \mathrm{j}=1,2, \ldots \ldots . \mathrm{m}$
(iii) $\mathrm{g}_{\mathrm{j}}(\mathrm{X}) \leq \mathrm{b}_{\mathrm{j}}, \mathrm{j}=1,2, \ldots . . \mathrm{m}$
(iv) $\lambda_{j} \geq 0, j=1,2, \ldots \ldots, m$
are satisfied at $\mathrm{X}_{0}$.
Introducing slack variables the inequality constraints bcomes

$$
\begin{equation*}
\mathrm{g}_{\mathrm{j}}(\mathrm{X})+s_{j}^{2}=\mathrm{b}_{\mathrm{j}}, \quad \mathrm{j}=1,2, \ldots \ldots, \mathrm{~m} \tag{1}
\end{equation*}
$$

or, $\quad g_{j}(X)+s_{j}^{2}-b_{j}=0, \quad j=1,2, \ldots \ldots .$.

In order to obtain all stationary points, we form the Lograngian function L given by
$\mathrm{L}(\mathrm{X}, \mathrm{I}, \mathrm{S})=\mathrm{f}(\mathrm{X})+\sum_{j=1}^{m} \lambda_{j}\left\{\mathrm{~g}^{j}(\mathrm{X})+s_{j}^{2}-\mathrm{b}^{\mathrm{i}}\right\}$
Then the stationary points are obtained by polving the equations

$$
\begin{aligned}
& \frac{\partial L}{\partial x_{i}}=0, i=1,2, \ldots \ldots, n \\
& \frac{\partial L}{\partial \lambda_{j}}=0, j=1,2, \ldots \ldots, m
\end{aligned}
$$

and $\quad \frac{\partial L}{\partial s_{j}}=0, j=1,2 . \ldots \ldots, m$
i.e., $\quad \frac{\partial f}{\partial x_{i}}+\sum_{j=1}^{m} \lambda, \frac{\partial g_{j}}{\partial x_{i}}=0, \quad \mathrm{i}=1,2, \ldots . . \mathrm{n}$

$$
\begin{align*}
& g_{j}+s_{j}^{2}-b_{j}=0  \tag{3}\\
& 2 \lambda_{j} s_{j}=0,
\end{align*}
$$

Multiplying (4) by $\frac{1}{2} \mathrm{~s}$ : we get,

$$
\lambda_{j} s_{j}^{2}=0
$$

Using (1) this gives

$$
\begin{align*}
\lambda_{j}\left\{b_{j}-g_{j}(X)\right\} & =0 \\
\text { or, } \lambda_{j}\left\{g_{j}(X)-b_{j}\right\} & =0, \quad j=1,2, \ldots \ldots . . m \tag{5}
\end{align*}
$$

From (5) we have when $\lambda_{j} \neq 0$ then $g_{j}(X)-b_{j}=0$ or, $g_{j}(X)=b_{j}$

$$
\text { or, } \frac{\partial g_{j}}{\partial b_{j}}=1
$$

Thus $\frac{\partial g_{k}}{\partial b_{j}}=s_{j k}$ where $s_{j k}=\left\{\begin{array}{c}1 \text { for } j=k \\ 0 \text { for } j \neq k\end{array}\right.$

Using chain rule of differential calculus we have

$$
s_{j k}=\frac{\partial g_{k}}{\partial b_{j}}=\sum_{i=1}^{n} \frac{\partial g_{k}}{\partial x_{i}} \frac{\partial x_{i}}{\partial b_{j}}
$$

Multiplying both sides by $\lambda k$ and summing over all values of $k$ we get

$$
\sum_{k=1}^{m} \lambda_{k} s_{j k}=\sum_{k=1}^{m} \lambda_{k}\left(\sum_{i=1}^{n} \frac{\partial g_{k}}{\partial x_{i}} \frac{\partial x_{i}}{\partial b_{j}}\right)
$$

or, $\quad \lambda_{j}=\sum_{k=1}^{m} \lambda_{0}\left(\sum_{i=1}^{n} \frac{\partial g_{k}}{\partial x_{j}} \frac{\partial x_{i}^{i}}{\partial b_{j}}\right)$
Again $\frac{\partial f}{\partial b_{j}}=\sum_{i=1}^{n} \frac{\partial f}{\partial x_{i}} \frac{\partial x_{i}}{\partial b_{j}}$
Adding (6) and (7) we get

$$
\begin{aligned}
& \frac{\partial f}{\partial b_{j}}+\lambda_{j}=\sum_{i=1}^{n}\left[\frac{\partial f}{\partial x_{i}}+\sum_{k=1}^{m} \lambda_{k} \frac{\partial g_{k}}{\partial x_{i}}\right] \frac{\partial x_{i}}{\partial b_{j}} \\
& \quad=0 \text { [using (2)] }
\end{aligned}
$$

or, $\frac{\partial f}{\partial b_{j}}=-\lambda_{j}$
Thus when $\lambda_{\mathrm{j}} \neq 0$ then we have $\lambda_{\mathrm{j}}=-\frac{\partial f}{\partial b_{j}}$
We now show that $\lambda_{\mathrm{j}}>0$. If possible let $\lambda_{\mathrm{j}}<0$. Then from (9) we have $\frac{\partial f}{\partial b_{j}}>0$
This implies that as $b_{j}$ is increased, the objective function increases. Now as $b_{j}$ optimal value of the objective function clearly cannot increase. This contradicts our assumption $\lambda_{j}>0$. Thus at an optimal solution we have $\lambda_{j}>0$ when $\lambda_{j} \neq$ 0 . Hence at the optimal solution we hour $\lambda j \geq 0$.

Note : For the problem
Maximize $\mathrm{f}=\mathrm{f}\left(x_{1}, x_{2}, \ldots \ldots, x_{\mathrm{n}}\right)$
subject to $g_{j}\left(x_{1}, x_{2}, \ldots . ., x_{\mathrm{n}}\right) \mathrm{b}_{\mathrm{j}}, \mathrm{i}=1,2, \ldots . . m$
the Kuhn-Tucker necessary conditions for $\left(x_{1}^{*}, x_{2}^{*}, \ldots \ldots x_{n}^{*}\right)$ to be a local maximum are that
(i) $\frac{\partial f}{d x_{i}}+\sum_{j=1}^{m} \lambda_{j} \frac{\partial g_{i}}{\partial x_{i}}=0, i=1,2, \ldots \ldots, n$
(ii) $\lambda_{j}\left[g_{j}-b_{j}\right]=0, j=1,2, \ldots \ldots, m$
(iii) $g_{j} \leq b_{j}, j=1,2, \ldots \ldots, m$
(iv) $\lambda_{j} 0, \mathrm{j}=1,2, \ldots \ldots, m$
are satified at $\left(x_{1}^{*}, x_{2}^{*}, \ldots . x_{n}^{*}\right)$
Sufficency of the Kuhn-Tucker conditions
The Kuhn-Tucker necessary conditions are also sufficient if the objective function and the solution space satisfy certain conditions regarding convexity anc concairty. For maximization problem the objective function should be concave and solution space should be convex set.

For minimization problem the objective function should be convex and the solution space should be convex set.

Example 1.4.1. Solve using Kuhn-Tucker conditions
Maximize $2=5+8 x_{1}+12 x_{2}-4 x_{1}^{2}-4 x_{2}^{2}-4 x_{3}^{2}$
subject to $x_{1}+x_{2} \leq 1$

$$
2 x_{1}+3 x_{2} \leq 6
$$

Here the constraints are

$$
\mathrm{g}_{1}=x_{1}+x_{2} \leq 1
$$

and $\quad g_{2}=2 x_{1}+3 x_{2} \leq 6$
The Kuhn-Tucker necessary conditions are

$$
\begin{align*}
& \\
& \frac{\partial Z}{\partial x_{i}}+\lambda_{1} \frac{\partial g_{1}}{\partial x_{i}}+\lambda_{2} \frac{\partial g_{2}}{\partial x_{i}}=0, i=1,2,3 \\
& \\
& \lambda_{\mathrm{j}}\left[g_{\mathrm{j}}-\mathrm{b}_{\mathrm{j}}\right]=0, \quad \mathrm{j}=1,2  \tag{1}\\
& \\
& \lambda_{\mathrm{j}} \leq 0, \quad \mathrm{j}=1,2 \\
& \text { i,.e. } \\
& 8-8 x_{1}+\lambda_{1}+2 \lambda_{2}=0
\end{align*}
$$

$$
\begin{align*}
& 12-8 x_{2}+\lambda_{1}+3 \lambda_{2}=0  \tag{2}\\
& -8 x_{1}=0  \tag{3}\\
& \lambda_{1}+\left(x_{1}+x_{2}-1\right)=0  \tag{4}\\
& \lambda_{2}+\left(2 x_{1}+3 x_{2}-6\right)=0  \tag{5}\\
& x_{1}+x_{2}-1=0  \tag{6}\\
& 2 x_{1}+3 x_{2}-6 \leq 0  \tag{7}\\
& \lambda_{1} \leq 0  \tag{8}\\
& \lambda_{2} \leq 0 \tag{9}
\end{align*}
$$

Four cases may arise.
case 1. $\lambda_{1}=0, \lambda_{2}=0$
case 2. $\quad \lambda_{1}=0, \lambda_{2} \neq 0$
case 3. $\lambda_{1} \neq 0, \lambda_{2}=0$
case 4. $\quad \lambda_{1} \neq 0, \lambda_{2} \neq 0$
Case 1. Here $\lambda_{1}=0, \lambda_{2}=0$
From (1) we get $x_{1}=1$
From (2) we get $x_{2}=3 / 2$
This solution does not satisfy (6). So this solution is discarded
Case 1. Here $\lambda_{1}=0, \lambda_{2} \neq 0$
From (5) we get $2 x_{1}+3 x_{1}-6=0$
(1) becomes $8-8 x_{1}+2 \lambda_{2}=0 \quad$ or, $x_{1}=\left(\lambda_{2}+4\right) / 4$
(2) becomes $12-8 x_{2}+3 \lambda_{2}=0$ or, $x_{2}=\left(3 \lambda_{2}+12\right) / 8$

Using (11) and (12) we get from.(10)

$$
\left(\lambda_{2}+4\right) / 2+\left(g \lambda_{2}+36\right) / \mathrm{S}-6=0
$$

or. $\quad 4 \lambda_{2}+16+g \lambda_{2}+36-48=0$
or, $\quad 13 \lambda_{2}=-4$
or, $\quad 12=-4 / 13<0$

From (11) we have $x_{1}=-\frac{1}{13}+1=\frac{12}{13}$
From (11) we have $x_{1}=-\frac{24}{104}+\frac{12}{18}=\frac{18}{13}$
This solution violets (6) and so is discorded.
Case. 3 Here $\lambda_{1} \neq 0$ and $\lambda_{2}=0$
From (4) we have $x_{1}+x_{2}-1=0$
(1) becomes $8-8 x_{1}+\lambda_{1}=0$ or, $x_{1}=\left(\lambda_{1}+8\right) / 8$
(2) becomes $12-8 x_{2}+\lambda_{1}=0$ or, $x_{2}=\left(\lambda_{1}+12\right) / 8$

Using (14). (15) in (13) we have
or, $\quad \lambda_{1}=-6$
From (14) and (15) we get $x_{1}=1 / 4, x_{2}=3 / 4$
From (3) we get $x_{3}=0$
$\therefore \quad x_{1}=1 / 4, x_{2}=3 / 4, x_{3}=0$
This solution satifies (6) and (7).
Hence this is the optimum solution.

### 1.5 Summary

This unit is devoted with the classical theory of optimization for locating the points of maxima and minima of constrained and unconstrained nonlinear problems. This theory deals with the use of diffrential calculus. The topics introduced includs the development of the necessary and sufficient conditions for locating the extreme points for unconstrained problems, the treatment of the constrained problem with equality constraints using Lagrangian methods, and the development of the KuhnTucker conditions for the general problem with inequality constraints. Though the classical optimization techniques are not suitable for obtaining real life problems, the underlying theory gives the basis for devising most of the non-linear programming algorithms.

### 1.6 Assesment Questions

1. Determine the extreme points of the function

$$
f=8 x_{1}^{3}+27 x_{2}^{3}+16 x_{1}^{2}+18 x_{2}^{2}+6
$$

2. Determine the extreme points of the function

$$
\mathrm{Z}=121+27 x_{1}^{3}+64 x_{2}^{3}+36 x_{1}^{2}+32 x_{2}^{2}
$$

3. Find the extreme points of the function

$$
\mathrm{f}=x_{1}^{3}+x_{2}^{3}+2 x_{1}^{2}+4 x_{2}^{2}+20
$$

4. The total profits $(z)$ of a firm depend upon the level of of output $(Q)$ and the advertising expenditure (A). Find the profit maximizing values of $Q$ (in thousand units) and A (Rs in thousand) given the following relationship.

$$
Z=800-3 Q^{2}-4 Q+2 Q A-5 A^{2}+48 A
$$

5. Using method of constrained variation and method of Lagrange multiplier
(i) Minimize $\mathrm{f}(x)=\frac{1}{2}\left(x_{1}^{2}+x_{2}^{2}+x_{3}^{2}\right)$

Subject to $x_{1}=x_{2}$

$$
x_{1}+x_{2}+x_{3}=1
$$

(ii) Minimize $\mathrm{f}=19-16 x_{1}+6 x_{2}-4 x_{3}+8 x_{1}^{2}+2 x_{2}^{2}+x_{3}^{2}-4 x_{1} x_{2}+4 x_{1} x_{3}$
subject to $2 x_{1}-x_{2}+2 x_{3}=3$
(iii) Maximize $\mathrm{f}=8 x_{1} x_{2} x_{3}$
subject to $x_{1}^{2}+x_{2}^{2}+x_{3}^{2}=1$
(iv) Minimize $\mathrm{f}=4 x_{1}^{2}+2 x_{2}^{2}+9 x_{3}^{2}$
subject to $4 x_{1}-4 x_{2}+9 x_{3}=9$

$$
8 x_{1}-8 x_{2}+15 x_{3}=17
$$

6. Using Kunh-Tucker condition determine the variable values to

$$
\text { Maximize } z=x_{1}^{2}-x_{2}^{2}-x_{3}^{2}+4 x_{1}+6 x_{2}
$$

subject to $x^{1}+x_{2} \leq 2$

$$
2 x_{1}+3 x_{2} \leq 12
$$

7. Use Kuhn-Tucker conditions of solve the following non-linear programming problems
(i) Maximize $\mathrm{Z}=x_{1}^{2}+6 x_{1}+5 x_{2}$
subject to $x_{1}+2 x_{2} \leq 10$

$$
x_{1}+3 x_{2} \leq 9
$$

(ii) Maximize $\mathrm{Z}=2 x_{1}-x_{1}^{2}+x_{2}$

$$
\text { subject to } 2 x_{1}+3 x_{2} \leq 6
$$

$$
2 x_{1}+x_{2} \leq 4
$$

(iii) Maximize $\mathrm{Z}=2 x_{1}^{2}+12 x_{1} x_{2}-7 x_{2}^{2}$

$$
\text { subject to } 2 x_{1}+5 x_{2} \leq 98
$$

$$
x_{1}+x_{2} \geq 0
$$

(iv) Maximize $Z=8 x_{1}+10 x_{2}-x_{1}^{2}-x_{2}^{2}$

$$
\text { subject to } 3 x_{1}+2 x_{2} \leq 6
$$

$$
x_{1}, x_{2} \geq 0
$$

## Unit $2 \square$ Revised Simplex Method

## Structure

### 2.1 Introduction

## 2:2 Revised Simplex Method

2.3 Standard Form for Revised Simplex Method
2.4 A Logarithm of Revised Simplex Method
2.5 Comparison of Simplex Method and Revised Simplex Method
2.6 Illustrative Examples
2.7 Summary
2.8 Self Assessment Questions

### 2.1. Introduction

The revised simplex method proceeds through the same steps as simplex method but keeps all important data in a smaller array. The 'revised' aspect concerns the procedure of changing the simplex tables only. The revised simplex method is thus an efficient computational procedure for solving a linear programming problem with less time and labour. For large size problem this method is found to be ..... useful as it reduces the cost of obtaining the solution.

### 2.2 Revised Simplex Method :

When a linear programming problem is solved simplex method, successive iterations are obtained by using suitable row operations so that the objective function reduces its value in each step if it is a problem of maximization. Also the net evaluations should remain always non-negative in every step. This method requires storing the entire table, in the memory of the computer. For large size problem it may not be feasible. So, it
requires to device a new method by modifying simplex method to handle LPP will large number of decision variables and constraints.

In fact, it is found tat it is not necessary to compute the entire simplex table during each iteration. The only informations needed to pass from one table to the next one are seen to be
(i) Net evaluations $z_{j}-e_{j}$ to determine the non-basic variable that enters the basis.
(ii) The key column.
(iii) The current basic variables and their values to determine the minimum positive ratio, and thereby to determine the basic variable that will leave the basis.
It is sown that all the above informations can be directly obtained from the original equations of the given LPP by making use of the inverse of the current basis matrix.

If $B$ be the current basis then we have

$$
\begin{aligned}
& x_{\mathrm{B}}=\mathrm{B}^{-1} b, y_{j}=\mathrm{B}^{-1} a_{j} \text { for all } j=1,2, \ldots \ldots, n \\
& z_{j}-c_{j}=C_{\mathrm{B}} \mathrm{~B}^{-1} a_{j}-c_{j} \text { for all } j=1,2, \ldots \ldots, n \\
& \text { and } \mathrm{z}=c_{\mathrm{B}} x_{\mathrm{B}} .
\end{aligned}
$$

We note that all these necessary informations can be calculated if the current value of $\mathrm{B}^{-1}$ is known. Much computational work is needed for transformation of all $\mathrm{y}_{\mathrm{j}}, j=$ 1, 2, $\qquad$
But all $y$; are not needed to go to next table. As noted above we need only to know the key column i.e. $y_{k}$. This will actively save our much labour. At each iteration $x_{B}$, $\mathrm{zc}_{\mathrm{B}} \mathrm{B}^{-1}$ and $\mathrm{B}^{-1}$ are transformed and not all the $y_{j}$ are transformed, only the key column $y_{k}$ is transformed in the revised simplex method. The criteria for selecting the entering and departing vectors in the revised simplex method precisely the same as that was in the simplex method. The labour saving point in this method lies in the fact of computing the inverse of the next basis directly from that of the current basis without actually having to invert the next basis.

### 2.3 Standard Form for Revised Simplex Method :

Let the linear programming problem be
Maximize $z=c x$,
subject to $\quad \mathrm{A} x=b$

$$
\begin{equation*}
x \geq 0 \tag{1}
\end{equation*}
$$

where $c, x^{\mathrm{T}} \in R^{\mathrm{n}}, b \mathrm{~T} \in \mathrm{R} m$ and A is an $m \times n$ real matrix. In the revised simplex method we consider the objective function equation $z=c x$ also one constraint. Thus the new system becomes a $(m+1)$ simultaneous lines equations in $n+1$ variables $z$, $x_{1}, x_{2}, \ldots \ldots, x_{n}$. The problem thus becomes to get the solution of this system such that $z$ is as large as possible. The simultaneous linear system thus becomes

$$
\begin{align*}
\mathrm{A} x+o z & =b \\
-c x+z & =0  \tag{2}\\
\quad x & \geq 0, z \text { is unrestricted. }
\end{align*}
$$

Hence the LPP (1) becomes equivalent to the problem of finding the solution of the system (2) such that $z$ is as large as possible.

In matrix notation (2) becomes

$$
\left[\begin{array}{cc}
\mathrm{A} & 0  \tag{3}\\
-c & 1
\end{array}\right]\left[\begin{array}{l}
x \\
z
\end{array}\right]=\left[\begin{array}{l}
b \\
0
\end{array}\right], x \geq 0
$$

Let B be the initial basis submatrix of A and $x_{\mathrm{B}}=\mathrm{B}^{-1} b$ be the initial basic feasible solution to the original LPP (1).

Since the values of the non-basic variables are always zero (2) becomes

$$
\begin{align*}
& \mathrm{B} x_{\mathrm{B}}+0 z=b \\
& -\mathrm{C}_{\mathrm{B}} x_{\mathrm{B}}+z=0 \tag{4}
\end{align*}
$$

$$
\begin{align*}
& \text { or, }\left[\begin{array}{cc}
\mathrm{B} & 0 \\
-\mathrm{C}_{\mathrm{B}} & 1
\end{array}\right]\left[\begin{array}{c}
x_{\mathrm{B}} \\
z
\end{array}\right]=\left[\begin{array}{l}
b \\
0
\end{array}\right] \\
& \text { or, } \hat{\mathbf{B}} \hat{x}_{\mathrm{B}}=\hat{b} \tag{5}
\end{align*}
$$

where, $\hat{\mathrm{B}}=\left[\begin{array}{cc}\mathrm{B} & 0 \\ -\mathrm{C}_{\mathrm{B}} & 1\end{array}\right], \hat{x}_{\mathrm{B}}=\left[\begin{array}{c}x_{\mathrm{B}} \\ z\end{array}\right]$ and $\hat{b}=\left[\begin{array}{l}b \\ 0\end{array}\right]$
From (4) we have

$$
\begin{equation*}
\hat{x}_{\mathrm{B}}=\hat{\mathrm{B}}^{-1} \hat{b} \tag{7}
\end{equation*}
$$

This is the initial basic feasible solution to the reformulated problem (2).
Computation of Inverse of $\hat{\mathbf{B}}$ by partitioning we have $\mathrm{B}=\left[\begin{array}{cc}B & 0 \\ -\mathrm{C}_{\mathrm{B}} & 1\end{array}\right]$.
Let $\hat{B}^{-1}=\left[\begin{array}{ll}\mathrm{P} & \mathrm{Q} \\ \mathrm{R} & \mathrm{S}\end{array}\right]$
Since $\hat{\mathbf{B}} \hat{\mathrm{B}}^{-1}=\mathrm{I}$, we have

$$
\begin{aligned}
& {\left[\begin{array}{cc}
\mathrm{B} & 0 \\
-\mathrm{C}_{\mathrm{B}} & 1
\end{array}\right]\left[\begin{array}{ll}
\mathrm{P} & \mathrm{Q} \\
\mathrm{R} & \mathrm{~S}
\end{array}\right]=\mathrm{I}_{m+1}} \\
& \text { or, }^{2}\left[\begin{array}{cc}
\mathrm{BP}+\mathrm{OR} & \mathrm{BQ}+\mathrm{OS} \\
-\mathrm{C}_{\mathrm{B}} \mathrm{P}+\mathrm{R} & -\mathrm{C}_{\mathrm{B}} \mathrm{QS}
\end{array}\right]=\left[\begin{array}{cc}
\mathrm{I}_{m} & 0 \\
0 & 1
\end{array}\right] \\
& \therefore \mathrm{BP}=\mathrm{I}_{m} \\
& \quad \mathrm{BQ}=\mathrm{O} \\
& \quad-\mathrm{C}_{\mathrm{B}} \mathrm{P}+\mathrm{R}=0 \\
& -\mathrm{C}_{\mathrm{B}} \mathrm{Q}+\mathrm{S}=1
\end{aligned}
$$

Since $\mathrm{B}^{-1}$ exists, we get from above

$$
\begin{aligned}
& \mathrm{P}=\mathrm{B}^{-1} I_{m}=\mathrm{B}^{-1} \\
& \mathrm{Q}=\mathrm{B}^{-1} \mathrm{O}=\mathrm{O} \\
& \mathrm{R}=\mathrm{C}_{\mathrm{B}} \mathrm{~B}^{-1} \\
& \mathrm{~S}=1+\mathrm{C}_{\mathrm{B}} \mathrm{O}=1
\end{aligned}
$$

Thus from (8) we get

$$
\mathrm{B}^{-1}=\left[\begin{array}{cc}
\mathrm{B}^{-1} & 0  \tag{9}\\
\mathrm{C}_{\mathrm{B}} \mathrm{~B}^{-1} & 1
\end{array}\right]
$$

We note that all the components of $\hat{\mathrm{B}}^{-1}$ are known.
Determination of net evaluations, key column and BFS :
We define $A=\left[\begin{array}{c}A \\ -C\end{array}\right]$

$$
\text { and } \hat{y}=\hat{B}^{-1} \hat{A}
$$

Then $\hat{y}=\left[\begin{array}{cc}B & 0 \\ C_{B} B^{-1} & 1\end{array}\right]\left[\begin{array}{c}A \\ -C\end{array}\right]$

$$
\begin{align*}
& =\left[\begin{array}{cc}
B^{-1} A & -O C \\
C_{B} B^{-1} A & -C
\end{array}\right] \\
& =\left[\begin{array}{ll}
B_{B}^{-1} A \\
C_{B}\left(B^{-1} A\right)-C
\end{array}\right] \tag{10}
\end{align*}
$$

we have

$$
A=B y
$$

$\therefore \mathrm{y}=\mathrm{B}^{-1} \mathrm{~A}$
$\therefore$ From (10) we have $\hat{y}=\left[\begin{array}{l}y \\ C_{B} y-C\end{array}\right]$
or, $\left[\hat{y}_{1} \hat{y}_{2} \ldots \ldots \hat{y}_{n}\right]=\left[\begin{array}{cccc}y_{1} & y_{2} & \cdots \cdots \cdots & y_{n} \\ z_{1}-c_{1} & z_{2}-c_{2} & \cdots \cdots \cdots & z_{n}-c_{n}\end{array}\right]$
Thus for $j=1,2, \ldots . . ., n$ we have

$$
\hat{y}_{j}=\left[\begin{array}{c}
y_{j} \\
z_{j}-c_{j}
\end{array}\right] \text { and } y_{j}=B^{-1} a_{j}
$$

Hence the net evaluation are the components of $\mathrm{C}_{\mathrm{B}} \mathrm{B}^{-1} \mathrm{~A}-\mathrm{C}$
i.e. $\mathrm{C}_{\mathrm{B}} \mathrm{B}^{-1} \mathrm{~A}-\mathrm{C}=\left[\begin{array}{ll}z_{1}-c_{1} & z_{2}-c_{2} \ldots \ldots . . \\ z_{n} & -\mathrm{C}_{n}\end{array}\right]$

Most negative $z_{j}-c_{j}$ will determine the key column. Let $z_{k}-c_{k}$ be the most negative $z_{j}-c_{j}$. Then the key column is

$$
\hat{y}_{k}=\left[\begin{array}{c}
y_{k}  \tag{11}\\
z_{k}-c_{k}
\end{array}\right]=\left[\begin{array}{c}
\mathrm{B}^{-1} a_{k} \\
z_{k}-c_{k}
\end{array}\right]
$$

From (7) and (6) we have

$$
\hat{x}_{\mathrm{B}}=\left[\begin{array}{c}
x_{\mathrm{B}} \\
z
\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}=\left[\begin{array}{cc}
\mathrm{B}^{-1} & 0 \\
\mathrm{C}_{\mathrm{B}} \mathrm{~B}^{-1} & 1
\end{array}\right]\left[\begin{array}{c}
b \\
0
\end{array}\right]=\left[\begin{array}{c}
\mathrm{B}^{-1} b \\
\mathrm{C}_{\mathrm{B}} \mathrm{~B}^{-1} b
\end{array}\right]=\left[\begin{array}{c}
\mathrm{B}^{-1} b \\
\mathrm{C}_{\mathrm{B}} x_{\mathrm{B}}
\end{array}\right]
$$

we note the important fact that all necessary informations can be obtained from the products $\hat{\mathrm{B}}^{-1} \hat{\mathrm{~A}}$ and $\hat{\mathrm{B}}^{-1} \hat{b}$.

Also we note that $\hat{A}$ and $\hat{b}$ remains same in all steps, only $\hat{\mathrm{B}}^{-1}$ changes in each step of simplex table depending on the current basis B.

The above discussion enables us now to state the algorithm of revised simplex method.

### 2.4 ALogarithin of Revised Simplex Method :

It stepwise procedure of revised simplex method are as follows.
Step 1. Introduce necessary slack and surplus variables. Convert the problem into a problem of maximization if it is in minimization form. Restate the LPP in the standard form of revised simplex method i.e. in the form $\left[\begin{array}{cc}A & 0 \\ -c & 1\end{array}\right]\left[\begin{array}{l}x \\ z\end{array}\right]=\left[\begin{array}{l}b \\ 0\end{array}\right], x \geq 0, z$ is unrestricted.

Step 2. Begin with the initial basis $\mathrm{B}=\mathrm{I}_{m}$ and form the auxiliary matrix $\hat{\mathrm{B}}=$ $\left[\begin{array}{cc}\mathrm{B} & 0 \\ -\mathrm{C}_{\mathrm{B}} & 1\end{array}\right]$ and write down

$$
\mathrm{B}^{-1}=\left[\begin{array}{cc}
\mathrm{B}^{-1} & 0 \\
\mathrm{C}_{\mathrm{B}} \mathrm{~B}^{-1} & 1
\end{array}\right] \text {. Form } \hat{\mathrm{A}}=\left[\begin{array}{c}
\mathrm{A} \\
-c
\end{array}\right] \text { and } \hat{b}=\left[\begin{array}{l}
b \\
0
\end{array}\right] \text {, }
$$

Also form $\hat{x}_{\mathrm{B}}=\left[\begin{array}{c}x_{\mathrm{B}} \\ z\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}$.

Step 3. Compute the net evaluations $z_{1}-c_{1}, z_{2}-c_{2}, \ldots . . z_{n}-c_{n}$ as the components of the product

$$
\left[\begin{array}{ll}
\mathrm{C}_{\mathrm{B}} \mathrm{~B}^{-1} & 1
\end{array}\right]\left[\begin{array}{c}
\mathrm{A} \\
-c
\end{array}\right]
$$

If all $z_{j}-c_{j}$ are non-negative, the current basic solution $\hat{x}_{\mathrm{B}}=\left[\begin{array}{c}x_{\mathrm{B}} \\ z\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}$ gives the optimal BFS and maximum value of the objective function.

If at least one $z_{j}-c_{j}$ is negative, determine the most negative of them. If $z_{k}-c_{k}$ is the most negative $z_{j}-c_{j}$ then find $\hat{y}_{k}=\left[\begin{array}{c}y_{k} \\ z_{k}-c_{k}\end{array}\right]=\hat{B}^{-1} \hat{a}_{k}$. Go to step 4. If there is a tie for the most negative $z_{j}-c_{j}$, resolve the tie by any standard method.

Take $x_{k}$ as the new basic variable. Go to step 4.
Step 4. If all $y_{i k} \leq 0$ there exists an unbounded solution to the given problem.
If at least one $y_{i k}>0$, consider the current $\mathrm{x}_{\mathrm{B}}$ and compute the replacement ratios.

$$
\left\{\frac{x_{\mathrm{B} i}}{y_{i k}}: y_{i k}>0\right\}
$$

If $\frac{x_{\mathrm{B} r}}{y_{r k}}$ is the minimum of all these ratios then the basic variable $x_{\mathrm{B} r}$ becomes nonbasic variable in the next table. ie. $x_{\mathrm{B}}$ is replaced by $x_{k}$. Go to step 5 .

Step 5. Write down the results obtained in steps 2, 3 and 4 in a able. This table is known as revised simplex table. This table is of the form


Step 6. Convert the key element $y_{r k}$ of $\hat{y}_{k}$ into unity and all other elements into zero by suitable row operations. Same operations are to be applied in the current $\hat{B}^{-1}$. These operation will change $\hat{\mathrm{B}}^{-1}$ to new $\hat{\mathrm{B}}^{-1}$ for the next table.

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Step 7. Consider new $\hat{\mathrm{B}}^{-1}$ obtained in step 6 as $\hat{\mathrm{B}}^{-1}$ and go to step 3. Repeat the procedure until an optimum basic feasible solution is obtained or there is an indication of an unbounded solution.

## Advantages of revised simplex method :

The advantages of the revised simplex method over the regular simplex method are
(i) fewer calculations are required.
(ii) less storage is needed when computing the problem on a computer.
(iii) the round off errors can be controlled as table entries are not repeatedly recalculated.

### 2.5 Comparison of Simplex Method and Revised Simplex Method :

Let us consider the LPP
Maximize $z=c x$
subject to $\mathrm{A} x=b, x \geq 0$
where A is a matrix of order $m \times n$. If initially artificial variables are not needed for obtaining the initial basis matrix, then for solving this problem by the simple $x$ method we have to transfer $(n+1)$ columns at each iteration. ( $n$ columns for A and one column for $x_{\mathrm{B}}$ ). Also, at each iteration one variable is introduced into the basis and one is removed from it. Thus, in total we compute for $(n-m+1)$ columns. Further more, for each of these columns, we have to transform $(m+1)$ elements. For moving from one iteration to another we also need to calculate minium ratio $x_{\mathrm{B}} / y_{i k}$. Hence in all we have to perform multiplication $(m+1)(n-m+1)$ times and addition $m(n-m+1)$ times.

In the revised simplex method, there are $(m+1)$ rows and $(m+2)$ columns. So, for moving from one iteration to another we have to make $(m+1)^{2}$ multiplication operations to get an improved solution in addition to $m$ ( $n-m$ ) operations for calculating $\left(z_{j}-c_{j}\right)$ 's.

In the revised simplex method we need to make $(m+1)(m+2)$ entries in each table while in simplex method there are $(m+1)(n+1)$ entries in each table.

If the number of variables $n$ is significantly larger than the number of constraints m , then the coputational efforts of the revised simplex method is smaller than that of the simplex method.

Revised simplex method reduces the cumulative round-off error while calculating ( $z_{j}-c_{j}$ )'s and updated column $y_{k}$ due to the use of original data.

The inverse of the current basis matrix is obtained automatically.

### 2.6 Illustrative Examples :

Example 2.6.1. Use revised simplex method to solve the LPP.

$$
\begin{array}{ll}
\text { Maximize } & z=2 x_{1}-3 x_{2}+x_{3} \\
\text { subject to } & 3 x_{1}+6 x_{2}+x_{3} \leq 6 \\
& 4 x_{1}+2 x_{2}+x_{3} \leq 4 \\
& x_{1}-x_{2}+x_{3} \leq 3 \\
& x_{1}, x_{2}, x_{3} \geq 0 .
\end{array}
$$

Solution : Introducing slack variables $x_{4} \geq 0, x_{5} \geq 0, x_{6} \geq 0$, the given LPP becomes in standard form as

Maximize $\quad z=2 x_{1}-3 x_{2}+x_{3}+0 x_{4}+0 x_{5}+0 x_{6}$
subject to

$$
\begin{aligned}
& 3 x_{1}+6 x_{2}+x_{3}+x_{4}=x_{6} \\
& 4 x_{1}+2 x_{2}+x_{3}+x_{5}=4 \\
& x_{1}-x_{2}+x_{3}+x_{6} \\
&=3 \\
& x_{1}, x_{2}, x_{3}, x_{4}, x_{5}, x_{6} \geq 0
\end{aligned}
$$

or, Maximize $z=c x$
subject to $\mathrm{A} x=b, x \geq 0$
where $A=\left[\begin{array}{cccccc}3 & 6 & 1 & 1 & 0 & 0 \\ 4 & 2 & 1 & 0 & 1 & 0 \\ 1 & -1 & 1 & 0 & 0 & 1\end{array}\right], c=0\left[\begin{array}{llllll}2 & -3 & 1 & 0 & 0 & 0\end{array}\right]$

$$
b=\left[\begin{array}{l}
6 \\
4 \\
3
\end{array}\right] \text { and } x=\left[x_{1} x_{2} x_{3} x_{4} x_{5} x_{6}\right]^{\mathrm{T}}
$$

$\therefore$ we have $\hat{A}=\left[\begin{array}{c}\mathrm{A} \\ -c\end{array}\right]=\left[\begin{array}{cccccc}3 & 6 & 1 & 1 & 0 & 0 \\ 4 & 2 & 1 & 0 & 1 & 0 \\ 1 & -1 & 1 & 0 & 0 & 1 \\ -2 & 3 & -1 & 0 & 0 & 0\end{array}\right], b=\left[\begin{array}{l}b \\ 0\end{array}\right]=\left[\begin{array}{l}6 \\ 4 \\ 3 \\ 0\end{array}\right]$
Initially

$$
\mathrm{B}=\left[\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right], x_{\mathrm{B}}=\left[\begin{array}{l}
x_{4} \\
x_{5} \\
x_{6}
\end{array}\right], c_{\mathrm{B}}=\left[\begin{array}{lll}
c_{4} & c_{5} & c_{6}
\end{array}\right]=\left[\begin{array}{lll}
0 & 0 & 0
\end{array}\right]
$$

Now $C_{B} B^{-1}=\left[\begin{array}{lll}0 & 0 & 0\end{array}\right]\left[\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right]=\left[\begin{array}{lll}0 & 0 & 0\end{array}\right]$
$\therefore B^{-1}=\left[\begin{array}{cc}\mathrm{B}^{-1} & 0 \\ \mathrm{C}_{\mathrm{B}} \mathrm{B}^{-1} & 1\end{array}\right]=\left[\begin{array}{llll}1 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ 0 & 0 & 0 & 1\end{array}\right]$
$\therefore x_{\mathrm{B}}=\left[\begin{array}{c}x_{\mathrm{B}} \\ z\end{array}\right]=\left[\begin{array}{l}x_{4} \\ x_{5} \\ x_{6} \\ -z\end{array}\right]=\mathrm{B}^{-1} b=\left[\begin{array}{llll}1 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ 0 & 0 & 0 & 1\end{array}\right]\left[\begin{array}{l}6 \\ 4 \\ 3 \\ 0\end{array}\right]=\left[\begin{array}{l}6 \\ 4 \\ 3 \\ 0\end{array}\right]$
i.e. $\left[\begin{array}{l}x_{4} \\ x_{5} \\ x_{6}\end{array}\right]=\left[\begin{array}{l}6 \\ 4 \\ 3\end{array}\right]$ and $z=0$

The net evaluation are the components of

$$
\left.\begin{array}{rl} 
& {\left[\begin{array}{lll}
\mathrm{C}_{8} \mathrm{~B}^{-1} & 1
\end{array}\right]\left[\begin{array}{c}
\mathrm{A} \\
-c
\end{array}\right]=\left[\begin{array}{lll}
0 & 0 & 0
\end{array}\right]\left[\begin{array}{cccccc}
3 & 6 & 1 & 1 & 0 & 0 \\
4 & 2 & 1 & 0 & 1 & 0 \\
-2 & 3 & -1 & 0 & 0 & 0 \\
-2 & 3 & -1 & 0 & 0 & 0
\end{array}\right]=\left[\begin{array}{lllll}
-2 & 3 & -1 & 0 & 0
\end{array} 0\right.}
\end{array}\right]
$$

Since there are negative net evaluations, the solution obtained is not optimal. The most negative net evaluation is $z_{1}-c_{1}=-2$. Therefore $x_{1}$ will be the new basic variable.

Now we compute

$$
y_{1}=\mathrm{B}^{-1} a=\left[\begin{array}{llll}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0 \\
0 & 0 & 0 & 1
\end{array}\right]\left[\begin{array}{l}
3 \\
4 \\
1 \\
2
\end{array}\right]=\left[\begin{array}{c}
3 \\
4 \\
1 \\
-2
\end{array}\right]
$$

These results are shown in the following initial revised simplex table

| Basic variables | Values | $\hat{\mathbf{B}}^{-1}$ |  |  |  | $\hat{y}^{-1}$ | $\min$ <br> ratio |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $x_{4}$ | 6 | 1 | 0 | 0 | 0 | 3 | 2 |
| $x_{5}$ | 4 | 0 | 1 | 0 | 0 | 4 | 1 |
| $x_{6}$ | 3 | 0 | 0 | 1 | 0 | 1 | 3 |
| $z$ | 0 | 0 |  | 0 | 1 | -2 |  |

Here the minimum ratio is $\operatorname{Min}\left\{\frac{x_{B i}}{y_{i k}}: y_{i k}>0\right\}=1$ and the corresponding variable is $x_{5}$. Therefore, the outgoing basic variable is $x_{5}$. So $x_{5}$ is replaced by $x_{1}$ in the next table.

Using elementary row operations $\hat{y}_{1}=\left[\begin{array}{c}3 \\ 4 \\ 1 \\ -2\end{array}\right]$ is converted to $\left[\begin{array}{l}0 \\ 1 \\ 0 \\ 0\end{array}\right]$ and the same operations are done for $\hat{\mathrm{B}}^{-1}$. This gives new $\hat{\mathrm{B}}^{-1}$ as follows

$$
\hat{\mathbf{B}}^{-1}=\left[\begin{array}{cccc}
1 & -\frac{3}{4} & 0 & 0 \\
0 & \frac{1}{4} & 0 & 0 \\
0 & -\frac{1}{4} & 1 & 0 \\
0 & \frac{1}{2} & 0 & 1
\end{array}\right]
$$

The new BFS is given by

$$
\begin{aligned}
& \hat{x}=\left[\begin{array}{l}
x_{4} \\
x_{1} \\
x_{6} \\
x
\end{array}\right]=\hat{\mathrm{B}}^{-1} \cdot \hat{b}=\left[\begin{array}{cccc}
1 & -\frac{3}{4} & 0 & 0 \\
0 & \frac{1}{4} & 0 & 0 \\
0 & \frac{1}{2} & 0 & 1
\end{array}\right]\left[\begin{array}{l}
6 \\
4 \\
3 \\
0
\end{array}\right]=\left[\begin{array}{l}
3 \\
1 \\
2 \\
2
\end{array}\right] \\
& \therefore x_{\mathrm{B}}=\left[\begin{array}{l}
x_{4} \\
x_{1} \\
x_{6}
\end{array}\right]=\left[\begin{array}{l}
3 \\
1 \\
2
\end{array}\right] \text { and } \mathrm{E}=2
\end{aligned}
$$

The net evaluations are given by

$$
\begin{aligned}
x_{\mathrm{B}}= & {\left[\begin{array}{lll}
\mathrm{C}_{\mathrm{B}} & \mathrm{~B}^{-1} & 1
\end{array}\right]=\left[\begin{array}{c}
\mathrm{A} \\
-\mathrm{C}
\end{array}\right]=\left[\begin{array}{llll}
0 & \frac{1}{2} & 0 & 1
\end{array}\right]\left[\begin{array}{cccccc}
3 & 6 & 1 & 1 & 0 & 0 \\
4 & 2 & 1 & 0 & 1 & 0 \\
1 & -1 & 1 & 0 & 0 & 1 \\
-2 & 3 & -1 & 0 & 0 & 0
\end{array}\right] } \\
& =\left[\begin{array}{llllll}
0 & 4 & -\frac{1}{2} & 0 & \frac{1}{2} & 0
\end{array}\right]
\end{aligned}
$$

Since there is negative not evaluation, the BFS obtained is not optimal. Here the only negative net evaluation is $z_{3}-c_{3}=-\frac{1}{2}$. So, $x_{3}$ is the next incoming basic variable.

Now we compute

$$
y_{3}-\mathrm{B}^{-1} \hat{a}_{3}=\left[\begin{array}{cccc}
1 & -\frac{3}{4} & 0 & 0 \\
0 & \frac{1}{4} & 0 & 0 \\
0 & -\frac{1}{4} & 1 & 0 \\
0 & \frac{1}{2} & 0 & 1
\end{array}\right]\left[\begin{array}{c}
1 \\
1 \\
1 \\
-1
\end{array}\right]=\left[\begin{array}{c}
\frac{1}{4} \\
\frac{1}{4} \\
\frac{3}{4} \\
-\frac{1}{2}
\end{array}\right]
$$

These results are shown in the following simplex table

| Basic variables | Values | $\hat{\mathbf{B}}^{-1}$ |  |  |  | $\hat{y}^{-1}$ | $\min$ <br> ratio |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $x_{4}$ | 6 | 1 | $-\frac{3}{4}$ | 0 | 0 | $\frac{1}{4}$ | 12 |
| $x_{1}$ | 1 | 0 | $\frac{1}{4}$ | 0 | 0 | $\frac{1}{4}$ | 4 |
| $x_{6}$ | 2 | 0 | $-\frac{1}{4}$ | 1 | 0 | $\frac{3}{4}$ | $\frac{8}{3}$ |
| $z$ | 2 | 0 | $\frac{1}{2}$ | 0 | 1 | $-\frac{1}{2}$ |  |

Here the minimum ratio is $\frac{8}{3}$ and is associated with the basic variable $x_{6}$. Therefore, the outgoing basic variable is $x_{6}$. So $x_{6}$ is replaced by $x_{3}$ is the next iteration. Using elementary row operations $\hat{y}_{3}=\left[\begin{array}{c}\frac{1}{4} \\ \frac{1}{4} \\ \frac{3}{4} \\ -\frac{1}{2}\end{array}\right]$ is converted to $\left[\begin{array}{l}0 \\ 0 \\ 1 \\ 0\end{array}\right]$ and the same operations
are performed in $\hat{\mathrm{B}}^{-1}$. This gives the new $\hat{\mathrm{B}}^{-1}$ as follows

Now $\hat{\mathrm{B}}^{-1}=\left[\begin{array}{cccc}1 & -\frac{2}{3} & -\frac{1}{3} & 0 \\ 0 & \frac{1}{3} & -\frac{1}{3} & 0 \\ 0 & -\frac{1}{3} & \frac{4}{3} & 0 \\ 0 & \frac{1}{3} & \frac{2}{3} & 1\end{array}\right]$

The next BFS is given by

$$
\begin{gathered}
\hat{x}_{\mathrm{B}}=\left[\begin{array}{l}
x_{4} \\
x_{1} \\
x_{3} \\
z
\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}=\left[\begin{array}{cccc}
1 & -\frac{2}{3} & -\frac{1}{3} & 0 \\
0 & \frac{1}{3} & -\frac{1}{3} & 0 \\
0 & -\frac{1}{3} & \frac{4}{3} & 0 \\
0 & \frac{1}{3} & \frac{2}{3} & 1
\end{array}\right]\left[\begin{array}{l}
6 \\
4 \\
3 \\
0
\end{array}\right]=\left[\begin{array}{c}
\frac{7}{3} \\
\frac{1}{3} \\
\frac{8}{3} \\
\frac{10}{3}
\end{array}\right] \\
\therefore x_{\mathrm{B}}=\left[\begin{array}{l}
x_{4} \\
x_{1} \\
x_{3}
\end{array}\right]=\left[\begin{array}{l}
\frac{7}{3} \\
\frac{1}{3} \\
\frac{8}{3}
\end{array}\right] \text { and } z=\frac{10}{3}
\end{gathered}
$$

The net evaluation are given by

$$
\left[\begin{array}{ll}
\mathrm{C}_{\mathrm{B}} \mathrm{~B}^{-1}
\end{array}\right]\left[\begin{array}{l}
\mathrm{A} \\
-c
\end{array}\right]=\left[\begin{array}{llll}
0 & \frac{1}{3} & \frac{2}{3} & 1
\end{array}\right]\left[\begin{array}{cccccc}
3 & 6 & 1 & 1 & 0 & 0 \\
4 & 2 & 1 & 0 & 1 & 0 \\
1 & -1 & 1 & 0 & 0 & 1 \\
-2 & 3 & -1 & 0 & 0 & 0
\end{array}\right]=\left[\begin{array}{llllll}
0 & 3 & 0 & 0 & \frac{1}{3} & \frac{2}{3}
\end{array}\right]
$$

Here all net evaluations are non-negative. Hence we have obtained the optimal solution. The optimal solution is $x_{1}=\frac{1}{3}, x_{2}=0, \frac{x}{3}=\frac{8}{3}$ and $z=\frac{10}{3}$.

## Example 2.6.2. Solve by revised simplex method

Maximize $z=5 x_{1}+3 x_{2}$
subject to

$$
\begin{gathered}
4 x_{1}+5 x_{2} \leq 10 \\
5 x_{1}+2 x_{2} \leq 10 \\
3 x_{1}+8 x_{2} \leq 12 \\
x_{1}, x_{2} \geq 0
\end{gathered}
$$

Solution : Introducing surplus variable $x_{3} \geq 0$, slack variables $x_{4} \geq 0, x_{5} \geq 0$ and artificial variable $x_{6} \geq 0$ the standard form of the given LPP is

Maximize

$$
\begin{aligned}
& z=5 x_{1}+3 x_{2}+0 x_{3}+0 x_{4}+0 x_{5}-M x_{6} \\
& 4 x_{1}+5 x_{2}-x_{3}+x_{6}=10 \\
& 5 x_{1}+2 x_{2}+x_{4}=10 \\
& 3 x_{1}+8 x_{2}+x_{5}=12 \\
& x_{1}, x_{2}, x_{3}, x_{4}, x_{5}, x_{6} \geq 0
\end{aligned}
$$

or, Maximize $z=c x$
subject to $\mathrm{A} x=b, x \geq 0$
or, Maxmize $z=c x$
subject to. $\mathrm{A} x=b, x \geq 0$
where $\left.\mathrm{A}=\left[\begin{array}{cccccc}4 & 5 & -1 & 0 & 0 & 1 \\ 5 & 2 & 0 & 1 & 0 & 0 \\ 3 & 8 & 0 & 0 & 1 & 0\end{array}\right], c=\left[\begin{array}{lllll}5 & 3 & 0 & 0 & 0\end{array}\right]-\mathrm{M}\right]$

$$
b=\left[\begin{array}{l}
10 \\
10 \\
12
\end{array}\right], x=\left[\begin{array}{llllll}
x_{1} & x_{2} & x_{3} & x_{4} & x_{5} & x_{6}
\end{array}\right]^{\mathrm{T}}
$$

$\therefore$ We have $\mathrm{A}=\left[\begin{array}{l}\mathrm{A} \\ -c\end{array}\right]=\left[\begin{array}{cccccc}4 & 5 & -1 & 0 & 0 & 1 \\ 5 & 2 & 0 & 1 & 0 & 0 \\ 3 & 8 & 0 & 0 & 1 & 0 \\ -5 & -3 & 0 & 0 & 0 & \mathrm{M}\end{array}\right], b=\left[\begin{array}{l}b \\ 0\end{array}\right]=\left[\begin{array}{c}10 \\ 10 \\ 12 \\ 0\end{array}\right]$
Initially, $\mathrm{B}=\left[\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right], x_{\mathrm{B}}=\left[\begin{array}{l}x_{6} \\ x_{4} \\ x_{5}\end{array}\right]=\left[\begin{array}{l}10 \\ 10 \\ 12\end{array}\right], C_{B}=\left[\begin{array}{lll}c_{6} & c_{4} & c_{5}\end{array}\right]=\left[\begin{array}{lll}-\mathrm{M} & 0 & 0\end{array}\right]$
$\therefore \mathrm{B}^{-1}\left[\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right]$
Now $C_{B} B^{-1}=\left[\begin{array}{lll}-M & 0 & 0\end{array}\right]\left[\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right]=\left[\begin{array}{lll}-M & 0 & 0\end{array}\right]$
$\therefore B^{-1}=\left[\begin{array}{cc}B^{-1} & 0 \\ C_{B} B^{-1} & 1\end{array}\right]=\left[\begin{array}{cccc}1 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ -M & 0 & 0 & 1\end{array}\right]$
$\therefore \hat{x}_{\mathrm{B}}=\left[\begin{array}{c}x_{\mathrm{B}} \\ z\end{array}\right]=\left[\begin{array}{c}x_{6} \\ x_{4} \\ x_{5} \\ z\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}=\left[\begin{array}{cccc}1 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ -\mathrm{M} & 0 & 0 & 1\end{array}\right]\left[\begin{array}{c}10 \\ 10 \\ 12 \\ 0\end{array}\right]=\left[\begin{array}{c}10 \\ 10 \\ 12 \\ -10 \mathrm{M}\end{array}\right]$
$\therefore\left[\begin{array}{l}x_{6} \\ x_{4} \\ x_{5}\end{array}\right]=\left[\begin{array}{l}10 \\ 10 \\ 12\end{array}\right]$ and $z=-10 \mathrm{M}$
The net evaluations are the components of

$$
\begin{aligned}
& {\left[\begin{array}{ll}
c_{\mathrm{B}} \mathrm{~B}^{-1}
\end{array}\right]\left[\begin{array}{c}
\mathrm{A} \\
-c
\end{array}\right]=\left[\begin{array}{llll}
-\mathrm{M} & 0 & 0 & 1
\end{array}\right]\left[\begin{array}{cccccc}
4 & 5 & -1 & 0 & 0 & 1 \\
5 & 2 & 0 & 1 & 0 & 0 \\
3 & 8 & 0 & 1 & 0 & 0 \\
-5 & -3 & 0 & 0 & 0 & \mathrm{M}
\end{array}\right]} \\
& =[-4 \mathrm{M}-5-5 \mathrm{M}-3 \mathrm{M} 00000] \\
& =\left[z_{1}-c_{1} z_{2}-c_{2} z_{3}-c_{3} z_{4}-c_{4} z_{5}-c_{5} z_{6}-c_{6}\right]
\end{aligned}
$$

Since there are negative net evaluations, the solution obtained is not optimal. The most negative net evaluation is $z_{2}-c_{2}=-5 \mathrm{M}-3$. Therefore $x_{2}$ will be the new basic variable.

Now we compute

$$
\hat{y}_{2}=\hat{\mathrm{B}}^{-1} a_{2}=\left[\begin{array}{cccc}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0 \\
-\mathrm{M} & 0 & 0 & 1
\end{array}\right]\left[\begin{array}{c}
5 \\
2 \\
8 \\
-3
\end{array}\right]=\left[\begin{array}{c}
5 \\
2 \\
8 \\
-5 \mathrm{M}-3
\end{array}\right]
$$

These results are shown in the following initial revised simplex table

| Basic variables | Values | $\hat{\mathbf{B}}^{-1}$ |  |  |  | $\hat{\mathbf{y}}^{-1}$ | $\min$ <br> ratio |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $x_{6}$ | 10 | 1 | 0 | 0 | 0 | 5 | 2 |
| $x_{4}$ | 10 | 0 | 1 | 0 | 0 | 2 | 5 |
| $x_{5}$ | 10 | 0 | 0 | 1 | 0 | 8 | $\frac{3}{2}$ |
| $z$ | -10 M | -M | 0 | 0 | 1 | $-5 \mathrm{M}-3$ |  |

Here the minimum ratio is $\min \left\{\frac{x_{\mathrm{B} i}}{y_{i k}}: y_{i k}>0\right\}=\frac{3}{2}$ and the corresponding variable is $x_{5}$. Therefore, the outgoing basic variable is $x_{5}$. So $x_{5}$ is replaced by $x_{2}$ in the next table.

Using elementary row operations $\hat{y}_{2}-\left[\begin{array}{l}5 \\ 2 \\ 8 \\ -5 \mathrm{M}-3\end{array}\right]$ is converted to $\left[\begin{array}{l}0 \\ 0 \\ 1 \\ 0\end{array}\right]$ and the same operations are done for $\hat{\mathrm{B}}^{-1}$. This gives new $\hat{\mathrm{B}}^{-1}$ as follows

$$
\hat{\mathbf{B}}^{-1}=\left[\begin{array}{cccc}
1 & 0 & -\frac{5}{8} & 0 \\
0 & 1 & -\frac{1}{4} & 0 \\
0 & 0 & \frac{1}{8} & 0 \\
-M & 0 & \frac{5 M+3}{8} & 1
\end{array}\right]
$$

The new BFS is given by

$$
x_{B}=\left[\begin{array}{l}
x_{6} \\
x_{4} \\
x_{2} \\
z
\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}=\left[\begin{array}{cccc}
1 & 0 & -\frac{5}{8} & 0 \\
0 & 1 & -\frac{1}{4} & 0 \\
0 & 0 & \frac{1}{8} & 0 \\
-M & 0 & \frac{5 \mathrm{M}+3}{8} & 1
\end{array}\right]\left[\begin{array}{c}
10 \\
10 \\
12 \\
0
\end{array}\right]=\left[\begin{array}{c}
\frac{5}{2} \\
7 \\
\frac{3}{2} \\
\frac{-5 \mathrm{M}+9}{8}
\end{array}\right]
$$

$\therefore x_{\mathrm{B}}=\left[\begin{array}{l}x_{6} \\ x_{4} \\ x_{2}\end{array}\right]=\left[\begin{array}{l}\frac{5}{2} \\ 7 \\ \frac{3}{2}\end{array}\right]$ and $z=\frac{-5 M+9}{8}$
The net evaluation are the componentts of

$$
\begin{aligned}
& {\left[c_{\mathrm{B}} \mathrm{~B}^{-1}\right]\left[\begin{array}{c}
\mathrm{A} \\
-c
\end{array}\right]=\left[-M 0 \frac{5 M+3}{8}\right]\left[\begin{array}{cccccc}
4 & 5 & -1 & 0 & 0 & 1 \\
5 & 2 & 0 & 1 & 0 & 0 \\
3 & 8 & 0 & 0 & 1 & 0 \\
-5 & -3 & 0 & 0 & 0 & \mathrm{M}
\end{array}\right] } \\
= & {\left[\frac{-17 \mathrm{M}-3}{8} 0 \mathrm{M} 0 \frac{5 \mathrm{M}+3}{8} 0\right] }
\end{aligned}
$$

Since there is negative net evaluation, the BFS obtained is not optimal. Here the only negative net evaluation is $z_{1}-c_{1}$. So $x_{1}$ is the next incoming basic variable.

Now we compute.

$$
\hat{y}_{1}+\hat{\mathbf{B}}^{-1} a_{1}=\left[\begin{array}{cccc}
1 & 0 & -\frac{5}{4} & 0 \\
0 & 1 & -\frac{1}{4} & 0 \\
0 & 0 & \frac{1}{8} & 0 \\
-M & 0 & \frac{5 M+3}{8} & 1
\end{array}\right]\left[\begin{array}{c}
4 \\
5 \\
3 \\
-5
\end{array}\right]=\left[\begin{array}{c}
\frac{17}{8} \\
\frac{17}{4} \\
\frac{3}{8} \\
\frac{-17 \mathrm{M}-31}{8}
\end{array}\right]
$$

These results are shown in the following revised simplex table

| Basic variables | Values | $\hat{\mathbf{B}}^{-1}$ |  |  |  | $\hat{y}^{-1}$ | min <br> ratio |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $x_{6}$ | $\frac{5}{2}$ | 1 | 0 | $-\frac{5}{8}$ | 0 | $\frac{17}{8}$ | 2 |
| $x_{4}$ | 7 | 0 | 1 | $-\frac{1}{4}$ | 0 | $\frac{17}{4}$ | 5 |
| $x_{2}$ | $\frac{3}{2}$ | 0 | 0 | $\frac{1}{8}$ | 0 | $\frac{3}{8}$ | $\frac{3}{2}$ |
| $z$ | $\frac{-5 M+9}{8}$ | $-M$ | 0 | $\frac{5 M+3}{8}$ | 1 | $\frac{-17 M-31}{8}$ |  |

Here the minimum ratio is $\frac{20}{17}$ and is associated with the basic variable $x_{6}$. So $x_{6}$ is replaced by $x_{1}$ is the next iteration. Using elementary row operation $\hat{y}_{1}$ is converted to $\left[\begin{array}{l}1 \\ 0 \\ 0 \\ 0\end{array}\right]$ and the same operaions are performed in $\hat{\mathrm{B}}^{-1}$ as follows

$$
\therefore \text { Now } \hat{\mathrm{B}}^{-1}=\left[\begin{array}{cccc}
\frac{8}{17} & 0 & -\frac{5}{17} & 0 \\
-2 & 1 & 1 & 0 \\
-\frac{3}{17} & 0 & \frac{4}{17} & 0 \\
\frac{31}{17} & - & -\frac{13}{17} & 1
\end{array}\right]
$$

The next BFS is given by

$$
\begin{aligned}
& \hat{x}_{\mathrm{B}}=\left[\begin{array}{l}
x_{1} \\
x_{4} \\
x_{2} \\
z
\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}=\left[\begin{array}{cccc}
\frac{8}{17} & 0 & -\frac{5}{17} & 0 \\
-2 & 1 & 1 & 0 \\
-\frac{3}{17} & 0 & \frac{4}{17} & 0 \\
\frac{31}{17} & 0 & -\frac{13}{17} & 1
\end{array}\right]\left[\begin{array}{c}
10 \\
10 \\
12 \\
0
\end{array}\right]=\left[\begin{array}{c}
\frac{20}{17} \\
2 \\
\frac{18}{17} \\
\frac{154}{17}
\end{array}\right] \\
& \therefore x_{\mathrm{B}}=\left[\begin{array}{l}
x_{1} \\
x_{4} \\
x_{2}
\end{array}\right]=\left[\begin{array}{c}
\frac{20}{17} \\
2 \\
\frac{18}{17}
\end{array}\right] \text { and } z=\frac{154}{17}
\end{aligned}
$$

The net evaluation are given by

$$
\begin{aligned}
& {\left[\begin{array}{lll}
c_{\mathrm{B}} & \mathrm{~B}^{-1} & 1
\end{array}\right]\left[\begin{array}{c}
\mathrm{A} \\
-c
\end{array}\right]=\left[\begin{array}{lll}
\frac{31}{17} & 0 & -\frac{13}{17} 1
\end{array}\right]\left[\begin{array}{cccccc}
4 & 5 & -1 & 0 & 0 & 1 \\
5 & 2 & 0 & 1 & 0 & 0 \\
3 & 8 & 0 & 0 & 1 & 0 \\
-5 & -3 & 0 & 0 & 0 & \mathrm{M}
\end{array}\right]} \\
& =\left[\begin{array}{lll}
0 & 0-\frac{31}{17} 0-\frac{13}{17} \frac{31}{17}+\mathrm{M}
\end{array}\right]
\end{aligned}
$$

Since there are negative net evaluation the BFS obtained is not optimal. The most negative $z_{j}-c_{j}$ is $z_{3}-c_{3}=-\frac{31}{17}$ so $x_{3}$ is the next incoming basic variable.

Now we compute

$$
\hat{y}_{3}=\hat{\mathrm{B}}^{-1} \hat{a}_{3}=\left[\begin{array}{cccc}
\frac{8}{17} & 0 & -\frac{5}{17} & 0 \\
-2 & 1 & 1 & 0 \\
-\frac{3}{17} & 0 & \frac{4}{17} & 0 \\
\frac{31}{17} & 0 & -\frac{13}{17} & 1
\end{array}\right]\left[\begin{array}{c}
-1 \\
0 \\
0 \\
0
\end{array}\right]=\left[\begin{array}{c}
-\frac{8}{17} \\
\frac{2}{3} \\
\frac{3}{17} \\
-\frac{31}{17}
\end{array}\right]
$$

These results are shown in the following revised simplex table.

| Basic variables | Values | $\hat{\mathbf{B}}^{-1}$ |  |  | $\hat{\boldsymbol{y}}^{-1}$ | min <br> ratio |  |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $x_{1}$ | $\frac{20}{17}$ | 8 | 0 | $-\frac{5}{17}$ | 0 | $-\frac{8}{17}$ | $\ldots$ |
| $x_{4}$ | 2 | -2 | 1 | 1 | 0 | 2 | 1 |
| $x_{2}$ | $\frac{18}{17}$ | $-\frac{3}{17}$ | 0 | $\frac{4}{17}$ | 0 | 3 | 6 |
| $z$ | $\frac{154}{17}$ | $\frac{31}{17}$ | 0 |  | 1 | $-\frac{38}{17}$ |  |

Obviously $x_{4}$ will be replaced by $x_{3}$.
Using elementary row operations $\hat{y}_{3}$ is converted to $\left[\begin{array}{l}0 \\ 1 \\ 0 \\ 0\end{array}\right]$ and the same operations are used on $\hat{\mathrm{B}}^{-1}$. This gives new $\hat{\mathrm{B}}^{-1}$ as follows.

New $\hat{\mathbf{B}}^{-1}=\left[\begin{array}{cccc}0 & \frac{4}{17} & -\frac{1}{17} & 0 \\ -1 & \frac{1}{2} & \frac{1}{2} & 0 \\ 0 & -\frac{3}{34} & \frac{5}{34} & 0 \\ 0 & \frac{31}{34} & \frac{5}{34} & 1\end{array}\right]$

The next BFS is given by

$$
\begin{aligned}
& \hat{x}_{\mathrm{B}}=\left[\begin{array}{l}
x_{1} \\
x_{3} \\
x_{2} \\
z
\end{array}\right]=\hat{\mathrm{B}}^{-1} \hat{b}=\left[\begin{array}{cccc}
0 & \frac{4}{17} & -\frac{1}{17} & 0 \\
-1 & \frac{1}{2} & \frac{1}{2} & 0 \\
0 & -\frac{3}{34} & \frac{5}{34} & 0 \\
0 & \frac{31}{34} & \frac{5}{34} & 1
\end{array}\right]\left[\begin{array}{c}
10 \\
10 \\
12 \\
0
\end{array}\right]=\left[\begin{array}{c}
\frac{28}{17} \\
1 \\
\frac{15}{17} \\
\frac{185}{17}
\end{array}\right] \\
& \therefore x_{\mathrm{B}}=\left[\begin{array}{l}
x_{1} \\
x_{3} \\
x_{2}
\end{array}\right]=\left[\begin{array}{c}
\frac{28}{17} \\
1 \\
\frac{15}{17}
\end{array}\right] \text { and } z=\frac{185}{17}
\end{aligned}
$$

The net evaluations are given by

$$
\left.\begin{array}{l}
{\left[\begin{array}{ll}
c_{\mathrm{B}} & \mathrm{~B}^{-1}
\end{array}\right]\left[\begin{array}{c}
\mathrm{A} \\
-c
\end{array}\right]=\left[\begin{array}{lll}
0 & \frac{31}{34} & \frac{5}{34}
\end{array}\right]\left[\begin{array}{cccccc}
4 & 5 & -1 & 0 & 0 & 1 \\
5 & 2 & 0 & 1 & 0 & 0 \\
3 & 8 & 0 & 0 & 1 & 0 \\
-5 & -3 & 0 & 0 & 0 & \mathrm{M}
\end{array}\right]} \\
\quad=\left[\begin{array}{llll}
0 & 0 & 0 & \frac{31}{34}
\end{array} \frac{5}{34} \mathrm{M}\right.
\end{array}\right]
$$

Here all net evaluation are found to be non-negative. Hence wehave obtained the optimal solution. The optimal solutions is given by

$$
x_{1}=\frac{28}{17}, x_{2}=\frac{15}{17} \text { and } z_{\max }=\frac{185}{17} .
$$

### 2.7 Summary :

Revised simplex method is an efficient method and is very useful for large problem. Only necessary part of the simplex table is calculated to pass from one table to the next table. Standard form of the revised simplex method is devised and computational procedure of revised simplex method is noted and is compared with simplex method. Finally, the method is used to solve some examples.

### 2.8 Self Assessment Questions :

Use revised simplex method to solve the following LPP

1. Maximize $\quad z=3 x_{1}+5 x_{2}$
subject to $\quad x_{1} \leq 4$

$$
\begin{aligned}
& x_{2} \leq 6 \\
& 3 x_{1}+2 x_{2} \leq 18 \\
& x_{1}, x_{2} \geq 0
\end{aligned}
$$

[Ans : $x_{1}=2, x_{2}=6, z_{\max }=36$ ]
2. Maximize $\quad z=6 x_{1}-2 x_{2}+3 x_{3}$
subject to

$$
2 x_{1}-x_{2}+2 x_{3} \leq 2
$$

$$
x_{1} \quad+4 x_{3} \leq 4
$$

$$
x_{1}, x_{2}, x_{3} \geq 0
$$

[Ans: $x_{1}=4, x_{2}=6, x_{3}=0 z_{\text {max }}=12$ ]
3. Maximize $\quad z=x_{1}+x_{2}$
subject to $\quad x_{1}+2 x_{2} \geq 7$
$4 x_{1}+x_{2} \geq 6$
$x_{1}, x_{2} \geq 0$
$\left[\right.$ Ans: $\left.x_{1}=\frac{5}{7}, x_{2}=\frac{22}{7}, z_{\min }=\frac{27}{7}\right]$
4. Maxmize $\quad z=2 x_{1}+x_{2}$
subject to $\quad 3 x_{1}+x_{2} \leq 3$

$$
\begin{aligned}
& 4 x_{1}+3 x_{2} \geq 6 \\
& x_{1}+2 x_{2} \leq 3 \\
& x_{1}, x_{2} \geq 0 \\
& {\left[\text { Ans: } x_{1}=\frac{3}{5}, x_{2}=\frac{6}{5}, z_{\text {min }}=\frac{12}{5}\right]} \\
& \text { 5. Minimize } \quad z=4 x_{1}+3 x_{2} \\
& \text { subject to } \quad 3 x_{1}+4 x_{2} \leq 12 \\
& 3 x_{1}+3 x_{2} \leq 10 \\
& 2 x_{1}+x_{2} \leq 4 \\
& x_{1}, x_{2} \geq 0 \\
& \text { [Ans: } \left.x_{1}=\frac{4}{5}, x_{2}=\frac{12}{5}, z_{\max }=\frac{52}{5}\right] \text {. }
\end{aligned}
$$

## Unit $3 \square$ Dual Simplex Method

Structure
3.1 Introduction
3.2 Comparison Between Simplex Method and Dual Simplex Method
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3.1 Introduction :

The Dual Simplex Method gives an algorithm in which we start with a basic optimal solution of the primal in which all $z j-e j \geq 0$ but not feasible is some basic solution are negative. At each iteration the number of negative basic variables are decreased while maintaining the optimality. An optimal solutionis reached in a finite number of steps. The benifit of this procedure lies in the fact that we need not take the help of any artificial variable and hence it reduces a lot of labour.

### 3.2 Comparison Between Simplex Method and Dual Simplex Method :

In simplex method the initial solution is basic feasible and non optimal. In subsequent tables the value of the objective function gradually increases and
finally reaches to its obtimal value. In each table the solution is basic feasible and nonoptimal.

In dual simplex method the initial solution is basic non-feasible and optimal. In subsequent tables the value of the objective function gradually decreases and finally reaches to its optimal value. In each table the solution is basic non-feasible and optimal (or better than optimal).

Basic non-feasible and better than optimal some $x_{B}<0$ and all $z j-c j \geq 0, z_{\mathrm{B}} \geq z_{\text {max }}$.
(Dual simplex method)
z gradually decreases
Basic feasible and optimal
$x_{\mathrm{B}} \geq 0$ and all $z j-c j \geq 0$ $z_{\mathrm{B}}=z_{\text {inax }}$
(Simplex method)
z gradually increases
Basic feasible and non optimal
$x_{\mathrm{B}} \geq 0$ and some $z j-c j<0, z_{\mathrm{B}} \leq z_{\text {max }}$

### 3.3 Applications of the Dual Simplex Method :

It the given LPP is optimal and infeasible then only dual simplex method is applicable for many practical problem the initial table does not satisty these conditions and as a consequence dual simplex method can not be applied. Simplex method has no such restnction and is applicable to any LPP: Hence as rule the regular simplexmethod preferred over the dual simplex method for solving the general LPP. However, there are instances when the dual simplex method has a distinct advantage over the regular simplex method. There are problems in which a dual feasible table is readily available to start the dual simplex method and for such problems the optimal BFS is obtained easily in comparison to simplex method. Some of the applications of dual simplex method are :
(i) Sensitivity analysis when the right hand side vector be changed or when new constraints are added.
(ii) Parametric programming.
(iii) Integer programming problem.
(iv) Some non-linear programming problem.

### 3.4 Criteria for Incoming and Outgoing basic Variable in Dual Simplex Method :

In the transmation formula of simplex method if the basic variable $x \mathrm{~B}_{r}$ is replaced by the non-basic variable $x_{k}$ then we have

$$
\begin{align*}
& \quad \hat{z}=z-\frac{{ }^{x} \mathrm{~B}_{r}}{{ }_{r} r_{k}}\left(z_{k}-c_{k}\right)  \tag{1}\\
& \text { and }\left(\hat{z}_{i}-\hat{c}_{j}\right)-\left(z_{j}-c_{i}\right)-\frac{z_{r i}}{y_{r k}}\left(z_{k}-c_{k}\right)
\end{align*}
$$

As we want to remove negative besic variables, we choose the most negative besic variable $x \mathrm{~B}_{r}$ (say) as outgoing basic variable from the list of all basic variables.

We know that if for some basic solution (not necessarily feasible) all components of net evaluations are non-negative then the value of the objective function to this basic solution is optimal or better than optimal. So we intend to lower down the value of the objective function to get $z_{\text {max }}$. For this from (1) we should have $y_{r k}<0$ as $x \mathrm{~B},<0$ and $z_{k}-c_{k} \geq 0$. This should be one criterion for incoming basic variable.

In the next table we want the solution to be optimal or better than optimal. So we should have $\hat{z}_{j}-\hat{c}_{j} \geq 0$ for all $j$.
$\therefore$ From (2) We have

$$
\begin{align*}
& z_{j}-c_{j}-\frac{y_{j j}}{y_{r k}}\left(z_{k}-c_{k}\right) \geq 0 \text { for all } j \\
& \text { or, } z_{j}-c_{j} \geq \frac{y_{j j}}{y_{r k}}\left(z_{k}-c_{k}\right) \text { for all } j \tag{3}
\end{align*}
$$

When $y j \geq 0$ then (3) is satisfied as $y_{n k}<0$ and all $z j-c j \geq 0$.
When $y j<0$ then (3) is satisfied if $\frac{z_{j}-c_{j}}{y_{r j}} \leq \frac{z_{k}-c_{k}}{y_{r k}}$ for all $j$
Hence we are $b$ choose $k$ such that

$$
\max _{y_{r j}<0}\left\{\frac{z_{j}-c_{j}}{y_{r j}}\right\}=\frac{z_{k}-c_{k}}{y_{r k}}
$$

### 3.5 Dual Simplex Algorithm :

The iterative procedure for dual simplex algorithm are as follows :
Step 1 : Convert the minimizatism LPP into that of maximization if it is in the minimization form.

Step 2: Convert the $\geq$ type inequalious, representing the constraints of the given LPP, if any, into those of $\leq$ type by multiplying the corresponding constraints by -1 .

Step 3 : Introduce slack variables in the constraints of the given LPP and obtain an initial basic solution. Put this solution in the starting dual simplex table.

Step 4 : Test the nature of the net evaluations $z j-c j$ in the starting simplex table.
(i) If all $z j-e j$ and $x \mathrm{~B} j$ are non negative for all $i$ and $j$, then an optimum basic fesible solution has been obtained.
(ii) It all $z j-c j$ are non negative and at least one basic variable, say $x \mathrm{~B}_{r^{\prime}}$ is negative then go to step 5 .
(iii) It at least one $z j-c j$ is negative then dual simplex method is not applicable. In this case we are to apply artificial constraint method.

Step 5 : Select the most negative basic variable, say $x \mathrm{~B}_{k}$, as outgoing basic variable.
Step 6 : Test the nature of all $y_{j}, j, j=1, z, \ldots, n$.
(i) It all $y_{k} j$ are non-negative, there does not exist any feasible solution to the given LPP.
(ii) It at least one $y_{k} j$ is negative, then comute

$$
\left\{\frac{z j}{y_{k j}}: y_{k j}<0\right\}, j=1,2, \ldots, n
$$

and choose the maximum of these, If the maximum of these be $\frac{z_{r}-c_{r}}{y_{k r}}$ then $x_{r}$ is the incoming basic variable i.e. $x \mathrm{~B}_{k}$ is replaced by $x_{r}$.

Step 7 : With $y_{k r}$ as the key element form the next table. Using elementary row operation convert the key element to unity and all other elements of the key column to zero to get the improoed solution.

Step 8: Repeat the steps 4 to 7 until either an optimun basic feasible solution is obtained or there is an indication of no feasible solution.

### 3.6 Illustrative Examples :

Example 3.6.1. Solve the following LPP by dual Simplex Method.

$$
\begin{array}{ll}
\text { Maximize } & z=2 x_{1}+x_{2} \\
\text { Subject to } & 3 x_{1}+x_{2} \geq 3 \\
& 4 x_{1}+x_{2} \geq 6 \\
& x_{1}+2 x_{2} \geq 3 \\
& x_{1}, x_{2} \geq 0 .
\end{array}
$$

Solution : Converting the given LPP into maximization and changing all $\geq$ type inequations to $\leq$ type and finally adding slack variables $x_{3} \geq 0, x_{4} \geq 0, x_{5} \geq 0$, the reformulated LPP in its standard form becomes.

$$
\begin{aligned}
\text { Maximize } \begin{aligned}
& z^{\prime}=-2 x_{1}-x_{2}+o x_{3}+o x_{4}+o x_{5} \\
& \text { Subject to }=3 \\
&-3 x_{1}-x_{2}+x_{3}=3 \\
&-4 x_{1}-x_{2}+x_{4}=6 \\
&-x_{1}-2 x_{2}+x_{5}=3 \\
& x_{1}, x_{2}, x_{3}, x_{4}, x_{5}, x_{6} \geq 0 .
\end{aligned}
\end{aligned}
$$

Solution : The solution of this LPP. by dual simplex method is shown in the following tables.


In the first table $\max \left\{-\frac{1}{2},-1\right\}=-\frac{1}{2}$ and is ascociated with $y_{1} \therefore y_{4}$ is replaced by $y_{1}$ for the second table.

In the second table there is only one ratio $-\frac{2}{7}$ and is associated with. $y_{2}$.
$\therefore y_{s}$ is replaced by $y_{2}$ for the third table.
In the third table all $x \mathrm{~B}_{i}$ are non negative. So. this is optimal table. The optimal solution is $x_{1}=\frac{9}{7}, x_{2}=\frac{6}{7}$ and $z_{\max }^{\prime}=-\frac{24}{7} \quad \therefore z_{\min }=-z_{\max }^{\prime}=\frac{24}{7}$.

Example 3.6.2. Solve the following LPP by dual Simplex Method.

$$
\begin{array}{ll}
\text { Maximize } & z=-2 x_{1}-2 x_{2}-4 x_{3} \\
\text { Subject to } & 2 x_{1}+3 x_{2}+5 x_{3} \leq 2 \\
3 x_{1}+x_{2}+2 x_{3} & \geq 3 \\
x_{1}+4 x_{2}+6 x_{3} & \geq 5 \\
x_{1}, x_{2}, x_{3} & \geq 0
\end{array}
$$

Solution : © onverting the $\geq$ type inequations into $\leq$ type and introducing the slack variable. $x_{4} \geq C, x_{5} \geq 0, x_{6} \geq 0$ the given LPP can be written in the staridard form as

$$
\begin{aligned}
& \text { Maximize } \begin{array}{ll}
z=-2 x_{1}-2 x_{2}-4 x_{3}+0 x_{4}+0 x_{5}+0 x_{6} \\
\text { Subject to } & 2 x_{1}+3 x_{2}+5 x_{3}+x_{4} \\
& =2 x_{1}-x_{2}-2 x_{3}+x_{5}=-3 \\
& -x_{1}-4 x_{2}-6 x_{3}+x_{6}=-5 \\
& x_{1}, x_{2}, x_{3}, x_{4}, x_{6} \geq 0 .
\end{array}
\end{aligned}
$$

The following tables are obtained by using dual simplex method to this LPP.


In the first table $x_{6}=-5$ is the most negative basic variable and max $\left\{-2,-\frac{1}{2},-\frac{2}{3}\right\}=-\frac{1}{2}$ which is associated with this non basic variable $x_{2}$. So $x_{6}$ is replaced by $x_{2}$.

In the second table $x_{4}=-\frac{7}{4}, x_{5}=-\frac{7}{4}$ are the most negative basic variables. We choose $x_{5}$ arbitrarily. Here $\max \left\{-\frac{6}{11},-2,-2\right\}=-\frac{6}{11}$ which is associated with the non basic variable $x_{1}$. So $x_{5}$ is replaced by $x_{1}$.

In the third table $x \mathrm{~B}_{1}=x_{4}<0$ and all $y_{1} j \geq 0$.
$\therefore$ The given LPP has no feasible solution.

### 3.7 Modification of Dual Simplex Method :

If the initial table of the dual simplex method contains some negative basic variables and some of the net-evaluations are negtive then the dual simplex method is not applicable. In such situation dual simplex method is to be modified to from an equivalent LPP in which some basic variables are negative but all netevaluations are non-negative. Hence standard dual simplex method can be applied to that equivalent LPP.

The artificial constraint is one such method. In this method we consider the variables corresponding to which the net evaluations are negative and the variable corresponding to the most negative component of net evaluations is noted. $\mathrm{Lt} z_{p}-c_{p}$ be the most negative net evaluation. So we consider the corresponding variable $x_{p}$. In this method we have to consider the antificial constraint.

$$
\sum x_{j} \leq M
$$

Where $\Sigma$ is extended over all $j$ 's for whioh $z j-c j<0$ and $M$ is a sufficiently large positive number. Adding slack variable $x_{\mathrm{M}}$ to this constraint we get

$$
\Sigma x j+r_{M}=M
$$

From this we find $x_{p}$ as

$$
x_{p}=\mathrm{M}-\left(x_{\mathrm{M}}+\sum_{j \neq p} x_{j}\right)
$$

This $x_{p}$ is then subptituted in the orginal objective function and in the set of all constraints. This new problem together with the new added artificial constraint is equivalent to the given problem. This equivalent LPP will have all $z j-c j \geq 0$. Thus dual samplex method can be applied.

### 3.8 Illustrative Examples :

Example 3.8.1 : Use the artificial constraint method to find the initial basic solution of the following problem and then apply the dual simplex algorithm to solve it

$$
\begin{array}{ll}
\text { Maximize } & z=-2 x_{1}-x_{2}-x_{3} \\
\text { Subject to } & 4 x_{1}+6 x_{2}+3 x_{3} \leq 8 \\
-x_{1}+9 x_{2}-x_{3} & \geq 3 \\
2 x_{1}+3 x_{2}-5 x_{3} & \geq 4 \\
x_{1}, x_{2}, x_{3} & \geq 0
\end{array}
$$

Solution : We first convert the minimization problem to maximization and then change the inequation of $\geq$ type into $\leq$ type. Finally adding slack variables $x_{4} \geq 0$, $x_{5} \geq 0, x_{6} \geq 0$ we get the standard form LPP in dual simplex method as

Maximize $z^{\prime}=2 x_{1}+x_{2}+x_{3}+o x_{4}+o x_{5}+o x_{6}$
Subject to $4 x_{1}+6 x_{2}+3 x_{3}+x_{4}=8$

$$
\begin{aligned}
& x_{1}-9 x_{2}+x_{3}+x_{5}=-3 \\
& -2 x_{1}-3 x_{2}+5 x_{3}+x_{6}=-4 \\
& x_{1}, x_{2}, x_{3}, x_{4}, x_{5}, x_{6} \geq 0 .
\end{aligned}
$$

The initial dual simplex table is

|  | $c j$ | 2 | 1 | 1 | 0 | 0 | 0 |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ |
| 0 | $y_{4}$ | 8 | 4 | 6 | 3 | 1 | 0 | 0 |
| 0 | $y_{5}$ | -3 | 1 | -9 | 1 | 0 | 1 | 0 |
| 0 | $y_{6}$ | -4 | -2 | -3 | 5 | 0 | 0 | 1 |
| $z^{\prime}=0$ |  | $2 j-c j$ | -2 | -1 | -1 | 0 | 0 | 0 |

Here there are negative net evaluation, so standard dual simplex method is not applicable.

The negative net evaluations are $z_{1}-c_{1}, z_{2}-c_{2}, z_{3}-c_{3}$ \& most negative net evaluation is $z_{1}-c_{1}=-2$.
$\therefore$ The artificial constraint is
$x_{1}+x_{2}+x_{3} \leq \mathrm{M}$ where M is a very large positive number. Adding slack variable $x_{\mathrm{m}}$ we have

$$
x_{1}+x_{2}+x_{3}+x_{\mathrm{M}}=\mathrm{M}
$$

From this we have $x_{1}=\mathrm{M}-x_{2}-x_{3}-x_{\mathrm{M}}$
Using this in the LPP and adding the artificial constraint we have.
Maximize $\quad z^{\prime}=2\left(\mathrm{M}-x_{2}-x_{3}-x_{M}\right)+x_{2}+x_{3}+o x_{4}+o x_{5}+o x_{6}$
Subject to $4\left(M-x_{2}-x_{3}-x_{M}\right)+6 x_{2}+3 x_{3}+x_{4}=8$

$$
\begin{array}{ll}
\left(\mathrm{M}-x_{2}-x_{3}-x_{\mathrm{M}}\right)-9 x_{2}+x_{3}+x_{5} & =-3 \\
-2\left(\mathrm{M}-x_{2}-x_{3}-x_{\mathrm{M}}\right)-3 x_{3}+5 x_{3}+x_{6} & =-4 \\
x_{1}+x_{2}+x_{3}+x_{\mathrm{M}} & =\mathrm{M} \\
x_{1}, x_{2}, x_{3}, x_{4}, x_{5}, x_{6}, x_{\mathrm{M}} \geq 0 &
\end{array}
$$

or, Maximize $z^{\prime}=-2 x_{M}-x_{2}-x_{3}+o x_{4}+o x_{5}+2 \mathrm{M}$

$$
\text { Subject to }-4 x_{M}+2 x_{2}-x_{3}+x_{4} \quad=8-4 M
$$

$$
\begin{array}{ll}
-x_{\mathrm{M}}-10 x_{2}+x_{3} & =-3-\mathrm{M} \\
2 x_{\mathrm{M}}-x_{2}+7 x_{3}+x_{6} & =-4+2 \mathrm{M} \\
x_{\mathrm{M}}+x_{1}+x_{2}+x_{3} & =\mathrm{M} \\
x_{1}, x_{2}, x_{3}, x_{4}, x_{5}, x_{6}, x_{\mathrm{M}} \geq 0 &
\end{array}
$$

The following tables are obtained using simplex method.


Here all basic variable are non-negative. So thes is the optimal table. The optimal solution is $x_{1}=\frac{9}{7} \quad x_{2}=\frac{10}{21}, x_{3}=0$ and $z_{\max }^{\prime}=\left(-2 M+\frac{64}{21}\right)+2 M=\frac{64}{21}$ Therefore $z_{\text {min }}=-z_{\text {max }}^{\prime}=-\frac{64}{21}$.

Example 3.8.2. Use the atificial constrant method to dind the initial basic solution of the following problem and then apply the dual simplex algorithm to solve it :

$$
\begin{aligned}
& \text { Maximize } \quad z=2 x_{1}-3 x_{2}-2 x_{3} \\
& \text { Subject to } x_{1}-2 x_{2}-3 x_{3}=8 \\
& 2 x_{2}+x_{3} \leq 10 \\
& \\
& x_{2}-2 x_{3} \geq 4 \\
& \\
& \\
& x_{1}, x_{2}, x_{3} \geq 0 .
\end{aligned}
$$

Solution : We first change the inequation of $\geq$ type into $\leq$ type. Adding slack variable $x_{4} \geq 0, x_{5} \geq 0$ we get the standard form of the LPP in dual simplex method as

$$
\begin{aligned}
& \text { Maximize } \quad z=2 x_{1}-3 x_{2}-2 x_{3} \\
& \text { Subject to } x_{1}-2 x_{2}-3 x_{3}=8 \\
& 2 x_{2}+x_{3}+x_{4}=10 \\
& -x_{2}+2 x_{3}+x 5=-4 \\
& \\
& x_{1}, x_{2}, x_{3}, x_{4}, x_{5} \geq 0
\end{aligned}
$$

The initial dual simplex table is

|  | $c j$ | 2 | -3 | -2 | 0 | 0 |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{3}$ |
| 2 | $y_{1}$ | 8 | 1 | -2 | -3 | 0 | 0 |
| 0 | $y_{4}$ | 10 | 0 | 2 | 1 | 1 | 0 |
| 0 | $y_{5}$ | -4 | 0 | -1 | 2 | 0 | 1 |
| $z=16$ |  | $z j-c j$ | 0 | -1 | -4 | 0 | 0 |

Since these are negative net evaluations, standard dual simplex method is not applicable. The negative net evaluations are $z_{2}-c_{2}$ and $z_{3}-c_{3}$ and most negative net evaluation is $z_{3}-c_{3}=-4$.
$\therefore$ The artificial constraint is $x_{2}+x_{3} \leq \mathrm{M}$ where M is a very large positive number. Adding plack variable $x_{\mathrm{M}}$ we have

$$
x_{2}+x_{3}+x_{M}=M
$$

From this we have $x_{3}=M-x_{2}-x_{M}$
Using this in the LPP and adding the artificial constraint we have the equivalent LPP as

Maximize $z=2 x_{\mathrm{M}}+2 x_{1}-x_{2}-2 \mathrm{M}$
Subject to

$$
\begin{array}{ll}
3 x_{\mathrm{M}}+x_{1}+x_{2} & =3 \mathrm{M}+8 \mathrm{a} \\
-x_{\mathrm{M}}+x_{2}+x_{4} & =-\mathrm{M}+10 \\
-2 x_{\mathrm{M}}-3 x_{2}+x \% & =-2 \mathrm{M}-4 \\
x_{\mathrm{M}}+x_{2}+x_{3} & =\mathrm{M} \\
x_{\mathrm{M}}, x_{1}, x_{2}, x_{3}, x_{4}, x_{5} \geq 0
\end{array}
$$

The dual simplex tables are as follows.


In this table all basic variables are non-negative. So this is the optimal table. The optimal solution is $x_{1}=\frac{94}{5}, x_{2}=\frac{24}{5}, x_{3}=\frac{2}{5}$ and $z_{\max }=\frac{10 M+112}{5}-2 M=\frac{112}{5}$.

### 3.9 Summary :

Dual simplex method is found to be very useful is a large class of LPP. It is simple to handle and sige of the tables are not large as no artificial variables are introduced, the method is illustrated through examples. The method is then modified to handle more LPR.

### 3.10 Self Assessment Questions :

1. Use dual simplex method to solve the LPP

Maximize $z=-2 x_{1}-3 x_{2}-x_{3}$
Subject to

$$
\begin{aligned}
& 2 x_{1}+x_{2}+2 x_{3} \geq 3 \\
& 3 x_{1}+2 x_{2}+x_{3} \geq 4 \\
& x_{1}, x_{2}, x_{3} \geq 0
\end{aligned}
$$

[Ans. $x_{1}=\frac{5}{4}, x_{2}=0, x_{3}=\frac{1}{4}, z_{\max }=-\frac{11}{4}$ ]
2. Use dual simplex method to solve the LPP

Maximize $z=10 x_{1}+6 x_{2}+2 x_{3}$
Subject to $\quad-x_{1}+x_{2}+x_{3} \geq 1$

$$
\begin{aligned}
& 3 x_{1}+x_{2}-x_{3} \geq 2 \\
& x_{1}, x_{2}, x_{3} \geq 0
\end{aligned}
$$

[Ans. $x_{1}=\frac{1}{4}, x_{2}=\frac{5}{4}, x_{3}=0, z_{\text {min }}=10$ ]
3. Solve by dual simplex method the fllowing LPP

Maximize $z=6 x_{1}+x_{2}$

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Subject to $\quad 2 x_{1}+x_{2} \geq 3$

$$
\begin{aligned}
& x_{1}+x_{2} \geq 0 \\
& x_{1}, x_{2} \geq 0
\end{aligned}
$$

[Ans. $\left.x_{1}=1, x_{2}=1, z_{\min }=7\right]$
4. Solve the following LPP by dual simplex method

Maximize $z=-3 x_{1}-2 x_{2}$
Subject to $\quad x_{1}+x_{2} \geq 1$
$x_{1}+x_{2} \leq 7$
$x_{1}+2 x_{2} \geq 10$
$x_{2} \leq 3$

$$
x_{1}, x_{2} \geq 0
$$

[Ans. $x_{1}=4, x_{2}=3, z_{\mathrm{ma}} x=-18$ ]
5. Solve by dual simplex method :

Maximize $z=2 x_{1}+3 x_{2}$
Subject to

$$
\begin{aligned}
& 2 x_{1}+3 x_{2} \leq 30 \\
& x_{1}+2 x_{2} \geq 10 \\
& x_{1}-x_{2} \geq 0 \\
& x_{1} \geq 5 \\
& x_{2} \geq 0
\end{aligned}
$$

[Ans. $x_{1}=5, x_{2}=\frac{5}{2}, z_{\text {min }}=\frac{35}{2}$ ]
6. Solve the following LPP by daul simplex method Maximize $z=x_{1}+x_{2}$

Subject to $\quad 2 x_{1}+x_{2} \geq 2$

$$
-x_{1}-x_{2} \geq 1
$$

$$
x_{1}, x_{2} \geq 0
$$

[Ans. No feasible solution]
7. Using artificial constraint procedure, solve the follwing problem by dual simplex method and show that the problem has no feasible solution

Maximize $z=-x_{1}+x_{2}$
Subject to $\quad x_{1}-4 x_{2} \geq 5$
$x_{1}-3 x_{2} \leq .1$
$2 x_{1}-5 x_{2} \geq 1$
$x_{1}, x_{2} \geq 0$
8. Use the artificial constraint method to find the initial basic solution of the following problem and then apply the dual simplex algorthm to solve it

Maximize $z=x_{1}-3 x_{2}-2 x_{3}$
Subject to $\quad x_{2}-2 x_{3} \geq 2$

$$
x_{1}-4 x_{2}-6 x_{3}=8
$$

$$
2 x_{2}+x_{3} \leq 5
$$

$$
x_{1}, x_{2}, x_{3} \geq 0
$$

[Ans. $x_{1}=\frac{94}{5}, x_{3}=\frac{12}{5}, x_{3}=\frac{1}{5}, z_{\max }=\frac{56}{5}$ ]

## Unit 4 Post Optimality Analysis

Structure
4.1 Introduction
4.2 Discrete changes In The Cost Vector
4.3 Illustrative Example
4.4 Discrete Change In The Requirement Vector
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4.13 Self Assessment Questions
4.1 Introduction :

In reality the problem accuring are in general large in size and often an error is discovered in the data after the attainment of an optimal solution to the problem. In such a situation there are two alternatives, either to solve the problem from begining or to device some method to use the optimal table. Undoubtedly the second one willsave time and space and is named as post optimality analysis. Also in practical situation the values of the co-efficient matrix A, the components of the requirement vector and the cost vector or neither known exactly nor they are constant for all time and or all situations. so it is important to know how sensitive the optimal solution is to small changes in these
parameters. By sensitiveness we mean fulfilments of the condition of optimality as well as determining the limits of variations of these parameters for the solution to remain optimal.

We shall study the following effects of changes in the
(i) co-efficients $c_{i}$ of the objective function.
(ii) components of the requirement vector to
(iii) addition of a new variable
(iv) deletion of a variable
(v) addition of a new constraint

### 4.2 Discrete Changes In The Cost Vector :

Let $x_{\mathrm{B}}$ be the optimal basic solution of the LPP
Maximize $z=\mathrm{cx}$
Subject to $\mathrm{Ax}=\mathrm{b}$

$$
x \geq 0
$$

Where $c, x^{\mathrm{T}} \in \mathrm{R}^{n}, b^{\mathrm{T}} \in \mathrm{R}^{m}$ and A is mxn an real matrix. Let $\Delta c_{k}$ be the amount by which $c_{k}$ is changed. So the new value of $c_{k}$ is $c_{k}^{*}=c_{k}+\Delta c_{k}$.

We know that $x_{\mathrm{B}}=\mathrm{B}^{-1} b$ and so it independent of $c$.
As initially $x_{\mathrm{B}}$ was BFS it will remain so after the change. The optimality condition is $z_{j}-c_{j} \geq 0$ for all $j$ i.e. $\left[c_{\mathrm{B}} \mathrm{B}^{-1} 1\right]\left[\begin{array}{c}\mathrm{A} \\ -c\end{array}\right] \geq 0$. It invalues $c$. So change in $c$ will affect this condition. Thus when $c_{x}$ is changed to $c_{k}^{*}$, the solution $x_{B}$ may or may not remain optimal solution though it remains BFS.

Two cases will arise
(i) $c_{k}$ is not in $c_{B}$
(ii) $c_{k}$ is in $c_{\mathrm{B}}$

Case (i). Here $c_{k}$ is not in $c_{B}$. The net evaluations are the components of $c_{B} B^{-1} A-c_{1}$ and as $x_{\mathrm{B}}=\mathrm{B}^{-1} b$ was optimal solution we have $c_{\mathrm{B}} \mathrm{B}^{-1} \mathrm{~A}-c \geq 0$. i.e. $z_{j}-c_{j} \geq 0 \forall j$.

We note that when $c_{k}$ is changed to $c_{k}^{*}$ only $k$ th component of net evaluation will change Thus for all $j=1,2, \ldots, k-1, k+1, \ldots, n$ i.e. for all $j \neq k$ we have new net evaluations.

$$
\begin{aligned}
& z_{j}^{*}-c_{j}^{*}=z_{j}-c_{j} \geq 0\left[\text { as } z_{j}-c_{j} \geq 0 \text { for all } j\right] \\
& \text { for } j=k \text { we have } z_{k}^{*}-c_{k}^{*}=z_{k}-\left(c_{k}+\Delta c_{k}\right)=\left(z_{k}-c_{k}\right)-\Delta c_{k}
\end{aligned}
$$

We have $z_{k}-c_{k} \geq 0$. Therefor for all $\Delta c_{k}, z_{k}^{*}-c_{k}^{*}$ will not remain non negative.
Thus $x_{\mathrm{B}}$ will remain optimal solution for the changed LPP if $z_{k}-c_{k}-\Delta c_{k} \geq 0$ i.e. if $\Delta c_{k} \leq z_{k}-c_{k}$.

Case (ii). Here $c_{k}$ is one component of $c_{\mathrm{B}}$. Let $c_{k}=c_{B_{k}}$ and so $x_{k}$ is a basic variable. Thus $y_{k}$ is a unit vector with its $\lambda_{t h}$ component as 1 .

The new value of $z_{k}-c_{k}$ is given by

$$
z_{k}^{*}-c_{k}^{*}=\sum_{\substack{i=1 \\ i \neq \lambda}}^{m} c_{B_{i}} y_{i_{k}}+c_{k}^{*} \cdot 1-c_{k}^{*} \cdot 1=0 \quad\left[\because y_{i k}=0 \forall \mathrm{i} \neq \lambda\right] .
$$

For $j \neq k$, new value of $z_{j}-c_{j}$ is given by

$$
\begin{aligned}
z_{j}^{*}-c_{j}^{*} & =\left(\sum_{\substack{i=1 \\
i \neq \lambda}}^{m} c_{B_{i}} y_{i_{j}}+c_{k}^{\prime \prime} \cdot y_{\lambda_{j}}\right)-c_{j}^{*} \\
& =\sum_{i=\lambda} c_{B i} y_{i j}+\left(c_{k}+\Delta c_{k}\right) y_{\lambda j}-c_{j}\left[\because c_{j}^{u}=c_{j} \forall j \neq k\right] \\
& \left.=\sum_{i \neq \lambda} c_{B i} y_{i j}+c_{B_{\lambda}} y_{\lambda_{j}}+\Delta c_{k} y_{\lambda j}-c_{j}\right) \cdot\left[\because c_{k}=y_{B \lambda}\right] \\
& =\sum_{i=1}^{m} c_{B j} y_{i j}+\Delta c_{k} y_{\lambda j}-c_{j} \\
& =z_{j}-c_{j}+\Delta c_{k} y_{\lambda i}\left[\because z_{j}=\sum_{i=1}^{m} c_{B i} y_{i j}\right]
\end{aligned}
$$

$\therefore x_{\mathrm{B}}$ remains optimal solution

$$
\begin{aligned}
& \text { if } z_{j}-c_{j}+\Delta c_{k} \lambda_{j} \geq 0 \forall i \neq k \\
& \text { i.e. if } \Delta c_{k} y_{\lambda_{j}} \geq-\left(z_{j}-c_{j}\right) \forall j \neq k
\end{aligned}
$$

Now for $y_{\lambda j}=0$ this condition is fulfilled automatically as $z_{j}-c_{j} \geq 0$.
For $y_{\lambda j}>0$ this condition is satisfied if $\Delta c_{k} \geq-\frac{z_{j}-c_{i}}{y_{\lambda}} \forall j \neq k$
i.e. if $-\frac{z_{j}-c_{i}}{y_{\lambda j}} \leq \Delta c_{k} \forall j \neq k$
$\therefore$ We must have $\max _{\substack{y_{j}>0 \\ j \neq k}}\left\{-\frac{z_{j}-c_{j}}{y_{\lambda i}}\right\} \leq \Delta c_{k}$

$$
y_{\lambda j}<U \text { this condition is satistied if } \Delta c_{k} \leq-\frac{z_{j}-c_{j}}{y_{\lambda j}} \forall j \neq k
$$

$\therefore$ We must have $\Delta c_{k} \leq \min _{\substack{y_{j}>0 \\ j \neq k}}\left\{-\frac{z_{j}-c_{j}}{y_{\lambda j}}\right\}$
These two conditions can be combined as

$$
\max _{\substack{y_{j}>0 \\ j \neq k}}\left\{-\frac{z_{j}-c_{j}}{y_{\lambda_{j}}}\right\} \leq \Delta c_{k} \leq \min _{\substack{y_{j}<0 \\ j \neq k}}\left\{-\frac{z_{j}-c_{j}}{y_{\lambda_{j}}}\right\}
$$

Hence if $\Delta c_{k}$ lies in this range then the solution $\mathrm{x}_{\mathrm{B}}$ remain optimal and if $\Delta c_{k}$ falls outsids this range then at least one $z_{j}-c_{j}$ will be negative and the solution will no longer remain optimal.

It no $y_{\lambda_{j}}>0$, then there is no lower bownd of $\Delta c_{k}$ and if no $y_{\lambda_{j}}<0$, then there is no upper bound of $\Delta c_{k}$.

### 4.3 Illustrative Example :

### 4.3.1 The optimal solution of the LPP :

Maximize $z=6 x_{1}-2 x_{2}+3 x_{3}$
Subject to $2 x_{1}-x_{2}+2 x_{3} \leq 2$

$$
\begin{aligned}
& x_{1} \quad+4 x_{3} \leq 4 \\
& x_{1}, x_{2}, x_{3} \geq 0 .
\end{aligned}
$$

is contained in the table.

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 6 | $y_{1}$ | 4 | 1 | 0 | 4 | 0 | 1 |
| -2 | $y_{2}$ | 6 | 0 | 1 | 6 | -1 | 2 |
| $z_{j}-c_{i}$ |  | $z=12$ | 0 | 0 | 9 | 2 | 2 |

Find the ranges of the cost components when (i) changed one at a time (ii) changed two at a time (iii) changed all three at a time to keep the optimal solution same.

Solution :
(i) When one component is changed at a time :

For change of $c_{1}=6$ to $c_{1}^{*}$ we have the corresponding changed table as

|  |  | $c_{j}$ | $c_{1}^{*}$ | -2 | 3 | 0 | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ |
| $c_{1}^{*}$ | $\mathrm{y}_{1}$ | 4 | 1 | 0 | 4 | 0 | 1 |
| -2 | $\mathrm{y}_{2}$ | 6 | 0 | 1 | 6 | -1 | 2 |
|  |  |  | 0 | 0 | $4 c_{1}^{*}-12$ | 2 | $c_{1}^{*}-4$ |

This table becomes optimal table
if $4 c_{1}^{*}-12 \geq 0$ and $c_{\mathrm{i}}^{*}-4 \geq 0$
i.e. if $c_{1}^{*} \geq 3$ and $c_{1}^{*} \geq 4$
i.e. if $c_{1}^{*} \geq 4$

For change of $c_{2}=-2$ to $c_{2}^{*}$ the table corresponding to the final table becomes.

|  |  |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ |
| 6 | $\mathrm{y}_{1}$ | 4 | 1 | 0 | 4 | 0 | 1 |
| $c_{2}^{*}$ | $\mathrm{y}_{2}$ | 6 | 0 | 1 | 6 | -1 | 2 |
|  |  |  | 0 | 0 | $24+6 c_{2}^{*}$ | $-c_{2}^{*}$ | $6+2 c_{2}^{*}$ |

This table becomes the optimal table

$$
y 24+2 c_{2}^{*} \geq 0 \text { and }-c_{2}^{*} \geq 0 \text { and } 6+2 c_{2}^{*} \geq 0
$$

i.e. if $c_{2}^{*} \geq-4$ and $c_{2}^{*} \leq 0$ and $c_{2}^{*} \geq-3$
i.e. if $-3 \leq c_{2}^{*} \leq 0$

For change of $c_{3}=3$ to $c_{3}^{*}$ the modified table is

| $c$ | $c$ | -2 | $c_{3}^{*}$ | 0 | 0 |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ |
| 6 | $y_{1}$ | 4 | 1 | 0 | 4 | 0 | 1 |
| -2 | $y_{2}$ | 6 | 0 | 1 | 6 | -1 | 2 |
|  |  |  | 0 | 0 | $12+c_{3}^{\circ}$ | 2 | 2 |

This table remains optimal table

$$
\text { if } 12-c_{3}^{*} \geq 0
$$

i.e. if $c_{3}^{*} \leq 12$
(ii) When two components are changed at a time.

For the change of $c_{1}=6$ and $c_{2}=-2$ to $c_{1}^{*}$ and $c_{2}^{*}$ the modified table is

|  |  |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $c_{2}^{*}$ | 3 | $y_{2}$ | $y_{3}$ |
| $c_{1}^{*}$ | $y_{1}$ | 4 | 1 | 0 | 4 | 0 | $y_{5}$ |
| $c_{2}^{*}$ | $y_{2}$ | 6 | 0 | 1 | 6 | -1 | 2 |
|  |  |  | 0 | 0 | $4 c_{1}^{*}+6 c_{2}^{*}-3$ | $-c_{2}^{*}$ | $c_{2}^{*}+2 c_{2}^{*}$ |

This table becomes optimal table if all $\varepsilon_{j}-c_{j} \geq 0$
i.e. if $4 c_{1}^{*}+6 c_{2}^{*}-3 \geq 0$ and $-c_{2}^{*} \geq 0$ and $c_{1}^{*}+2 c_{2}^{*} \geq 0$
i.e. if $c_{2}^{*} \geq \frac{3-4 c_{1}^{*}}{6}$ and $c_{2}^{*} \leq 0$ and $c_{2}^{*} \geq-\frac{c_{1}^{*}}{2}$
i.e. $\max \left\{\frac{3-4 c_{1}^{*}}{6}, \frac{-c_{1}^{*}}{2}\right\} \leq c_{2}^{*} \leq 0$ and $c_{1}^{*}$ any real number.

For the change of $c_{1}=6$ and $c_{3}=3$ to $c_{1}^{*}$ and $c_{3}^{*}$ respectively the modified table is

|  |  |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}^{*}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}^{*}$ | $y_{5}$ |
| $c_{1}^{*}$ | $y_{1}$ | 4 | 1 | 0 | 4 | 0 | 1 |
| -2 | $y_{2}$ | 6 | 0 | 1 | 6 | -1 | 2 |
|  |  |  | 0 | 0 | $4 c_{1}^{*}-12 c_{2}^{*}-c_{3}^{*}$ | 2 | $c_{1}^{*}-4$ |

This table remains optimal table if all $z_{j}-c_{j} \geq 0$
i.e. if $4 c_{\mathrm{i}}^{\circ}-12-c_{3}^{*} \geq 0$ and $c_{\mathrm{i}}^{*}-4 \geq 0$
i.e. if $c_{1}^{*} \geq \frac{12+c_{3}^{*}}{4}$ and $c_{1}^{*} \geq 4$
i.e. if $c_{1}^{*} \geq \max \left\{4,3+\frac{c_{3}^{*}}{4}\right\}$ and $c_{3}^{*}$ any real number.

For the change of $c_{2}=-2$ and $c_{3}=3$ to $c_{2}^{*}$ and $c_{3}^{*}$ respectively the modified table is

|  | 6 |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}^{*}$ | $y_{1}$ | $y_{2}$ | $c_{3}^{*}$ | 0 | 0 |
| 6 | $y_{1}$ | 4 | 1 | 0 | 4 | $y_{4}$ | $y_{5}$ |
| $c_{2}^{*}$ | $y_{2}$ | 6 | 0 | 1 | 6 | 0 | 1 |
|  |  |  | 0 | 0 | $24+6 c_{2}^{*}-c_{3}^{*}$ | $-c_{2}^{*}$ | $6+2 c_{2}^{*}$ |

This table remains optimal table if all $z j-c j \geq 0$
i.e. if $24+6 c_{2}^{*}-c_{3}^{*} \geq 0$ and $-c_{2}^{*} \geq 0$ and $6+2 c_{2}^{*} \geq 0$
i.e. if $c_{2}^{*} \geq \frac{2 c_{3}^{*}-24}{6}$ and $c_{2}^{*} \leq 0$ and $c_{2}^{*} \geq-3$
i.e. if $\max \left\{\frac{c_{3}^{*}-24}{6},-3\right\} \leq c_{2}^{*} \leq 0$ and $c_{3}^{*}$ any real number.

When all the three components are changed together :
It $c_{1}=6, c=-2, c_{3}=3$ be changed respectively to $c_{1}^{*}, c_{2}^{*}, c_{3}^{*}$.
The modified table obtained from old optimal table is

| $c_{B}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}^{*}$ | $c_{3}^{*}$ | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{1}^{*}$ | $y_{1}$ | 4 | 1 | 0 | 4 | $y_{4}$ | $y_{5}$ |
| $c_{2}^{*}$ | $y_{2}$ | 6 | 0 | 1 | 6 | -1 | 2 |
|  |  |  | 0 | 0 | $4 c_{1}^{*}-6 c_{2}^{*}-c_{3}^{*}$ | $-c_{2}^{*}$ | $c_{1}^{*}+2 c_{2}^{*}$ |

This table becomes an optimal table if all $z_{j}-c_{j} \geq 0$.
i.e. if $4 c_{1}^{*}+6 c_{2}^{*}-c_{3}^{*} \geq 0$ and $-c_{2}^{*} \geq 0$ and $c_{1}^{*}+2 c_{2}^{*} \geq 0$
i.e. if $c_{2}^{*} \geq \frac{6 c_{3}^{*}-4 c_{1}^{*}}{6}$ and $c_{2}^{*} \leq 0$ and $c_{2}^{*} \geq-\frac{c_{1}^{*}}{2}$
i.e. if $\max \left\{\frac{c_{3}^{*}-4 c_{i}^{*}}{6},-\frac{c_{i}^{*}}{2}\right\} \leq c_{3}^{*} \leq 0$ and $c_{i}^{*}$ any real number and $c_{3}^{*}$ any real number.

### 4.4 Discrete Change In The Requirement Vector :

Let $x_{B}$ be the optimal BFS of the LPP
Maximize

$$
z=c x
$$

$$
\text { subject to } \quad A x=b, x \geq 0
$$

where $c, x^{\mathrm{T}} \in \mathrm{R}^{n}, b^{\mathrm{T}} \in \mathrm{R}^{m}$ and A is an max real matrix. We have $x_{\mathrm{B}}=\mathrm{B}^{-1} b$ and the net evaluations are the components $\left[c_{B_{B}} B^{-1} 1\right]\left[\begin{array}{c}A \\ -c\end{array}\right]$. From these we see that $x_{\mathrm{B}}$ depends on $b$ but net evaluations are independent of $b$ : So change made in $b$ will not affect optimality conditions ie. optimal solution will remain optimal but it will change the solution $x_{\mathrm{B}}$ and it may become negative i.e. infeasible.

Let the component $b_{k}$ of $b$ be changed to $b_{k}^{*}=b_{k}+\Delta b_{k}$.
So the old solution $x_{\mathrm{B}}=\mathrm{B}^{-1} b$ becomes

$$
x_{\mathrm{B}}^{*}=\mathrm{B}^{-1} b^{\bullet} \text { where } b^{*}=\left[b_{1}, b_{2}, \ldots, b_{k-1}, b_{k}+\Delta b_{k}, b_{k+1}, \ldots, b_{m}\right]^{\mathrm{T}}
$$

$$
\text { Let }\left[\begin{array}{cccc}
b_{11} & b_{12} & \ldots & b_{1 m} \\
b_{21} & b_{22} & \cdots & b_{2 m} \\
\because:! & :: & & : \because \\
b_{m 1} & b_{m 2} & & b_{m m}
\end{array}\right]
$$

$$
\begin{aligned}
\therefore x_{\mathrm{B}}^{*} & =\left[\begin{array}{cccc}
b_{11} & b_{12} & \cdots & b_{1 m} \\
b_{21} & b_{22} & \cdots & b_{2 m} \\
\cdots: & \cdots: & & \cdots: \\
b_{m 1} & b_{m 2} & & b_{m m}
\end{array}\right]\left[\begin{array}{c}
b_{2} \\
\vdots \\
b_{k-1} \\
b_{k}+\Delta b_{k} \\
b_{k+1} \\
\vdots \\
b_{m}
\end{array}\right] \\
& =\left[\begin{array}{cccc}
b_{11} & b_{12} & \cdots & b_{1 m} \\
b_{21} & b_{22} & \cdots & b_{2 m} \\
\cdots & \cdots & & \cdots \\
\cdots & \cdots & \cdots & \cdots \\
\cdots & \cdots & & \cdots \\
b_{m 1} & b_{m 2} & & b_{m m n}
\end{array}\right]\left[\begin{array}{c}
b_{1} \\
b_{2} \\
\vdots \\
b_{k-1} \\
b_{k} \\
b_{k+1} \\
\vdots \\
b_{m}
\end{array}\right]+\left[\begin{array}{cccccc}
b_{11} & b_{12} & \cdots & b_{1 k} & \cdots & b_{1 m} \\
b_{21} & b_{22} & \cdots & b_{2 k} & \cdots & b_{2 m} \\
\cdots & \cdots & & \cdots & \cdots & \\
\cdots & \cdots & & \cdots & \cdots \\
\cdots & \cdots & \cdots & \cdots \\
b_{m 1} & b_{m 2} & & b_{m k} & \cdots & b_{m n}
\end{array}\right]\left[\begin{array}{c}
0 \\
0 \\
\vdots \\
0 \\
0 \\
0 \\
\vdots \\
0
\end{array}\right]
\end{aligned}
$$

$=\mathrm{B}^{-1}+\left[\begin{array}{cc}b_{1 k} & \Delta b_{k} \\ b_{2 k} & \Delta b_{k} \\ \dddot{b_{m k}} & \dddot{\Delta} \ddot{b}_{k}\end{array}\right]$
Thus $\left[\begin{array}{c}x_{b_{1}}^{*} \\ x_{B_{2}}^{*} \\ \vdots \\ x_{B_{m}}^{\theta_{m}}\end{array}\right]=\left[\begin{array}{c}x_{B_{1}} \\ x_{B_{2}} \\ \vdots \\ x_{b_{m}}\end{array}\right]+\left[\begin{array}{cc}b_{1 k} & \Delta b_{k} \\ b_{2 k} & \Delta b_{k} \\ \vdots \\ b_{k} & \Delta b_{k}\end{array}\right]$
$x_{\mathrm{B}_{i}}^{*}=x_{\mathrm{B}_{j}}+b_{i k} \Delta b_{k}$ for all $\mathrm{i}=1,2, \ldots, \mathrm{M}$
As we have noted this solution is optimal or better than optimal but may not be feasible shough basic.

Thus $x_{\mathrm{B}}^{*}$ will be an optimal BFS if $x_{\mathrm{B}_{i}}^{*} \geq 0$ for all $i=1,2, \ldots, m$
i.e. if $x_{B_{i}}+b_{i k} \Delta b_{k} \geq 0$ for all $i=1,2, \ldots, m$
i.e. if $b_{i k} \Delta b_{k} \geq-x_{\mathrm{B}_{j}}$, for all $i=1,2, \ldots, m$

For all $b_{i k}=0$ this condition is satisfied.
For ali $b_{i k}>0$ this condition is satisfied if $\Delta b_{k} \geq-\frac{x_{B_{j}}}{b_{i k}}$
$\therefore$ We need $-\frac{x_{B_{i}}}{b_{i k}} \leq \Delta b_{k}$ for all $b_{i k}>0$
i.e. we need $\max \left\{-\frac{x_{B_{i}}}{b_{i k}}: b_{l k}>0\right\} \leq \Delta b_{k}$

Again for all $b_{i k}<0$ this condition is satisfied if $\Delta b_{k} \leq-\frac{x_{i_{i}}}{b_{i k}}$
$\therefore$ We need $\Delta b_{k} \leq-\frac{x_{b_{i}}}{b_{i k}}$ for all $b_{i k}<0$
i.e. we need $\Delta b_{k} \leq \min \left\{-\frac{x_{B_{i}}}{b_{i k}}: b_{i k}<0\right\}$

Hence $x_{B_{i}}$ will be optimal basefeasible solution if $\Delta \mathrm{b}_{\mathrm{k}}$ is selected satisfying the condition.

$$
\max \left\{-\frac{x_{B_{i}}}{b_{i k}} ; b_{i k}>0\right\} \leq \Delta b_{k} \leq \min \left\{-\frac{x_{B_{i}}}{b_{i k}}: b_{i k}>0\right\}
$$

### 4.5 Illustrative Examples :

Example 4.5.1. Given the LPP
Maximize $\quad z=-x_{1}+2 x_{2}-x_{3}$
Subject to $3 x_{1}+x_{2}-x_{3} \leq 10$
$-x_{1}+4 x_{2}+x_{3} \geq 6$
$x_{2}+x_{3} \leq 4$
$x_{1}, x_{2}, x_{3} \geq 0$
Determine the ranges for discrete changes of the components of $b$ when changed one at a time, so as to maintain the optimality of the current optimum solution for the LPP.

Solution : Introducing slack variables $x_{4} \geq 0, x_{6} \geq 0$, surplus variable $x_{5} \geq 0$ and artificial variable $x_{7} \geq 0$ we have the standard form as follows

Maximize $\quad z=-x_{1}+2 x_{2}-x_{3}+o x_{4}+o x_{5}+o x_{6}-\mathrm{M} x_{7}$
Subject to $3 x_{1}+x_{2}-x_{3}+x_{4} \quad=10$

$$
\begin{array}{cc}
-x_{1}+4 x_{2}+x_{3}-x_{5}+x_{7}=6 \\
x_{2}+x_{3} & =6 \\
x_{1}, x_{2}, x_{3}, x_{4}, x_{5}, x_{6}, x_{7} \geq 0 & =4 \\
&
\end{array}
$$

The tables obtained by simplex method are as follows :

|  |  | $c_{j}$ | -1 | 2 | -1 | 0 | 0 | 0 | -M |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |
| 0 | $y_{4}$ | 10 | 3 | 1 | -1 | 1 | 0 | 0 | 0 |
| -M | $y_{7}$ | 6 | -1 | 4 | 1 | 0 | -1 | 0 | 1 |
| 0 | $y_{6}$ | 4 | 0 | 1 | 1 | 0 | 0 | 1 | 0 |
|  |  |  | $\mathrm{M}+1$ | $-4 \mathrm{M}-2$ | $-\mathrm{M}+2$ | 0 | M | 0 | 0 |
| 0 | $y_{4}$ | $\frac{17}{2}$ | $\frac{13}{4}$ | 0 | $-\frac{5}{4}$ | 1 | $\frac{1}{4}$ | 0 | $-\frac{1}{4}$ |
| 2 | $y_{2}$ | $\frac{3}{2}$ | $-\frac{1}{4}$ | 1 | $\frac{1}{4}$ | 0 | $-\frac{1}{4}$ | 0 | $\frac{1}{4}$ |
| 0 | $y_{6}$ | $\frac{5}{2}$ | $\frac{1}{4}$ | 0 | $\frac{3}{4}$ | 0 | $\frac{1}{4}$ | 1 | $-\frac{1}{4}$ |
|  |  |  | $\frac{1}{2}$ | 0 | $\frac{3}{2}$ | 0 | $-\frac{1}{2}$ | 0 | $\mathrm{M}+\frac{1}{2}$ |
| 0 | $y_{4}$ | 6 | 3 | 0 | -2 | 1 | 0 | -1 | 0 |
| 2 | $y_{2}$ | 4 | 0 | 1 | 1 | 0 | 0 | 1 | 0 |
| 0 | $y_{5}$ | 10 | 1 | 0 | 3 | 0 | 1 | 4 | -1 |
|  |  |  | 1 | 0 | 3 | 0 | 0 | 2 | M |

In this final table the basis in $\mathrm{B}=\left[a_{4} a_{2} a_{5}\right]$ and in the initial table the basis is I $=\left[\begin{array}{lll}a_{4} & a_{7} & a_{6}\end{array}\right]$

The inverse of the basis in the final table is given by

$$
\mathrm{B}^{-1}=\left[\begin{array}{ccc}
1 & 0 & -1 \\
0 & 0 & 1 \\
0 & -1 & 4
\end{array}\right]=\left[\begin{array}{lll}
b_{11} & b_{12} & b_{13} \\
b_{21} & b_{22} & b_{23} \\
b_{31} & b_{32} & b_{33}
\end{array}\right]
$$

When $b_{1}$ is changed to $b_{1}+\Delta b_{1}$ then the range $b \Delta b_{1}$ such that the optimality of the new BFS is not violated is given by

$$
\max \left\{-\frac{x_{B i}}{b_{i 1}}: b_{i k}>0\right\} \leq \Delta b_{1} \leq \min \left\{-\frac{x_{B i}}{b_{i 1}}: b_{i 1}<0\right\}
$$

$\therefore \max \left\{-\frac{x_{B 1}}{b_{11}}\right\} \leq \Delta b_{1}\left[\because\right.$ only $b_{11}=1>0$ and there are $x_{0}$ negative $\left.b_{i 1}\right]$
or, $-\frac{6}{1} \leq \Delta b_{1}$ or, $\Delta \mathrm{b}_{1} \geq-6$
$\therefore \mathrm{b}_{1}+\Delta \mathrm{b}_{1} \geq \mathrm{b}_{1}-6$
or, $b_{1}^{*} \geq 10-6$
or, $b_{1}^{*} \geq 4$
When $b_{2}$ is changed to $b_{2}+\Delta b_{2}$ then the range of $\Delta b_{2}$ such that the optimality of the new BFS is not violated is given by

$$
\begin{aligned}
& \max \left\{-\frac{x_{\mathrm{B} i}}{b_{i 2}}: b_{i 2}>0\right\} \leq \Delta \mathrm{b}_{2} \leq \min \left\{-\frac{x_{\mathrm{B} i}}{b_{i 2}}: b_{i 2}>0\right\} \\
& \therefore \Delta b_{2} \leq \min \left\{-\frac{x_{B i}}{b_{i 2}}\right\}\left[\therefore \text { only } b_{32}=-1<0 \text { and there is no positive } b_{i 2}\right]
\end{aligned}
$$

or, $\Delta b_{2} \leq-\frac{10}{-1}$
or, $\Delta b_{2} \leq 10$
$\therefore b_{2}+\Delta b_{2} \leq b_{2}+10$
or, $b_{2}^{0} \leq 6=10$
or, $b_{2}^{*} \leq 16$
When $b_{3}$ is changed to $b_{3}+\Delta b_{3}$ thin the ranges of $\Delta b_{3}$ such that the optimality of the new BFS is not violated are given by

$$
\begin{aligned}
& \qquad \max \left\{-\frac{x_{\mathrm{B} i}}{b_{i 3}}: b_{i 3}>0\right\} \leq \Delta b_{3} \leq \min \left\{-\frac{x_{\mathrm{B} i}}{b_{i 3}}: b_{i 3}<0\right\} \\
& \text { or, } \max \left\{-\frac{x_{\mathrm{B} 2}}{b_{23}}, \frac{x_{\mathrm{B} 3}}{b_{33}}\right\} \leq \Delta b_{3} \leq \min \left\{-\frac{x_{\mathrm{B} 1}}{b_{13}}\right\}
\end{aligned}
$$

or, $\max \left\{-\frac{4}{1}, \frac{-10}{4}\right\} \leq \Delta b_{3} \leq \min \left\{\frac{-6}{-1}\right\}$
or, $-\frac{5}{2} \leq \Delta b_{3} \leq 6$
$\therefore-\frac{5}{2}+b_{3} \leq b_{3}+\Delta b_{3} \leq 6+b_{3}$
or, $-\frac{5}{2}+4 \leq b_{3} \leq 6+4$
or, $\frac{3}{2} \leq b_{3}^{*} \leq 10$
Example 4.5.2 Consider the LPP

$$
\begin{array}{cc}
\text { Maximize } & z=2 x_{1}+x_{2}+4 x_{3}-x_{4} \\
\text { subject to } & x_{1}+2 x_{2}+x_{3}-3 x_{4} \leq 8 \\
x_{2}+x_{3}+2 x_{4} \leq 0 \\
2 x_{1}+7 x_{2}-5 x_{3}-10 x_{4} \leq 21 \\
x_{1}, x_{2}, x_{3}, x_{4} \geq 0
\end{array}
$$

The optimal solution is it is contained in the following table

| $c_{\mathrm{B}}$ |  |  |  |  |  |  |  | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |  |  |  |
| 0 | $y_{1}$ | 1 | 1 | 0 | 3 | 1 | 1 | 2 | 0 |
| 1 | $y_{2}$ | 0 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |
| 0 | $y_{7}$ | 5 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |

For each of the parameter change listed below, make the necessary correction in the optimal table and solve the resulting problem.
(a) change $c_{1}$ to 1
(b) change $c$ to $\left[\begin{array}{llll}1 & 2 & 3 & 4\end{array}\right]$
(c) change $b$ to $[3-24]^{\mathrm{T}}$
(d) change $b_{2}$ to 11
(e) How much $c_{1}$ be changed without affecting the optimal solution.

Solution : (a) When $c_{1}$ is changed to 1 the modified form of the optimal table becomes

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | $y_{1}$ | 8 | 1 | 0 | 3 | 1 | 1 | 2 | 0 |
| 1 | $y_{2}$ | 0 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |
| 0 | $y_{7}$ | 5 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | -2 | 0 | 1 | 1 | 0 |

From this table we see that changed solution is not optimal as $z_{3}-c_{3}<0$. So we are to apply simplex method to get the optimal solution

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | $y_{1}$ | 8 | 1 | 0 | 3 | 1 | 1 | 2 | 0 | min <br> 1 |
| $y_{2}$ | 0 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |  |  |
| 0 | $y_{7}$ | 5 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |  |
|  |  |  | 0 | 0 | -2 | 0 | 1 | 1 | 0 |  |
| 4 | $y_{3}$ | 8 | $\frac{8}{3}$ |  |  |  |  |  |  |  |
| 1 | $y_{2}$ | $\frac{1}{3}$ | 0 | 1 | $\frac{1}{3}$ | $\frac{1}{3}$ | $\frac{2}{3}$ | 0 |  |  |
| 0 | $y_{7}$ | $\frac{95}{3}$ | $\frac{4}{3}$ | 0 | 0 | $\frac{10}{3}$ | $-\frac{2}{3}$ | $\frac{17}{3}$ | 1 |  |
| $z=$ | $\frac{40}{3}$ | $z_{\mathrm{j}}-c_{\mathrm{j}}$ | $\frac{2}{3}$ | 0 | 0 | $\frac{2}{3}$ | $\frac{5}{3}$ | $\frac{7}{3}$ | 0 |  |

Since all $z_{j}-c_{j} \geq 0$, this optimality conditions are satisfied. The optimal solution is $x_{1}=0, x=\frac{8}{3}, x_{3}=\frac{8}{3}, z_{\text {max }}=\frac{40}{3}$.

When $c$ is changed from [ $\left.\begin{array}{lll}1 & 4 & -1\end{array}\right]$ to $\left[\begin{array}{lll}1 & 2 & 3\end{array}\right]$ this modified form of the optimal table becomes

|  | $c_{j}$ |  | 1 | 2 | 3 | 4 | 0 | 0 | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |
| 1 | $y_{1}$ | 8 | 1 | 0 | 3 | 1 | 1 | 2 | 0 |
| 2 | $y_{2}$ | 0 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |
| 0 | $y_{7}$ | 5 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | -2 | -7 | 1 | 0 | 0 |

We see that there are negative $z_{j}-c_{j}$ viz $z_{3}-c_{3}=-2$ and $z_{4}-c_{4}=-7$. Hence the solution is not optimal. We apply simplex method to get the optimal solution.

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | $y_{1}$ | 8 | 1 | 0 | 3 | 1 | 1 | 2 | 0 |
| 2 | $y_{2}$ | 0 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |
| 0 | $y_{7}$ | 5 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |
| Min |  |  |  |  |  |  |  |  |  |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | -2 | -7 | 1 | 0 | 0 |
| 1 | $y_{1}$ | $\frac{11}{2}$ | 1 | 0 | 5 | 0 | 2 | $\frac{1}{2}$ | $-\frac{1}{2}$ |
| 2 | $y_{2}$ | 5 | 0 | 1 | -5 | 0 | -2 | 2 | 1 |
| 4 | $y_{7}$ | $\frac{5}{2}$ | 0 | 0 | -2 | 1 | -1 | $\frac{3}{2}$ | $\frac{1}{2}$ |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | -16 | 0 | -6 | $\frac{21}{2}$ | $\frac{7}{2}$ |
| 2 | $y_{3}$ | $\frac{11}{10}$ | $\frac{1}{5}$ | 0 | 1 | 0 | $\frac{2}{5}$ | $\frac{1}{10}$ | $-\frac{1}{10}$ |
| 2 | $y_{2}$ | $\frac{21}{2}$ | 1 | 1 | 0 | 0 | 0 | $\frac{5}{2}$ | $\frac{1}{2}$ |
| 4 | $y_{7}$ | $\frac{47}{10}$ | $\frac{2}{5}$ | 0 | 0 | 1 | $\frac{1}{5}$ | $\frac{17}{10}$ | $\frac{3}{10}$ |
| $z=\frac{431}{10}$ | $z_{j}-c_{j}$ | $\frac{16}{5}$ | 0 | 0 | 0 | $\frac{2}{5}$ | $\frac{121}{10}$ | $\frac{19}{10}$ |  |

Since all $z_{j}-c_{j} \geq$ we have obtained this optimal table. The optimal solution is

$$
\begin{aligned}
x_{1} & =0 \\
x_{2} & =\frac{21}{12}, \\
x_{3} & =11 \\
x_{4} & =\frac{47}{10} \\
\text { and } z_{\max } & =\frac{431}{10}=43 \frac{1}{10} .
\end{aligned}
$$

..... initial table the basis is $\mathrm{I}=\left[\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right]=\left[\begin{array}{lll}a_{5} & a_{6} & a_{7}\end{array}\right]$ and so inverse of the basis of the final table is given by $B^{-1}=\left[\begin{array}{ccc}1 & 2 & 0 \\ 0 & -1 & 0 \\ -2 & 3 & 1\end{array}\right]$.

The new solution when $b$ is changed from $[8021]^{\mathrm{T}}$ to $[3-24]^{\mathrm{T}}$ is given by

$$
\begin{aligned}
x_{\mathrm{B}}^{*}=\mathrm{B}^{-1} b^{*} & =\left[\begin{array}{ccc}
1 & 2 & 0 \\
0 & -1 & 0 \\
-2 & 3 & 1
\end{array}\right]\left[\begin{array}{c}
3 \\
-2 \\
4
\end{array}\right] \\
& =\left[\begin{array}{c}
-1 \\
2 \\
-8
\end{array}\right]=\left[\begin{array}{l}
x_{1} \\
x_{2} \\
x_{7}
\end{array}\right]
\end{aligned}
$$

This solution is not feasible but optimal. Hence to get the optimal solution we are to apply dual simplex method. The following are the modified optimal table and tables obtained by dual simplex method.

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | $y_{1}$ | -1 | 1 | 0 | 3 | 1 | 1 | 2 | 0 |
| 1 | $y_{2}$ | 2 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |
| 0 | $y_{7}$ | -8 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | 1 | 1 | 2 | 3 | 0 |
| $\frac{z_{j}-c_{1}}{y_{3 j}}: y_{3 j}<0$ |  |  | $-\frac{1}{4}$ |  | -1 |  |  |  |  |
| 2 | $y_{1}$ | -7 | 1 | 0 | 0 | $\frac{5}{2}$ | $-\frac{1}{2}$ | $\frac{17}{4}$ | $\frac{3}{4}$ |
| 1 | $y_{2}$ | 4 | 0 | 1 | 0 | $-\frac{5}{2}$ | $\frac{1}{2}$ | $-\frac{7}{4}$ | $-\frac{1}{4}$ |
| 4 | $y_{3}$ | 2 | 0 | 0 | 1 | $-\frac{1}{2}$ | $\frac{1}{2}$ | $-\frac{3}{4}$ | $-\frac{1}{4}$ |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | 1 | $\frac{3}{2}$ | $\frac{3}{2}$ | $\frac{15}{4}$ | $\frac{1}{4}$ |
| $\frac{z_{j}-c_{j}}{y_{j j}}: y_{1 j}<0$ |  |  | 0 |  | -3 |  |  |  |  |
| 0 | $y_{5}$ | -14 | -2 | 0 | 0 | -5 | 1 | $\frac{17}{4}$ | $-\frac{3}{2}$ |
| 1 | $y_{2}$ | -3 | 1 | 1 | 0 | 0 | 0 | $\frac{5}{2}$ | $-\frac{1}{2}$ |
| 4 | $y_{3}$ | -5 | 1 | 0 | 1 | 2 | 0 | $\frac{7}{2}$ | $\frac{1}{2 a}$ |

We note here that $x_{\mathrm{B}_{3}}=-5<0$ but all $y_{3 j} \geq 0$. Hence this changed problem has no feasible solution.

When $b_{2}$ changed to 11 , the new solution is given by

$$
x_{\mathrm{B}}^{*}=\left[\begin{array}{l}
x_{1} \\
x_{2} \\
x_{7}
\end{array}\right]=\mathrm{B}^{-1} b^{*}=\left[\begin{array}{ccc}
1 & 2 & 0 \\
0 & -1 & 0 \\
-2 & 3 & 1
\end{array}\right]\left[\begin{array}{c}
8 \\
11 \\
21
\end{array}\right]=\left[\begin{array}{c}
30 \\
-11 \\
38
\end{array}\right]
$$

Since $x_{B_{2}=-1<0}$, the solution is not feasible but optimal. So to get optimal solution we are to apply dual simplex method in the modified optimal table. The dual simplex tables are as follows :

|  | $c_{j}$ | 2 | 1 | 4 | -1 | 0 | 0 | 0 |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{B}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |
| 2 | $y_{1}$ | 30 | 1 | 0 | 3 | 1 | 1 | 2 | 0 |
| 1 | $y_{2}$ | -11 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |
| 0 | $y_{7}$ | 38 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | 1 | 1 | 2 | 3 | 0 |
| $\frac{z_{j}-c_{j}}{y_{2 j}}: y_{2 j}<0$ |  |  | -1 | $-\frac{1}{2}$ |  | 3 |  |  |  |
| 2 | $y_{1}$ | $\frac{49}{2}$ | 1 | $\frac{1}{2}$ | $\frac{5}{2}$ | 0 | 1 | $\frac{3}{2}$ | 0 |
| -1 | $y_{4}$ | $\frac{11}{2}$ | 0 | $-\frac{1}{2}$ | $\frac{1}{2}$ | 1 | 0 | $\frac{1}{2}$ | 0 |
| 0 | $y_{7}$ | 27 | 0 | 1 | -5 | 0 | -2 | 2 | 1 |
| $z=\frac{87}{2}$ |  | $z_{j}-c_{j}$ | 0 | $\frac{1}{2}$ | $\frac{1}{2}$ | 0 | 2 | $\frac{5}{2}$ | 0 |

In this tabl all $x_{\mathrm{B}_{i}}>0$ and all $z_{j}-c_{j} \geq 0$. So we have reached to the optimal table. The optimal solution is $x_{\mathrm{i}}=\frac{49}{2}, x_{2}=0, x_{3}=0, x_{4}=\frac{11}{2}$ and $z_{\max }=\frac{87}{2}$.
(e) When $c_{1}=1$ is replaced by $c_{1}^{*}$ the modified form of the optimal table is given by

|  | $c_{i}^{*}$ | 1 | 4 | 4 | -1 | 0 | 0 | 0 |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{B}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ |
| $c_{i}^{*}$ | $y_{1}$ | 8 | 1 | 0 | 3 | 1 | 1 | 2 | 0 |
| 1 | $y_{2}$ | 0 | 0 | 1 | -1 | -2 | 0 | -1 | 0 |
| 0 | $y_{7}$ | 5 | 0 | 0 | -4 | 2 | -2 | 3 | 1 |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | $3 c_{i}^{*}-5$ | $c_{i}^{*}-1$ | $c_{i}^{*}$ | $2 c_{i}^{*}-1$ | 0 |

This table remains as optimal tablie if all $z_{j}-c_{j} \geq 0$
i.e. if $3 c_{i}^{*}-5 \geq$ and $c_{i}^{*}-1 \geq 0$ and $c_{1}^{*} \geq 0$ and $2 c_{i}^{*}-1 \geq 0$
i.e. if $c_{1}^{*} \geq \frac{5}{3}$ and $c_{1}^{*} \geq$ and $c_{1}^{*}$ and $c_{1}^{*} \geq \frac{1}{2}$
i.e. if $c_{1}^{*} \geq \frac{5}{3}$.
(e) Alternative method using formula :

Since $c_{1} \in c_{\mathrm{B}}$, the range of $\Delta c_{1}$ for which the optimality of the solution is maintained is given by
$\max \left\{\frac{z_{j}-c_{j}}{y_{1 j}}: y_{1 j}>0\right\} \leq \Delta c_{1} \leq \min \left\{\frac{z_{j}-c_{j}}{y_{1 j}}: y_{1 j}<0\right\}$
i.e. $\max \left\{-\frac{z_{3}-c_{3}}{y_{13}},-\frac{z_{4}-c_{4}}{y_{14}},-\frac{z_{5}-c_{5}}{y_{15}},-\frac{z_{5}-c_{5}}{y_{16}}\right\} \leq \Delta c_{1}$
i.e. $\max \left\{-\frac{1}{3},-\frac{1}{1},-\frac{2}{1},-\frac{3}{2}\right\} \leq \Delta c_{1}$
i.e. $-\frac{1}{3} \leq \Delta c_{1}<\propto$
$\therefore c_{1}-\frac{1}{3} \leq c_{1}+\Delta c_{1}<c_{1}+\infty$
or, $2-\frac{1}{3} \leq c_{1}^{*}<\infty$
or, $\frac{5}{3} \leq c_{1}^{*}<\infty$
$\therefore$ If $c_{1}^{*} \geq \frac{5}{3}$ the optimal solution remain optimal.m $\alpha$

### 4.6 Addition Of A Single Variable :

Let the optimal solution of the given LPP
Maximize $z=c x$

$$
\text { subject to } A x=b, x \geq 0
$$

be known. Let $x_{n+1}$ be added with it and the coefficient vector associated with $x_{n+1}$ be $a_{n+1}$ and the cost coefficient for $x_{n+1}$ be $c_{n+1}$.

Since $b$ is not changed the old optimal solution will be feasible solution of the new LPP but it may not be optimal. Let $B$ be the optimal basis and $C_{B}$ be the associated cost vector of the old LPP. Then they are also the same for the new LPP. It is optimum for the new LPP if $z_{n+1}-\mathrm{C}_{n+1} \geq 0$.

In case $z_{n+1}-\mathrm{C}_{n+1}<0, x_{n+1}$ will enter the solution and simplex method is to be applied to the old optimal table added with $(n+1)$ th column as $y_{n+1}=\mathrm{B}^{-1} a_{n+1}$.

### 4.7 Illustrative Example :

Example 4.7.1: Consider the LPP

$$
\begin{aligned}
\text { Maximize } & z=x_{1}+2 x_{2}+x_{3} \\
\text { subject to } & 2 x_{1}+x_{2}-x_{3} \leq 2 \\
& 2 x_{1}-x_{2}+5 x_{3} \leq 6 \\
& 4 x_{1}+x_{2}+x_{3} \leq 6 \\
& x_{1}, x_{2}, x_{3} \geq 0
\end{aligned}
$$

Let a new variable $x_{3}^{\prime} \geq 0$ be introduced with cost (i) 3 (ii) 5 and $a_{3}^{\prime}=[2-1-$ 4]. Discuss the effect.

The solution of the LPP is obtained by simplex method. The following are the tables.


The inverse of the basis in the optimal table is

$$
\mathrm{B}^{-1}=\left[\begin{array}{ccc}
\frac{5}{4} & \frac{1}{4} & 0 \\
\frac{1}{4} & \frac{1}{4} & 0 \\
-\frac{3}{2} & -\frac{1}{2} & 2
\end{array}\right]
$$

The added column for $x_{3}^{\prime}$ is $a_{3}^{\prime}=\left[\begin{array}{c}2 \\ -1 \\ 4\end{array}\right]$

The corresponding column in the final table is given by

$$
y_{3}^{\prime}=\mathrm{B}^{-1} a_{3}^{\prime}=\left[\begin{array}{ccc}
\frac{5}{4} & \frac{1}{4} & 0 \\
\frac{1}{4} & \frac{1}{4} & 0 \\
-\frac{3}{2} & -\frac{1}{2} & 1
\end{array}\right]=\left[\begin{array}{c}
2 \\
-1 \\
4
\end{array}\right]=\left[\begin{array}{c}
\frac{9}{4} \\
\frac{1}{4} \\
\frac{6}{4}
\end{array}\right]
$$

(i) When $c_{3}^{\prime}=3$, we have

$$
z_{3}^{\prime}-c_{3}^{\prime}-c_{\mathrm{B}} y_{3}^{\prime}-c_{3}^{\prime}=\left[\begin{array}{lll}
2 & 1 & 0
\end{array}\right]\left[\begin{array}{l}
\frac{9}{4} \\
\frac{1}{4} \\
\frac{6}{4}
\end{array}\right]-3=\frac{19}{4}-3=\frac{7}{4}>0
$$

$\therefore$ The optimatity condition is satisfied for the changed problem also. The optimal solution is $x_{1}=0, x_{2}=4, x_{3}=2$.
(ii) When $z_{3}^{\prime}-c_{3}^{\prime}=c_{\mathrm{B}} y_{3}^{\prime}-c_{3}^{\prime}=\left[\begin{array}{lll}2 & 1 & 0\end{array}\right]\left[\begin{array}{l}\frac{9}{4} \\ \frac{1}{4} \\ \frac{6}{4}\end{array}\right]=\frac{19}{4}-5=-\frac{1}{4}<0$
$\therefore$ Optimatity condition is not satisfied here.
We shall modify the optimal table of the old problem with added column $y_{3}^{\prime}=\left[\begin{array}{c}\frac{9}{4} \\ \frac{1}{4} \\ \frac{6}{4}\end{array}\right]$
and $z_{3}^{\prime}-c_{3}^{\prime}=-\frac{1}{4}$ and $c_{3}^{\prime}=5$. Then to get the optimal solution we are to apply simplex method. The tables obtained are as follows.

|  |  | $c_{j}$ | 1 | 2 | 1 | 5 | 0 | 0 | 0 | $\min _{\text {ratio }}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\text {B }}$ | $y_{B}$ | $x_{B}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{3}^{\prime}$ | $y_{4}$ | $y_{s}$ | $y_{6}$ |  |
| 2 | $y_{2}$ | 4 | 3 | 1 | 0 | $\frac{9}{4}$ | $\frac{5}{4}$ | $\frac{1}{4}$ | 0 | $\begin{gathered} 16 \\ 9 \end{gathered}$ |
| 1 | $y_{3}$ | 2 | 1 | 0 | 1 | $\frac{1}{4}$ | $\frac{1}{4}$ | $\frac{1}{4}$ | 0 | 8 |
| 0 | $y_{6}$ | 0 | 0 | 0 | 0 | $\left[\begin{array}{l}6 \\ 4\end{array}\right]$ | $-\frac{3}{2}$ | $-\frac{1}{2}$ | 1 | $0 \rightarrow$ |
|  |  | $z_{j}-c_{j}$ | 6 | 0 | 0 | $-\frac{1}{4}$ | $\frac{11}{4}$ | $\frac{3}{4}$ | 0 |  |
| 2 | $y_{2}$ | 4 | 3 | 1 | 0 | 0 | $\frac{7}{2}$ | 1 | $-\frac{3}{2}$ |  |
| 1 | $y_{3}$ | 2 | 1 | 0 | 1 | 0 | $\frac{1}{2}$ | $\frac{1}{3}$ | $-\frac{1}{6}$ |  |
|  | $y_{3}^{\prime}$ | 0 | 0 | 0 | 0 | 1 | -1 | $-\frac{1}{3}$ | $\frac{2}{3}$ |  |
| $z=10$ |  | $z_{j}-c_{j}$ | 6 | 0 | 0 | 0 | $\frac{5}{2}$ | $\frac{2}{3}$ | $\frac{2}{3}$ |  |

In this table all $x_{\mathrm{B}_{i}} \geq 0$ and all $z_{j}-c_{j} \geq$. So the optimatily conditions are satisfied. The optimal solution is given by $x_{1}=0, x_{2}=4, x_{3}-2, x_{3}^{\prime}=0$ and $z_{\text {max }}=10$.

### 4.8 Deletion of A Variable :

From a LPP if we delete a variable them two cases any arise.
Case 1. If this variable deleted is non basic then the feasibility and optimality conditions are not affected. So the optimal solution of the old problem is the optimal solution of the new problem.

Case 2. If the variable deleted is basic then the conditions of optimality may be affected and so a new solution is to be obtained. For this new optimal solution, we are assign a cost - M corresopnding to the basic variable to be deleted and apply simplex method after modifying the old optimal table.

### 4.9 Illustrative Example :

## Example 4.9.1 For the LPP

$$
\begin{array}{lr}
\text { Maximize } \quad \mathrm{z}= & x_{1}+2 x_{2}+x_{3} \\
\text { subject to } \quad & 2 x_{1}+x_{2}-x_{3} \leq 2 \\
& 2 x_{1}-x_{2}+5 x_{3} \leq 6 \\
& 4 x_{1}+x_{2}+x_{3} \leq 6 \\
& x_{1}, x_{2}, x_{3} \geq 0
\end{array}
$$

the optimal table is

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | $y_{2}$ | 4 | 3 | 1 | 0 | $\frac{5}{4}$ | $\frac{1}{4}$ | 0 |
| 1 | $y_{3}$ | 2 | 1 | 0 | 1 | $\frac{1}{4}$ | $\frac{1}{4}$ | 0 |
| 0 | $y_{6}$ | 0 | 0 | 0 | 0 | $-\frac{3}{2}$ | $-\frac{1}{2}$ | 1 |
| $z=10$ |  | $z_{\mathrm{j}}-c_{j}$ | 6 | 0 | 0 | $\frac{1}{4}$ | $\frac{3}{4}$ | 0 |

Discuss the effect of deletion of the variable (i) $x_{1}$ (ii) $x_{2}$ (iii) $x_{3}$.
(i) From the optimal table we see that $x_{1}$ is a deleted the optimal solution remains unaffected. Hence old optimal solution is also the new optimal solution is

$$
x_{1}=0, x_{2}=4, x_{3}=2 \& z_{\max }=10
$$

(ii) From the optimal table we see that $x_{2}$ is a basic variable. Hence we make a new starting table by changing $c_{2}=2$ by $-M$, where $M$ is a big positive number. As M is very large the optimatity conditions are not affected and once it goes out from the basis it never reappears in the basis in the simplex method.

The modified table is

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | min ratio |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| -M | $y_{2}$ | 4 | 3 | 1 | 0 | $\frac{5}{4}$ | $\frac{1}{4}$ | 0 | $\frac{4}{3}$ |
| 1 | $y_{3}$ | 2 | 1 | 0 | 1 | $\frac{1}{4}$ | $\frac{1}{4}$ | 0 | 2 |
| 0 | $y_{6}$ | 0 | 0 | 0 | 0 | $-\frac{3}{2}$ | $-\frac{1}{2}$ | 1 | $\ldots$ |
| 1 | $y_{1}$ | $\frac{4}{3}$ | 1 | $\frac{1}{3}$ | 0 | $\frac{5}{12}$ | $\frac{1}{12}$ | 0 |  |
| 1 | $y_{3}$ | $\frac{2}{3}$ | 0 | $-\frac{1}{3}$ | 1 | $-\frac{1}{6}$ | $\frac{1}{6}$ | 0 |  |
| 0 | $y_{6}$ | 0 | 0 | 0 | 0 | $-\frac{3}{2}$ | $-\frac{1}{2}$ | 1 |  |
| $z=2$ |  | $z j-c j$ | 0 | M | 0 | $\frac{1}{4}$ | $\frac{1}{6}$ | 0 |  |

In this table all $z_{j}-c_{j} \geq$ and all $x_{\mathrm{B}_{i}} \geq 0$, so we have reached to optimal table. The optimal solution is

$$
x_{1}=\frac{4}{3}, x_{2}=0, x_{3}=\frac{2}{3}
$$

$$
\text { and } z_{\max }=2
$$

(iii) From the optimal table we see that $x_{3}$ is a basic variable. Hence we make a new starting table by changing $c_{1}=1$ by $-M$, where $M$ is a big positive number. As $\mathbf{M}$ is very large the optimality conditions are satisfied. Also once it goes out from the basis it never reappears in the basis in the simplex method. The modified table and other simplex tables are as follows :

$$
\begin{array}{lllllll}
c_{j} & 1 & -\mathrm{M} & 1 & 0 & 0 & 0
\end{array}
$$



The optimal table is obtained and this optimal solution is $x_{1}=0, x_{2},=2, x_{3}=0$ and $z_{\max }=4 . \alpha$

### 4.10 Addition Of A New Constraint :

Addition of a new constraint may or may not affect the current optimal solution. Two cases will arise.
(i) If the added constraint is satisfied by the old optimal solution then teh old optimal solution is also the new optimal solution.
(ii) If the added constraiant is not satisfied by the old optimal solution, then this old optimal solution becomes an infeasible solution for the new problem.
To obtain the optimal solution for the changed problem we are first to modify the fimal table and then apply dual simplex method.

The following three situations will arise depending on the nature of the solution to the original LPP.

If original LPP has an optimal solution then the modified LPP may have an optimal solution or it will give no F.S.

If the original LPP has unbounded solution then the modified LPP may have optimal solution or it will have no F.S. or it will have unbounded solution.

If the original LPP has no F.S, then the modified LPP will have also no F.S. $\alpha$

### 4.11 Illustrative Examples :

Example 4.11.1 Let us consider the final table of a LPP

| $c_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ | $y_{7}$ | $y_{8}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 2 | $y_{1}$ | 3 | 1 | 0 | 0 | -1 | 0 | 5 | 2 | -1 |
| 4 | $y_{2}$ | 1 | 0 | 1 | 0 | 2 | 1 | -1 | 0 | -5 |
| 1 | $y_{3}$ | 7 | 0 | 0 | 1 | 1 | -2 | 5 | --3 | 2 |
|  |  | $z_{j}-c_{j}$ | 0 | 0 | 0 | -1 | 0 | 2 | 1 | 2 |

where $y_{6}, y_{7}$ and $y_{8}$ are slack variables.
If the constraint
(i) $2 x^{1}+3 x^{2}-x^{3}+2 x^{4}+4 x^{5} \leq 5$
(ii) $2 x^{1}+3 x^{2}-x^{3}+2 x^{4}+4 x^{5} \leq 1$
is added then find the solution of the changed LPP.

## Solution :

From the final table we see that the optimal solution of the old LPP is

$$
x_{1}=3, x_{2}=1, x_{3}=7, x_{4}=0, x_{5}=0, x_{6}=0, x_{7}=0, x_{8}=0
$$

(i) The added constraint is

$$
2 x_{1}+3 x_{2}-x_{3}+2 x_{4}+4 x_{5} \leq 5
$$

Putting $x_{1}=3, x_{2}=1, x_{3}=7, x_{4}=0, x_{5}=0, x_{6}=0, x_{7}=0, x_{8}=0$ in this constraint we have

$$
\begin{aligned}
& 2.3+3.1-7+2.0+4.0 \leq 5 \\
& \text { or, } 6+3-7 \leq 5 \\
& \text { or, } 2 \leq 5
\end{aligned}
$$

This is true. So the solution satisfies the added constraint. Hence the old optimal solution is also optimal solution to the new problem.

The added constraint is

$$
2 x_{1}+3 x_{2}-x_{3}+2 x_{4}+4 x_{5} \leq 1
$$

Putting $x_{1}=3, x_{2}=1, x_{3}=7, x_{4}=0, x_{5}=0, x_{6}=0, x_{7}=0, x_{8}=0$ in this constraint we get

$$
\begin{aligned}
& 2.3+3.1-7+2.0+4.0 \leq 1 \\
& \text { or, } 6+3-2 \leq 1 \\
& \text { or, } 2 \leq 1
\end{aligned}
$$

This is not time i.e. the optimal solution to the old problem does not satisfy the added constraint. To get the solution of the new LPP we introduces the new constraint with a new slack variable in the optimal table of the old problem. We then modify this table to have a unit basis and then apply dual simplex method to it. The following are the tables.


The second table is obtained by the operation $R_{4}^{1}=R_{4}-2 R_{1}-3 R_{2}+R_{3}$. The third table is obtained by using dual simplex method to the second table and is the the final table. The optimal solution is $x_{1}=3, x_{2}=0, x_{3}=9, x_{4}=0, x_{5}=1$.

### 4.12 Summary :

The usefulness of post-optimality analysis is discussed. Then onle by one the different situations viz discrete changes in the cost vector and requirement vector,
addition and deletion of a single variable, and addition of a new constraint are discussed. Each situation is illustrated by examples.

### 4.13 Self Assessment Questions :

1. For the LPP

Maximize $\quad z=15 x_{1}+45 x_{2}$
subject to $\quad 5 x_{1}+2 x_{2} \leq 162$

$$
\begin{gathered}
x_{1}+16 x_{2} \leq 240 \\
x_{2} \leq 50 \\
x_{1}, x_{2} \geq 0
\end{gathered}
$$

find the optimal solution. Find the range of each cost coefficient (changed one at a time) to give same optimal solution.
[ Ans : $x_{1}=352 / 13, x_{2}=173 / 13, z_{\max }=1005$ ]
2. Find how much the 7 in the first constraint of the problem

Minimize $\quad z=x_{1}-3 x_{2}+2 x_{3}$
subject to $\quad 3 x_{1}-x_{2}+2 x_{3} \leq 7$
$-2 x_{1}+4 x_{2} \leq 12$
$-4 x_{1}+3 x_{2}+8 x_{3} \leq 10$

$$
x_{1}, x_{2}, x_{3} \geq 0
$$

be changed before the basis of the optimal table would change.
3. Find the optimal solution of the LPP and the separate ranges of variations of $b_{2}$ and $b_{3}$ consistent with the optimatity of the solution

Minimize

$$
\begin{gathered}
z=-x_{1}+2 x_{2}-x_{3} \\
3 x_{1}+x_{2}-x_{3} \leq 10 \\
-x_{1}+4 x_{2}+x_{3} \geq 6 \\
x_{2}+x_{3} \leq 4 \\
x_{1}, x_{2}, x_{3} \geq 0 .
\end{gathered}
$$

Determine also this efficient discrete changes in the components of the cost vector which correspond to the basic variables.

$$
\text { [Ans: } x_{1}=0, x_{2}=4, x_{3}=0 ; \Delta b_{2} \leq 10,-5 / 2 \leq \Delta b_{3} \leq 6,-2 \leq \Delta c_{2} \text { ] }
$$

4. Following is the optimal table for an LPP

|  |  | $c_{j}$ | 2 | 1 | 1 | 2 | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $c_{\mathrm{B}}$ | B | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ |
| 2 | $a_{1}$ | 3 | 1 | 0 | -1 | 3 | 2 |
| 1 | $a_{2}$ | 4 | 0 | 1 | 4 | -1 | -2 |

(i) Find the limitations of this values of $c_{3}, c_{4}, c_{5}$ (taking one at a time) for which the current solution will remain optimal.
(ii) Find the optimal solution to the problem, if $c_{3}$ is changed to 3 .
(iii) Find the limitations of the values of $c_{1}$ for which the current solution remains optimal.
(iv) Find the optimal solution to this problem, if $c_{1}$ is changed to 5 .
[ Ans: (i) $-\alpha<c_{3} \leq 2,-\alpha<c_{4} \leq 5,-\alpha<c_{5} \leq 2$
(ii) $x_{1}=4, x_{3}=1 ; x_{2}=0, x_{4}=0$
(iii) $1 \leq c_{1} \leq 3$
(iv) $x_{1}=13 / 4, x_{2}=0, x_{3}=1, x_{4}=0$ ]
5. Find the optimal solution of the IPP

$$
\begin{array}{lc}
\text { Maximize } & z=4 x_{1}+3 x_{2} \\
\text { subject to } & x_{1}+x_{2} \leq 5 \\
& 3 x_{1}+x_{2} \leq 7 \\
& x_{1}+2 x_{2} \leq 10 \\
& x_{1}, x_{2} \geq 0
\end{array}
$$

Show how to find the optimal solution of the problem, if
(i) the first component of the original requirement vector be increased by one unit and the third component be decreased by one unit.
(ii) the second component of the original requirement vector be decreased by two units.
(Ans: (i) $x_{1}=1, x_{2}=4, z_{\text {max }}=16$
(ii) $\left.x_{1}=0, x_{2}=5, z_{\text {max }}=15\right]$

## Unit 5 Q Quadratic Programming Problem

## Structure

5.1 Introduction
5.2 Kuhn-Tucker Conditions for Quadratic Programming Problem
5.3 Wolfe's Modified Simplex Method
5.4 Beale's Method .
5.5 Summary
5.6 Self Assessment Questions

### 5.1. Introduction :

Quadratic programiing problem is the most well behaved nonlinear programming problem. Quadratic programming deals with non-linear programming problem of maximizing (or minimizing) quadratic objective function subject to a set of linear inequality constraints. The solution of this problem is based on the Kuhn-Tucker conditions. The quadratic objective function to be optimized is taken as strictly convex for minimization and strictly concave for maximization. As the solution space is always convex, the optimal the solution obtained is global is nature.

Definition 5.1.1 : Let $x^{\mathrm{T}}$ and $\mathrm{C} \in \mathrm{R}^{\mathrm{p}}$ and Q be a symmetric $n \times n$ real matrix then, the problem quadratic pogamming problem is

$$
\begin{aligned}
& \text { Maximize (or minimize) } f(x)=c x+\frac{1}{2} x^{\top} Q x \\
& \text { subject to, Ax: } \leq b \\
& x \geq 0 \\
& \text { where } \quad x=\left[x_{1}, x_{2}, \ldots, x_{n}\right]^{\mathrm{T}} \\
& \left.\qquad \begin{array}{rl}
c & =\left[c_{1}, c_{2}, \ldots, c_{n}\right] \\
b & =\left[b_{1}, b_{2}, \ldots\right. \\
b_{m}
\end{array}\right]^{\mathrm{T}}
\end{aligned}
$$

$$
\mathrm{A}=\left[\begin{array}{cccc}
a_{11} & a_{12} & \ldots . & a_{1 n} \\
a_{21} & a_{22} & \ldots . & a_{2 n} \\
\ldots & \ldots & \ldots . & \ldots . \\
\ldots . & \ldots . & \ldots & \ldots . \\
a_{m 1} & a_{m 2} & \ldots . & a_{n m}
\end{array}\right] \text { and } \mathrm{Q}=\left[\begin{array}{cccc}
c_{11} & c_{12} & \ldots . & c_{1 n} \\
c_{21} & c_{22} & \ldots . & c_{2 n} \\
\ldots . & \ldots . & \ldots & \ldots \\
\ldots . & \ldots . & \ldots . & \ldots \\
c_{n 1} & c_{n 2} & \ldots & c_{n n}
\end{array}\right]
$$

The function $x^{\mathrm{T}} \mathrm{Q} x$ defines a quadratic form when Q is a symmetric matrix.
The quadratic form $x^{\mathrm{T}} \mathrm{Q} x$ is said to be positive-definite if $x^{\mathrm{T}} \mathrm{Q} x \geq$ for all $\boldsymbol{x} \neq 0$.

The quadratic form $x^{\mathrm{T}} \mathrm{Q} x$ is said to be positive semi definite if $x^{\mathrm{T}} \mathrm{Q} x \geq$ for at one $x \neq 0$.

The quadratic form $x^{\mathrm{T}} \mathrm{Q} x$ is said to be negative definite and negative semidefinite if $-x^{\top} \mathrm{Q} x$ is positive definite and positive semi-definite respectively.

In quadratic programming problem $x^{T} Q x$ is assmed to be negative definite in the maximization case, and positive definite in the minimization case. These means that $f(x)=c x+\frac{1}{2} x^{T} Q x$ is assumed to be strictly convex function for minimmzation case and strictly concave fomaximization case.

As the constraints are always assumed to be linear, the solution space of a quadratic programming problem is always convex.

Thus the solution obtained using Kuhn-Tucker conditions given global optimum of the quadatic programming problem.

### 5.2 Kuhn-Tucker Conditions for Quadratic Programming Problem:

Let the quadratic programming problem be
Maximize $\mathrm{f}(\mathrm{x})=\sum_{j=1}^{n} c_{j} x_{j} \frac{1}{2} \sum_{j=1}^{n} \sum_{j=1}^{n} c_{j k} x_{j} x_{k}$
subject to the constraints

$$
\sum_{j=1}^{n} a_{i j} x_{j} \leq b_{j}, i=1,2, \ldots \ldots, m
$$

$$
\text { and } x_{j} \geq 0, j=1,2, \ldots \ldots, n
$$

where $c_{j k}=c_{k j}$ for all $j$ and $k$.
Introducing slack variables $q_{i}^{2}$ and $r_{j}^{2}$ the problem reduces to

$$
\begin{aligned}
& \text { Maximize } f=\sum_{j=1}^{n} c_{j} x_{j}+\frac{1}{2} \sum_{j=1}^{n} \sum_{k=1}^{n} c_{j k} x_{j} x_{k} \\
& \text { subject to } \sum_{j=1}^{n} a_{i j} x_{j}-b_{i}+q_{i}^{2}=0, i=1,2, \ldots . . m \\
& \quad-x_{j}+r_{j}^{2}=0, j=1,2, \ldots \ldots . .
\end{aligned}
$$

The Lagangian function is given by

$$
\begin{aligned}
& L\left(x_{1}, x_{2}, \ldots, x_{n}, q_{1}, q_{2}, \ldots, q_{m}, r_{1}, r_{2}, \ldots, r_{n}, \lambda_{1}, \lambda_{2}, \ldots, \lambda_{m}, \mu_{1}, \mu_{2}, \ldots, \mu_{n}\right) \\
& =\sum_{j=1}^{n} c_{j} x_{j}+\frac{1}{2} \sum_{j=1}^{n} \sum_{j=1}^{n} c_{j k} x_{j} x_{k}-\sum_{i=1}^{m} \lambda_{j}\left(\sum_{j=1}^{n} a_{i j} x_{j}-b_{i}+q_{i}^{2}\right)-\sum_{j=1}^{n} \mu_{j}\left(-x_{j}+r_{j}^{2}\right)
\end{aligned}
$$

The Kuhn-Tucher conditions are given by

$$
\begin{aligned}
& \frac{\partial f}{\partial x_{j}}-\sum_{i=1}^{m} \lambda_{i} a_{i j}-\mu_{j}(-1)=0, \quad j=1,2, \ldots \ldots, n \\
& \lambda_{i}\left(\sum_{j=1}^{n} a_{i j} x_{j}-b_{j}\right)=0, \quad i=1,2, \ldots \ldots, m \\
& \mu_{j} x_{j}=0, \quad j=1,2, \ldots \ldots, n \\
& \sum_{j=1}^{n} a_{i j} x_{j} \leq b_{j}, \quad i=1,2, \ldots \ldots, m \\
& \quad x_{j} \geq 0, \quad j=1,2, \ldots ., m \\
& \lambda_{i} \geq 0, \quad i=1,2, \ldots, m \\
& \mu_{j} \geq 0, \quad j=1,2, \ldots ., n
\end{aligned}
$$

Letting $q_{i}^{2}=s_{i} \geq 0$ these equations becomes

$$
\left.\begin{array}{l}
\left.\begin{array}{c}
c_{j}+\sum_{k=1}^{n} c_{j k} x_{k}-\sum_{i=1}^{m} \lambda_{i} a_{i j}+\mu_{j}=0 \\
\sum_{j=1}^{n} a_{i j} x_{j}-b_{i}+s_{i}=0, i=1,2, \ldots, m
\end{array}\right\} \\
\left.\begin{array}{l}
\lambda_{1} s_{i}=0, i=1,2, \ldots, m \\
\mu_{j} x_{j}=0, j=1,2, \ldots, n
\end{array}\right\} \\
\left.\begin{array}{l}
\lambda_{i} \geq 0, i=1,2, \ldots, m \\
\mu_{j} \geq 0, j=1,2, \ldots, n \\
x_{j} \geq 0, j=1,2, \ldots, n \\
s_{i} \geq 0, i=1,2, \ldots, n
\end{array}\right\}
\end{array}\right\}
$$

(1) is a system of $m+n$ linear equations in $x_{j}, \lambda_{i}, \mu_{i}$ and $s_{i}$.

The solution of these system which will satisfy also (2) and (3) is the required optimal solution of the quadrative programming problem.

### 5.3 Wolfe's Modified Simplex Method :

To solve the system (1) satisfying the conditions (2) and (3) Wolfe suggested to introduce the non-negative artificial variables $\beta_{1}, \beta_{2}, \ldots . . . \beta_{n}$ in the Kuhn-Tucker conditions (1) and to constract an objective function $z=-\beta_{1}-\beta_{2}-\ldots .-\beta_{n}$ and to consider the following LPP with complementary slackness condition.

Maximize $z=-\beta_{1}-\beta_{2}-\ldots . . .-\beta_{n}$
subject to $\sum_{k=1}^{n} c_{j k} x_{k}-\sum_{i=1}^{m} \lambda_{i} a_{i j}+\mu_{j}=-c_{j}, j=1, c, \ldots, n$

$$
\begin{aligned}
& \sum_{j=1}^{n} a_{i j} x_{j}+s_{i}=b_{i}, i=1,2, \ldots . m \\
& \lambda_{1}, s_{i}, x_{i}, \mu_{i}, \mu_{j}, \beta_{j} \geq 0, i=1,2, \ldots, m, j=1,2, \ldots . n
\end{aligned}
$$

and satisfying the complementarhy slackness conditions

$$
\begin{aligned}
& \lambda_{i} s_{i}=0, i=1,2, \ldots \ldots, m \\
& \mu_{j} x_{j}=0, j=1,2, \ldots \ldots, n
\end{aligned}
$$

The optimum solution of theis LPP gives the optimum solution of the given QPP.

Note : To maintain the condition $\lambda_{i} s_{i}=0=\mu_{j} x_{j}$ all the time we should note that if $\lambda_{i}$ is in the basic solution with positive value then $s_{i}$ can not be basic with positive value. Similarly $\mu_{j}$ and $x_{i}$ cannot be in the basic solution (i.e. positive) simultaneously.

Example 5.3.1 Using Wolfe's method solve the quadratic programming problem
Maximize $\mathrm{z}=2 x_{1}+x_{2}-x_{1}^{2}$
subject to $\quad 2 x_{1}+3 x_{2} \leq 6$

$$
2 x_{1}+x_{2} \leq 4
$$

$$
x_{1}, x_{2} \geq 0
$$

Solution : First be write all constraints with ' $Z$ ' sign to get the problem as
Maximize $\mathrm{z}=2 x_{1}+x_{2}-x_{1}^{2}$
subject to $\quad 2 x_{1}+3 x_{2} \leq 0$

$$
\begin{aligned}
2 x_{1}+x_{2} & \leq 4 \\
x_{1} \quad & \leq 0 \\
-x_{2} & \leq 0
\end{aligned}
$$

Introducing slack variable $q_{1}^{2}, q_{2}^{2}, r_{1}^{2}$ and $r_{2}^{2}$ we get
Maximize $z=2 x_{1}+x_{2}-x_{1}^{2}$
subject to

$$
\begin{aligned}
2 x_{1}+3 x_{2}+q_{1}^{2} & =6 \\
2 x_{1}+x_{2}+q_{2}^{2} & =4 \\
-x_{1}+r_{1}^{2} & =0 \\
-x_{2}+r_{2}^{2} & =0
\end{aligned}
$$

We now constuct the Largrange function
$L\left(x_{1}, x_{2}, q_{1}, q_{2}, r_{1}, r_{2}, \lambda_{1}, \lambda_{2}, \mu_{1}, \mu_{2}\right)$
$=\left(2 x_{1}+x_{2}-x_{1}^{2}\right)-\lambda_{1}\left(2 x_{1}+3 x_{2}+q_{1}^{2}-6\right)-\lambda_{2}\left(2 x_{1}+x_{2}+q_{2}^{2}-4\right)$

$$
-\mu_{1}\left(-x_{1}+r_{1}^{2}\right)-\mu_{2}\left(-x_{2}+r_{2}^{2}\right)
$$

The Kuhn-Tucker's necessay and sufficient conditions gives

$$
\begin{aligned}
& \frac{\partial L}{\partial x_{1}}=0 \text { or, } 2-2 x_{1}-2 \lambda_{1}-2 \lambda_{2}+\mu_{1}=0 \\
& \frac{\partial L}{\partial x_{2}}=0 \text { or, } 1-3 \lambda_{1}-\lambda_{2}+\mu_{2}=0 \\
& \frac{\partial L}{\partial \lambda_{1}}=0 \text { or, } 2 x_{1} 3 x_{2}+q_{1}^{2}-6=0 \\
& \frac{\partial L}{\partial \lambda_{2}}=0 \text { or, } 2 x_{1}+x_{2}+q_{2}^{2}-4=0 \\
& \lambda_{1} q_{1}^{2}-0, \lambda_{2} q_{2}^{2}=0, \mu_{1} x_{1}=0, \mu_{2} x_{2}=0 \\
& x_{1}, x_{2}, \lambda_{1}, \lambda_{2}, \mu_{1}, \mu_{2} \geq 0
\end{aligned}
$$

Taking $q_{1}^{2}=s_{1}$ and $q_{2}^{2}=s_{2}$ we get

$$
\begin{aligned}
& 2 x_{1}+2 \lambda_{1}-\mu_{1}=2 \\
& 3 \lambda_{1}+\lambda_{2}-\mu_{2}=1 \\
& 2 x_{1}+3 x_{2}+s_{i}=6 \\
& 2 x_{1}+x_{2} \quad+s_{2}=6 \\
& \lambda_{1} s_{1}=0, \lambda_{2} s_{2}=0, \mu_{1} x_{1}=0, \mu_{2} x_{2}=0 \\
& x_{1}, x_{2}, \lambda_{1}, \lambda_{2}, \mu_{1}, \mu_{2}, s_{1}, s_{2} \geq 0
\end{aligned}
$$

With necessary modification we use phase I of two phase method to solve this system Introducing artificial variables $\beta 1$ and $\beta 2$ the modified LPP become

$$
105
$$

$$
\begin{aligned}
& \text { Maximize } z^{\prime}=-\beta_{1}-\beta_{2} \\
& \text { subject to } 2 x_{1}+2 \lambda_{1}+2 \lambda_{2}-\mu_{1}+\beta_{1}+\quad=2 \\
& 3 \lambda_{1}+\lambda_{2}-\mu_{2}+\beta_{2}=1 \\
& \begin{array}{lll}
2 x_{1}+3 x_{2} & s_{1} & =6 \\
2 x_{1}+x_{2} & s_{2} & =4
\end{array} \\
& \mu_{1} x_{1}=0, \mu_{2} \dot{x}_{2}=0, \lambda_{1} s_{1}=0, \lambda_{2} s_{2}=0 \\
& x_{1}, x_{2}, \lambda_{1}, \lambda_{2}, \mu_{1}, \mu_{2}, \beta_{1}, \beta_{2}, s_{1}, s_{2} \geq 0
\end{aligned}
$$

Initiat talle of Phase-I is

|  |  | $\mathrm{C}_{\mathrm{j}}$ | 0 | 0 | 0 | 0 | 0 | 0 | -1 | -1 | 0 | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\mathrm{C}_{\mathrm{B}}$ | B.V | $X_{\mathrm{B}}$ | $x_{1}$ | $x_{2}$ | $\lambda_{1}$ | $\lambda_{2}$ | $\mu_{1}$ | $\mu_{2}$ | $\beta_{1}$ | $\beta_{2}$ | $s_{1}$ | $s_{2}$ |
| -1 | $\beta_{1}$ | 2 | 2 | 0 | 2 | 2 | -1 | 0 | 1 | 0 | 0 | 0 |
| -1 | $\beta_{2}$ | 1 | 0 | 0 | 3 | 1 | 0 | -1 | 0 | 1 | 0 | 0 |
| 0 | $s_{1}$ | 6 | 2 | 3 | 0 | 0 | 0 | 0 | 0 | 0 | 1 | 0 |
| 0 | $s_{2}$ | 4 | 2 | 1 | 0 | 0 | 0 | 0 | 0 | 0 | 0 | 1 |
| $z^{\prime}=$ | -3 |  | -2 | 0 | -5 | -3 | 1 | 1 | 0 | 0 | 0 | 0 |$\rightarrow$

Accoding to the regular pocedure $\lambda_{1}$ enters and $\beta_{2}$ leave the basis is $\lambda_{1}>0$ \& $\beta_{2}=0$. But $s_{1}=6 \quad \therefore \lambda_{1} s 1 \neq 0$.
$\therefore \lambda_{1}$ cannot enter the basis.
Next negative $z ;-e$, is associated with $\lambda_{2}$. If $\lambda_{2}$ enters the basis then $\beta_{1}$ and $\beta_{2}$ will leave the basis is $\lambda_{1}>0$.

Since $s_{2}=4$ we have $\lambda_{2} s_{2} \neq 0$. So $\lambda_{2}$ cannot enter the basis.
Next negative $z$; - ; is associated with $x_{1}$. If $x_{1}$, eneris the basis then $\beta_{1}$, leaves the basis ie. $x_{1} \geq 0$. This is accepted since $\mu_{1}=0 \& \mu_{1}, x_{1}=0$ is satisfied.

The next table is

|  |  | $\mathrm{C}_{\mathrm{j}}$ | 0 | 0 | 0 | 0 | 0 | 0 | -1 | -1 | 0 | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\mathrm{C}_{\mathrm{B}}$ | $\mathrm{B} \cdot \mathrm{V}$ | $X_{\mathrm{B}}$ | $x_{1}$ | $x_{2}$ | $\lambda_{1}$ | $\lambda_{2}$ | $\mu_{1}$ | $\mu_{2}$ | $\beta_{1}$ | $\beta_{2}$ | $s_{1}$ | $s_{2}$ |
| -1 | $x_{1}$ | 1 | 1 | 0 | 1 | 1 | $-1 / 2$ | 0 | $1 / 2$ | 0 | 0 | 0 |
| -1 | $\beta_{2}$ | 1 | 0 | 0 | 3 | 1 | 0 | -1 | 0 | 1 | 0 | 0 |
| 0 | $s_{1}$ | 4 | 0 | 3 | -2 | -2 | 1 | 0 | -1 | 0 | 1 | 0 |
| 0 | $s_{2}$ | 2 | 0 | 1 | -2 | -2 | 1 | 0 | -1 | 0 | 0 | 1 |
| $\mathrm{z}^{\prime}=$ | -3 |  | 0 | 0 | -3 | -1 | 0 | 1 | 1 | 0 | 0 | 0 |

Here $\lambda_{1}$ enters and $\beta_{1}$ leaves the basis i.e.. $\lambda_{1}>0, \beta_{2}=0$

This is not accepted since $s_{1}=4 \quad \therefore \lambda, s \neq 0$.
If $\lambda_{2}$ enters the basis then $x_{1}$ or $\beta_{2}$ leaves the basis..
This is not also accepted since $s_{2}=2$ \& so $\lambda_{2} s_{2} \neq 0$
We select $x_{2}$ to enter the basis. Then $s_{1}$ leaves the basis.
The next table is

|  |  | $\mathrm{C}_{\mathrm{j}}$ | 0 | 0 | 0 | 0 | 0 | 0 | -1 | -1 | 0 | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\mathrm{C}_{\mathrm{B}}$ | B.V | $X_{\mathrm{B}}$ | $x_{1}$ | $x_{2}$ | $\lambda_{1}$ | $\lambda_{2}$ | $\mu_{1}$ | $\mu_{2}$ | $\beta_{1}$ | $\beta_{2}$ | $s_{1}$ | $s_{2}$ |
| 0 | $x_{1}$ | 1 | 1 | 0 | 1 | 1 | $-1 / 2$ | 0 | $1 / 2$ | 0 | 0 | 0 |
| -1 | $\beta_{2}$ | 1 | 0 | 0 | 3 | 1 | 0 | -1 | 0 | 1 | 0 | 0 |
| 0 | $x_{2}$ | $4 / 3$ | 0 | 1 | $-2 / 3$ | $-2 / 3$ | $1 / 3$ | 0 | $-1 / 3$ | 0 | $1 / 3$ | 0 |
| 0 | $s_{2}$ | $2 / 3$ | 0 | 0 | $-4 / 3$ | $-4 / 3$ | $2 / 3$ | 0 | $-2 / 3$ | 0 | $-1 / 3$ | 1 |
| $\mathrm{z}^{\prime}=$ | -1 |  | 0 | 0 | -3 | -1 | 0 | 1 | 1 | 0 | 0 | 0 |

Hare $\lambda_{1}$ enters the basis and $\beta_{2}$ leaves the basis. This is acceptable since $s_{1}$ $=0 \quad \therefore \lambda_{1} s_{1}=0$.

The next, table is

|  |  | $\mathrm{C}_{\mathrm{j}}$ | 0 | 0 | 0 | 0 | 0 | 0 | -1 | -1 | 0 | 0 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\mathrm{C}_{\mathrm{B}}$ | B.V | $X_{\mathrm{B}}$ | $x_{1}$ | $x_{2}$ | $\lambda_{1}$ | $\lambda_{2}$ | $\mu_{1}$ | $\mu_{2}$ | $\beta_{1}$ | $\beta_{2}$ | $s_{1}$ | $s_{2}$ |
| 0 | $x_{1}$ | $2 / 3$ | 1 | 0 | 0 | $2 / 3$ | $-1 / 2$ | $1 / 3$ | $1 / 2$ | $-1 / 3$ | 0 | 0 |
| 0 | $\lambda_{1}$ | $1 / 3$ | 0 | 0 | 1 | $1 / 3$ | 0 | $-1 / 3$ | 0 | $1 / 3$ | 0 | 0 |
| 0 | $x_{2}$ | $14 / 9$ | 0 | 0 | 0 | $-4 / 9$ | $1 / 3$ | $-2 / 9$ | $-1 / 3$ | $2 / 9$ | $1 / 3$ | 0 |
| 0 | $s_{2}$ | $10 / 9$ | 0 | 0 | 0 | $-8 / 9$ | $2 / 3$ | $-4 / 9$ | $-2 / 3$ | $4 / 9$ | $-1 / 3$ | 1 |
| $z^{\prime}=$ | 0 |  | 0 | 0 | 0 | 0 | 0 | 0 | 1 | 1 | 0 | 0 |

In this table $\beta_{1}=0$ and $\beta_{2}=0$. So this is the final table.
The optimal solution is
$x_{1}=2 / 3, x_{2}=14 / 9, \lambda_{1}=1 / 3, \lambda_{2}=0, s_{1}=0, s_{2}=10 / 9, \mu_{1}=0, \mu_{2}=0$
The complementay stachness conditions
$\mu_{1} x_{1}=0, \mu_{2} x_{2}=0, \lambda_{1} s_{1}=0 \& \lambda_{2} s_{2}=0$ are satisfied.
$\therefore$ The optimal solution of the given quadratic programming problem is
$x_{1}=2 / 3, x_{2}=14 / 9$
and $z_{\text {man }}=2(2 / 3) \mid 14 / 9-2 / 3=22 / 9$

### 5.4 Beale's Method

Beale suggested another approach to solve quadratic programming problem (QPP)
Let the QPP be of the from

$$
\begin{aligned}
& \text { Maximize } \mathrm{f}(\mathrm{x})=c x+\frac{1}{2} x^{\mathrm{T}} \mathrm{Q} x \\
& \text { subject to } \mathrm{A} x=b, x \geq 0
\end{aligned}
$$

Where $x=\left\{x_{1}, x_{2}, \ldots \ldots . . x_{n}\right]^{\top}, \mathrm{C}=\left\{c_{1}, c_{2}, \ldots . . c_{\mathrm{n}}\right], \mathrm{A}$ is mxn matrix and Q is symmetric matrix.

In This method the variables are partitioned into basic and non-basic variables. At each iteration, the objective function is expressed in terms of te non-basic variables.

The Beale's iteative procedure of solving QPP is stated below :
Step 1. Express the constaints of the given QPP as equations by introducing slack / surplus variables to get $\mathrm{Ax}=\mathrm{b}$.

Step 2. Select arbitrarily $m$ variables as basic and the remaining $n-m$ variables as non-basic. With this partitioning, the constraint equation $\mathrm{Ax}=\mathrm{b}$ can be written as

$$
\left[\begin{array}{ll}
\mathrm{B} & \mathrm{R}
\end{array}\right]\left[\begin{array}{l}
x_{B} \\
x_{\mathrm{B}}
\end{array}\right]=\mathrm{b}
$$

or, $\mathrm{B} x_{\mathrm{B}}+\mathrm{R} x_{\mathrm{R}}=\mathrm{b}$
Where $x_{\mathrm{B}}$ and $x_{\mathrm{R}}$ denote the basic and non-basic vectors respectively. Thus we get

$$
x_{\mathrm{B}}=\mathrm{B}^{-1} \mathrm{~b}-\mathrm{B}^{-1} \mathrm{R} x_{\mathrm{R}}
$$

Step 3. Express the basic $x_{\mathrm{B}}$ in terms of non-basic $x_{\mathrm{R}}$ only, using the given and additional constraint equations, if any.

Step 4. Express the objective function $f(x)$ in terms of $x R$ only using the given and additional constraints, if. As $x_{B} \geq 0$ we have $B^{-1} R x R \leq B^{-1} b$. Thus, any component of $x R$ can increase only, until $\delta f / \delta x_{R}$ becomes zero, or one or more components of $\mathrm{x}_{\mathrm{B}}$ are reduced to zero.

Note that we face the possibilit of having moer than m non-zero variables at any step of teration. This stage comes when the new point generated at some step occurs were $\delta f / \delta x_{R}$ becomes zero. Geometrically, this means that we are no longe at an extreme point of the convese set formed by the constaints, and thus no longer have a basic solution with respect to the original constraint set. When this happens, we simply define a new variables $s_{i}$ as $s_{\mathrm{i}}=\delta / f \delta x_{\mathrm{Ri}}$ and a new constraint $s_{\mathrm{i}}=0$.

Step 5. At this stage, we have $m+1$ non-zero varibles and $m+1$ constraints, which is a basic solution to the extended set of constaints.

Step. Repeat the above procedure until no further improvement of the objective function may be obtained by increasing one of the non-basic variables.

Example 5.4.1. Using Beale's method solve the QPP
Maximize $z=5+4 x_{1}+6 x_{2}-2 x_{1}^{2}-2 x_{1} x_{2}-2 x_{2}^{2}$
subject to

$$
\begin{aligned}
& x_{1}+2 x_{2} \leq 0 \\
& x_{1}, x_{2} \geq 0
\end{aligned}
$$

## Solution :

Introducing slacle variable $x_{3} \geq 0$, the given QPP becomes
Maximize $z=5+4 x_{1}+6 x_{2}-2 x_{1}^{2}-2 x_{1} x_{2}-2 x_{2}^{2}$
subject to $\quad x_{1}+2 x_{2}+x_{3}=2$

$$
x_{1}, x_{2}, x_{3} \geq 0
$$

We choose $x_{1}$ arbitarily as basic variable and express it in terms of $x_{2}$ and $x_{3}$. Thus

$$
x_{1}=2-2 x_{2}-x_{3}
$$

We now express the objective functions $z$ in terms of $x_{R} z=5+4\left(2-2 x_{2}\right.$ $\left.x_{3}\right)+6 x_{2}-2\left(2-2 x_{2}-x_{3}\right)^{2}-2\left(2-2 x_{2}-x_{3}\right) x_{2}-2 x_{2}^{2}$
$\therefore \frac{\partial z}{\partial x_{2}}=-8+6-4\left(2-2 x_{2}-x_{3}\right)(-2)-2\left(2-4 x_{2}-x_{3}\right)-4 x_{2}$
At $x_{2}=0 \mathrm{~m} x_{3}=0$ We have $\frac{\partial z}{\partial x_{2}}=-8+6+14-4=10$

This means $z$ will increase if $x_{2}$ is increased from zero.
Also $\frac{\partial z}{\partial x_{3}}=-4+4\left(2-2 x_{2}-x_{3}\right)+2 x_{2}$
$\therefore$ At $x_{2}=0, x_{3}=0$ we have $\frac{\partial z}{\partial x_{3}}=-4+8=4$
We see that the rate of increase of $z$ with respect to $x_{2}$ is more.
Hence incease in $x_{2}$ will give better improvement in the objective function.
To find how much $x_{2}$ should or may increase, we check two quantities.
(i) the value of $x_{2}$ for which $\delta z / \delta x_{2}$ vanishes.
(ii) the largest value of $x_{2}$ attained without deriving the basic variable $x_{1}$ negative.

Then $x_{2}$ will be minimum of these two.
Now $\delta z / \delta x_{2}=0$ gives for $x_{3}=0$

$$
-2+8\left(2-2 x_{2}\right)-2\left(2-4 x_{2}\right)-4 x_{2}=0
$$

or, $-2+16-16 x_{2}-4+8 x_{2}-4 x_{2}=0$
or, $-12 x_{2}+10=0$
or, $x_{2}=5 / 6$
And for $x_{3}=0, x_{1}<0$ gives $2-2 x_{2}<0$ or, $x_{2}>1$
We have $\min \{5 / 6,1\}=5.6$. Thus the new basic variable is $x_{2}$.
Expressing $x_{2}$ is terms of $x_{1}$ and $x_{3}$ we get

$$
x_{2}=1-x_{1 / 2}-x_{3 / 2}
$$

We now express $z$ in terms of $x_{1}$ and $x_{3}$ as

$$
\begin{array}{r}
z=5+4 x_{1}+6\left(1-x_{1 / 2}-x_{3 / 2}\right)-2 x_{1}^{2}-2 x_{1}\left(1-x_{1 / 2}-x_{3 / 2}\right) \\
-2\left(1-x_{1 / 2}-x_{3 / 2}\right)^{2}
\end{array}
$$

$$
\text { Now } \cdot \frac{\partial z}{\partial x_{1}}=4-6(-1 / 2)-4 x_{1}-2 x_{1}(-1 / 2)-2\left(1-x_{1 / 2}-x_{3 / 2}\right)
$$

$$
-4\left(1-x_{1 / 2}-x_{3 / 2}\right)(-1 / 2)
$$

$$
=1-3 x_{1}
$$

$$
\begin{aligned}
\frac{\partial z}{\partial x_{3}} & =6(1-1 / 2)-2 x_{1}(-1 / 2) 4\left(1-x_{1 / 2}-x_{3 / 2}\right)(-1 / 2) \\
& =-1-x_{3}
\end{aligned}
$$

At $x_{2}=0, x_{3}=0$ We have $\frac{\partial z}{\partial x_{1}}=1$ and $\frac{\partial z}{\partial x_{3}}=-1$
Ths $z$ increases as $x_{1}$ is increases. So $x_{1}$ can be introduced to incease $z$.
To find how much $x_{1}$ should or may increase, we check two quantities.
(i) the value of $x_{1}$ for which $\delta z / \delta x_{1}$ vanishes.
(ii) the largest value of $x_{1}$ attained without deriving the basic variable $x_{2}$ negative.

The $x_{1}$ will be minimum of these two.
Fo $x_{3}=0, \delta z / \delta x_{1}=0$ gives $1-3 x_{1}=0$ or $x_{1}=1 / 3$
For $x_{2}=0, x_{2}<0$ gives $1-x_{1 / 2}<0$ or, $x_{1}>2$
We have $\min \{1 / 3,2\}=1 / 3$
Hence we find $x_{1}=1 / 3$ and the new basic vaiable is $x_{1}$.
At $x_{1}=\frac{1}{3}, x_{3}=0$ we have $\frac{\partial z}{\partial x_{1}}=0, \frac{\partial z}{\partial x_{3}}=-1$. Thus the optimal solution has been attained \& the optimal solution is $x_{1}=1 / 3, x_{2}=1-1 / 6-0=5 / 6, x_{3}=0$ and $\operatorname{man} x=5+4 / 3+6 \times 5 / 6-2 \times(1 / 3)^{2}-2(1 / 3)(5 / 6)-2 \times(5 / 6)^{2}=55 / 6$

### 5.5 Summary

Quadratic programming problem is concerned with non linear progamming problem of maximizing (or minimizing) the quadratic objective function subject to a set of linear inequality constaints. Wolfe's modified simplex method and Beale's method are discussed here with examples.

### 5.6 Self Assessment Questions

1. Applying wolfe's method solve the following quadatic pogramming problems
(i) Maximize $\mathrm{f}=4 x_{1}+6 x_{2}-2 x_{1}^{2}-2 x_{1} x_{2}-2 x_{2}^{2}$
subject to $\quad x_{1}+2 x_{2} \leq 2$

$$
x_{1}, x_{1} \geq 0
$$

(ii) Maximize $\mathrm{z}=12 x_{1}+12 x_{2}-18 x_{1}^{2}-1 \angle x_{1} x_{2}-8 x_{2}^{2}$
subject to $\quad 3 x_{1}+4 x_{2} \leq 2$

$$
x_{1}, x_{1} \geq 0
$$

(iii) Maximize $\mathrm{f}=3 x_{1}+2 x_{2}-2 x_{2}^{2}$
subject to $\quad 4 x_{1}+x_{2} \leq 4$

$$
\begin{aligned}
& 2 x_{1}+x_{2} \leq 2 \\
& x_{1}, x_{2} \geq 0
\end{aligned}
$$

(iv) Maximize $\mathrm{z}=10 x_{1}+6 x_{2}-50 x_{1}^{2}$
subject to

$$
\begin{aligned}
& 5 x_{1}+8 x_{2} \leq 4 \\
& 5 x_{1}+4 x_{2} \leq 2 \\
& x_{1}, x_{2} \geq 0
\end{aligned}
$$

(iv) Maximize $\mathrm{f}=-4 x_{1}+x_{1}^{2}-2 x_{1} x_{2}+2 x_{2}^{2}$
subject to

$$
\begin{aligned}
& 2 x_{1}+x_{2} \leq 6 \\
& x_{1}-4 x_{2} \leq 0 \\
& x_{1}, x_{2} \geq 0
\end{aligned}
$$

(iv) Máximize $z=2 x_{1}+3 x_{2}-2 x_{1}^{2}$
subject to $\quad x_{1}+4 x_{2} \leq 4$

$$
x_{1}+x_{2} \leq 2
$$

$$
x_{1}, x_{2} \geq 0
$$

2. Use Beale's method of solve the following quadiratic linear programming problems
(i) Maximize $\mathrm{z}=6-6 x_{1}+2 x_{1}^{2}-2 x_{1} x_{2}+2 x_{2}^{2}$
subject to $\quad x_{1}+x_{2} \leq 2$

$$
x_{1}, x_{2} \geq 0
$$

(ii) Maximize $\mathrm{z}=2 x_{1}+3 x_{2}-x_{1}^{2}$
subject to $\quad x_{1}+2 x_{2} \leq 4$

$$
x_{1}, x_{2} \geq 0
$$

(iii) Maximize $\mathrm{f}=2 x_{1}+3 x_{2}-2 x_{2}^{2}$

$$
\begin{array}{ll}
\text { subject to } & x_{1}+4 x_{2} \leq 4 \\
& x_{1}+x_{2} \leq 2 \\
& x_{1}, x_{2} \geq 0
\end{array}
$$

(iv) Maximize $f=12 x_{1}+6 x_{2}-18 x_{1}^{2}-6 x_{1} x_{2}-2 x_{2}^{2}$
subject to

$$
3 x_{1}+2 x_{2} \leq 2
$$

$$
x_{1}, x_{2} \geq 0
$$

## Unit 6 Integer Programming Problem

## Structure

6.1 Introduction
6.2 Need for Integer. Programming
6.3 Gomory's cutting plane method for all IPP
6.3.1 Construction of Gomory's constraints
6.3.2 Gomory's cutting Plane Algorithm
6.4 The Branch and Bound Method
6.4.1 Branch and Bound Algorithm
6.5 Summary
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### 6.1 Introduction

Integer Programming Problem (IPP) is a special class of Linear Programming Problem where all or some of the variables in the optimal solution are restrieted to the integers. If all the variables are restricted to take integral values the IPP is termed as pure IPP. On the other hand, if only some variables are restricted to take only integer values then the problem is called mixed IPP.

In 1956, R. E. Gomory developed a method to solve pure IPP. Later, he extended the method to solve mixed IPP. Another important approach, called the "branch and bound" technique was developed for solving both the all integer and he mixed integer programming problems.

Several algorillms have yet been developed for solving both types of IPP. We shall discuss only.
(i) Gomory's cutting plane method for pure IPP. and
(ii) Branch and bound method.

### 6.2 Need for Integer Programming

To solve an IPP one may think to get the optimal solution just by rounding down the optimal solution of the corresponding LPP obtained by regular simplex method. But there is no gaurantee for this. It may or may not happen so. The integer solution obtaind by rounding down the optimal solution of the corresponding LPP will not always satisfy all constraints or will not give the actual optimal solution of the IPP. These are explained by following examples.

## Example 6.2.1

$$
\begin{aligned}
\text { Maximize } z= & 3 x_{1}-2 x_{2} \\
\text { subject to } \quad & 12 x_{1}+7 x_{2} \leq 28 \\
& x_{1}, x_{2} \geq 0 \\
& x_{1}, x_{2} \text { are integers. }
\end{aligned}
$$

Ignoring the integer restriction here the optimal solution is $x_{1}=2 \frac{1}{3}, x_{2}=0$ with $\max z=7$.

The solution obtained by rounding down this optimal solution is $x_{1}=2, x_{2}=$ 0 this solution is the optimal solution of the given Integer programming problem.

Example 6.2.2
Minimize $z=2 x_{1}+3 x_{2}$
subject to $\quad 80 x_{1}+31 x_{2} \geq 248$

$$
x_{1}, x_{2} \geq 0, x_{1}, x_{2} \text { are integers. }
$$

Here, ignoring the integer restriction, the optimal solution is $x_{1}=3 \frac{1}{10}, x_{2}=$ 0 with $\min z=6 \frac{1}{5}$

Rounding down the solution we get $x_{1}=3, x_{2}=0$
But this point does not lie in the feasible region since $80 \times 3+31.0=240$ $<248$.

Hence just rounding the optimal solution of the corresponding LPP to the given IPP we may not get the optimal solution of the IPP.

## Example 6.2.3

$$
\begin{aligned}
\text { Maximize } z= & 3 x_{1}+4 x_{2} \\
\text { subject to } & 4 x_{1}+6 x_{2} \leq 15 \\
& x_{1}, x_{2} \geq 0 \\
& x_{1}, x_{2} \text { are integers. }
\end{aligned}
$$

Ignoring the integer-valued restriction The optimal solution of the problem is $x_{1}$ $=3 \frac{3}{4}, x_{2}=0$ with $\max z=11 \frac{1}{4}$

Rounding off this solution we get $x_{1}=3, x_{2}=$ or, $x_{1}=4, x_{2}=0$.
For $x_{1}=3, x_{2}=0$ we have $z=3 \times 3+4 \times 0=9$
$x_{1}=4, x_{2}=0$ does no satisfy $4 x+6 x_{2} \leq 15$. Here the actual solution tothis IPP is $x_{1}=2, x_{2}=1$ with $\max z=10$.

### 6.3 Gomory's cutting plane method for all IPP

In this method we first find the optimal solution to the IPP by simplex method ingoring the integer valued restriction. If in the optimal solution all the variables have integer values, then it is also the optimum solution of the given IPP. But if not, then a new constraint, called secondary an Gomory's constraint is introduced to the problem which slice away non-integer optimal solution exhibited by the extreme point of the feasitle region of the associated LPP and at the same time leave all feasible integer solutions untouched. The new related LPP is then solved as usual. If the new optimal solution obtained does not satisfy the integer requirement, then another Gomory's constraint is added and the process is repeated iteratively until the required integer valued optimum solution is obtained. As each introduced Gomory's constraint cut off a portion of the feasible region of the related LPP, the method is called Gomory's cutting plane method.

### 6.3.1 Construction of Gomory's constraints

Ignoring the integer restriction let the optimal solution of the given IPP using simplex method be $x_{\mathrm{B}}$. Also let this optimal solution has at least one non-integer
component. If more than one basic variable are fractional, we select that non-integral variable which involves the largest fractional part.

As $x_{\mathrm{Br}}$ corresponds to the rth now of simplex table we consider the rh now. of the final tables as

$$
\begin{equation*}
\sum_{j=1}^{n} y_{r i} x_{j}=b_{r} \quad \ldots \quad \ldots \tag{1}
\end{equation*}
$$

Let $\left[y_{\mathrm{r} j}\right]$ denote the greatest integer less than $y_{\mathrm{rj}}$ and fri denote the positive fractional part of $y_{\mathrm{rj}}$. Similarly, let $\left[b_{\mathrm{r}}\right]$ and $f_{\mathrm{r}}$ be resputively the greatest integer less than $b_{r}$ and the positive fractional part of $b r$.

Then we have $y_{\mathrm{rj}}=\left[y_{\mathrm{r} j}\right]+f_{\mathrm{rj}}$

$$
\text { and } b_{\mathrm{r}}=\left[b_{\mathrm{r}}\right]+f_{\mathrm{r}} \text { where } 0<f_{\mathrm{rj}}<1 \text { and } 0<f_{\mathrm{r}}<1 \text {. }
$$

From (1) we have thus

$$
\begin{align*}
& \quad \sum_{j=1}^{n}\left[y_{r j}\right] x_{j}+\sum_{j=1}^{n} f_{r j} x_{j}=\left[b_{r}\right]+f_{r} \\
& \text { and } f_{r}-\sum_{j=1}^{n} f_{r i} x_{j}=\left[b_{r}\right]-\sum_{j=1}^{n}\left[y_{r i}\right] x_{j} . . . \tag{2}
\end{align*}
$$

For integer value of $x_{j}$ the RHS of (2) is an integer. So LHS of (2) must be an integer. Now $f_{r}$ is a proper fraction i.e. $0<f_{r}<1$ and $\sum_{j=1}^{n} f_{r j} x_{j}$ is positive thus (2) gives.
(A proper fraction) - (positive number) $=$ (integer)
Hence RHS is either zero or negative integes.
So LHS is also either zero or negative integer
i.e. LHS $\leq 0$

$$
\text { or, } \sum_{j=1}^{n} f_{i n} x_{j} \leq 0
$$

$$
\text { or, } \quad-\sum_{n=1}^{n} f_{i j} x_{j} \leq-f_{r}
$$

Introducing plack variable $x_{\mathrm{s}}$ this becomes

$$
-\sum_{n=1}^{n} f_{i j} x_{j}+x_{s}=-f_{r}
$$

This is the Gomory's constraints which is to be introduced to the given problem to form a new LPP to be solved the dual simplex method.

### 6.3.2 Gomory's cutting Plane Algorithm

The following are the four steps of solving ail integer IPP by Gomory's cutting plane method.

Step 1. Using simplex method find the optimal solution of the IPP ignoringthe integral value restructions.

Step 2. If all the variables have integral values, take this solution as the optimal solution of the given IPP.

If at least one varible in the optimal solution obtained in setp 1 has fractional value then identify the now involving the largest fractional part. Using this row from the Gomory's constraint.

Step 3. Augment the IPP by introducing the Gomory's constraint formed in step 2 and modify the table. Using dual simplex method find the new optimal solution of the augmented LPP.

Step 4. If all variables of the optimal solution obtained in setp 3 are integers, then this is the required optimal solution of the original IPP. Otherwise go to step 2 and again augment the IPP by a new Gomory's constraint.

Example 6.3.1 Use Gomory's cutting plane method to find the optimal solution of the IPP

$$
\begin{array}{ll}
\text { Maximize } z= & x_{1}+x_{2} \\
\text { subject to } \quad & 2 x_{1}+5 x_{2} \leq 16 \\
& 6 x_{1}+5 x_{2} \leq 30 \\
& x_{1}, x_{2} \geq 0 \\
& x_{1}, x_{2} \text { are integers. }
\end{array}
$$

Solution : Ignoring the intergal value restriction we solve itby simplex method. Introducing slack variables $x_{3}$ and $x_{4}$ the LPP becomes

$$
\text { Maximize } z=x_{1}+x_{2}+0 x_{3}+0 x_{4}
$$

subject to

$$
\begin{array}{ll}
2 x_{1}+5 x_{2}+x_{3} & =16 \\
6 x_{1}+5 x_{2}++x_{4}=30 \\
x_{1}, x_{2}, x_{3}, x_{4} \geq 0
\end{array}
$$

Using simplex method the tables are obtained

| $c_{\mathrm{j}}$ | 1 | 1 | 0 | 0 |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | $y_{B}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | min ratio |
| 0 | $y_{4}$ | 30 | 6 | 5 | 0 | 1 | $5 \rightarrow$ |
| $z=0$ |  |  | -1 | -1 | 0 | 0 |  |
| 0 | $y_{3}$ | 6 | 0 | $10 / 3$ | 1 | $-1 / 3$ | $9 / 5 \rightarrow$ |
| 1 | $y_{1}$ | 5 | 1 | $5 / 6$ | 0 | $1 / 6$ | 6 |
| $z=5$ |  |  | 0 | $-1 / 6$ | 0 | $1 / 6$ |  |
| 1 | $y_{2}$ | $9 / 5$ | 0 | 1 | $3 / 10$ | $-1 / 10$ |  |
| 1 | $y_{1}$ | $7 / 2$ | 1 | 0 | $-1 / 4$ | $1 / 4$ |  |
| $z=53 / 10$ |  |  | 0 | 0 | $1 / 20$ | $3 / 20$ |  |

In this object table we see that both the variables are fractional and are $9 / 5=$ $1+4 / 5,7 / 2=3+1 / 2$. The largest fractional part is $4 / 5$ and is associated with the first row. The first row written in the form of equation is

$$
x_{2}+(3 / 10) x_{3}-(1 / 10) x_{4}=9 / 5
$$

Writing $3 / 10=0+3 / 10,-1 / 10=-2+9 / 10$ and $9 / 5=1+4 / 5$ this becomes

$$
x_{2}+0 x_{3}+(3 / 10) x_{3}-2 x_{4}+(9 / 10) x_{4}=1+4 / 5
$$

$\therefore$ The Gomory's constraint is
$-(3 / 10) x_{3}-(9 / 10) x_{4}-\leq(4 / 5)$
Introducing plack variable $x_{5} \leq 0$ we get
$-(3 / 10) x_{3}-(9 / 10) x_{4}+x_{5}=-(4 / 5)$

Adding this Gomory's constraint to the above optimum table, we get modified table as follows :

|  | $c_{j}$ | 1 | 1 | 0 | 0 | 0 |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $C_{B}$ | $y_{B}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ |
| 1 | $y_{2}$ | $9 / 5$ | 0 | 1 | $3 / 10$ | $-1 / 10$ | 8 |
| 1 | $y_{1}$ | $7 / 2$ | 1 | 0 | $-1 / 4$ | $1 / 4$ | 0 |
| 0 | $y_{5}$ | $-4 / 5$ | 0 | 0 | $-3 / 10$ | $-9 / 10$ | 1 |
|  |  | $z_{j}-c_{\mathrm{j}}$ | 0 | 0 | $1 / 20$ | $3 / 20$ | 1 |
| $\frac{\left(z_{j}-c_{j}\right)}{y_{3 j}}: y_{3 j}<0$ |  |  |  | $1 / 20$ | $3 / 20$ |  |  |
| 1 | $y_{2}$ | 1 | 0 | 1 | 0 | -1 | 1 |
| 1 | $y_{1}$ | $25 / 6$ | 1 | 0 | 0 | 1 | $-5 / 6$ |
| 0 | $y_{3}$ | $8 / 3$ | 0 | 0 | 1 | 3 | $-10 / 3$ |
|  |  | $z_{j}-c_{\mathrm{j}}$ | 0 | 0 | 0 | 0 | $1 / 6$ |

In this optimal table the basic variable $x_{1}$ is fractional (it is a variable of the original given IPP). It is associated with second row. We consider the second row and write it as equation to form Gomory's second constraint.

$$
x_{1}+x_{5}-(5 / 6) x_{5}=25 / 6
$$

or, $x_{1}+x_{5}+(-1) x_{5}+(1 / 6) x_{5}=4+1 / 6$.
The Gomory's constraint is

$$
-(1 / 6) x_{5} \leq-(1 / 6)
$$

or. $-x_{5} \leq-1$
Adding slack variable $x_{6} 0$ we get

$$
-x_{5}+x_{6}=-1
$$

Adding the Gomory's constraint to the above optimum table and modifying the table we get

| $c_{\mathrm{j}}$ | 1 | 1 | 0 | 0 | 0 | 0 |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\mathrm{C}_{\mathrm{B}}$ | $y_{\mathrm{B}}$ | $x_{\mathrm{B}}$ | $y_{1}$ | $y_{2}$ | $y_{3}$ | $y_{4}$ | $y_{5}$ | $y_{6}$ |
| 1 | $y_{2}$ | 1 | 0 | 1 | 0 | -1 | 1 | 0 |
| 1 | $y_{1}$ | $25 / 6$ | 1 | 0 | 0 | 1 | $-5 / 6$ | 0 |
| 0 | $y_{3}$ | $8 / 3$ | 0 | 0 | 1 | 3 | $-10 / 3$ | 0 |
| 0 | $y_{6}$ | -1 | 0 | 0 | 0 | 0 | -1 | 1 |
|  |  | $z_{\mathrm{j}}-c_{\mathrm{j}}$ | 0 | 0 | 0 | 0 | $1 / 6$ | 0 |
| $\frac{\left(z_{j}-c_{j}\right)}{y_{3 j}}$ | $: y_{4 \mathrm{j}}<0$ |  |  |  |  |  | $1 / 6$ |  |
| 1 | $y_{2}$ | 0 | 0 | 1 | 0 | -1 | 0 | 1 |
| 1 | $y_{1}$ | 5 | 1 | 0 | 0 | 1 | 0 | $-5 / 6$ |
| 0 | $y_{3}$ | 6 | 0 | 0 | 1 | 3 | 0 | $-10 / 3$ |
|  | $y_{5}$ | 1 | 0 | 0 | 0 | 0 | 1 | -1 |

As the original variables are integers this is the final table of the IPP. The optimal solution is $x_{1}=5, x_{2}=0$ and $\max z=5$.

### 6.4 The Branch and Bound Method

The Branch and Bound method is most powerful method and is applicable to both pure as well as mixed integer programming prolbems. This method was developed by Landand Doig. The principal idea underlying the branch and bound method is an follows. First we are to solve the problem ignoring the integer valued restriction. If the optimal solution has non-integral value, say $x_{\mathrm{j}}$, then there is an integer $k$ such that $k<x_{\mathrm{j}}<k+1$. As we want $x_{\mathrm{j}}$ to have integer value, the value
of $x_{\mathrm{j}}$ must satisfy either $x_{\mathrm{j}} \leq k$ or $\mathrm{xj} \geq k+1$ but not noth. Adding these constraints individually to the constraints of the given problem two subproblems are obtained, These two subproblems are solved. Repating the branching, the desired optimal solution is obtained.

### 6.4.1 "ranch and Bound Algorithm

The step by step procedure of branch and bound algorithm is as follows :
Let the IPP be
Maximize $z=c x$
subject to $\quad \mathrm{A} x=\mathrm{b}$

$$
x \geq 0
$$

$x_{\mathrm{j}}$ is integer for $\mathrm{j} \in I$
Where $c=\left\{c_{1}, c_{2}, \ldots ., c_{n}\right], x=\left[x_{1}, x_{2}, \ldots ., x_{n}\right]^{\top}, b=\left[b_{1}, b_{2}, \ldots ., b_{m}\right]^{\top}$
$\mathrm{A}=\left[a_{\mathrm{ij}}\right]_{\mathrm{mxn}}$
If $\mathrm{I}=(1,2, \ldots ., n\}$ then it is a pure (or all) IPP and if 1 is a proper subset of $(1,2, \ldots \ldots, n)$ then it is a mixed IPP.

Step 1. Ignoringthe integer restriction solve the IPP. If the optimal solution be such that all $x_{j}, j \in I$ are integers, then this is the required optimal solution. If at least one $x_{\mathrm{j}}, \mathrm{J} \in \mathrm{I}$ be non-integer then go to next step.

Step 2. Among hon-integer $x_{\mathrm{j}}, \mathrm{j} \in \mathrm{I}$ chope any one, Then there exists integer k such that

$$
\mathrm{K}<x_{\mathrm{j}}<k+1
$$

As we want $x_{\mathrm{j}}$ to be an integer, the integer solution must satisfy one of the following

$$
x_{\mathrm{j}} \leq k \text { or } x_{\mathrm{j}} \geq k+1
$$

Add these constraints indirectly to the constraints of the current problem and get two sub-problems. Solve these two sub-problems.

Step 3. If for any of the subproblem integer solution is obtained then that problem is not further branched.

But if any subproblem involves some non-integer variable, then it is again branched. This process of branching is continued, until each subproblem either admits an integer valued solution or there is eirdence that it cannot yield a better solution or it gives no feasible solution.

Among all subproblems select that integer valued solution which gives the over all maximum value of the object function.

Note : Main disadvantage of this method is that it requires the optimal solution of each subproblem. For large size problem this become very tedions job. Inspite of this drawback it is most effective method for solving IPP. Also the method is applicable for both all and mixed IPP

Example 6.4.1 Using Branch and Bound technique solve the following IPP

$$
\begin{array}{ll}
\operatorname{Maximize} z= & x_{1}+x_{2} \\
\text { subject to } & 3 x_{1}+2 x_{2} \leq 12 \\
& x_{1}, x_{2} \leq 0 \\
& x_{1}, x_{2} \geq 0 \\
& x_{1}, x_{2} \text { are integers. }
\end{array}
$$

Solition : Ignoring the integer valued restriction the solution of the given IPP by graphical method is $x_{1}=8 / 3, x_{2}=2$, the value of $z$ is $4 \frac{2}{3}$. We call the LPP corresponding to this IPP as LPPI.

The value of $x_{1}$ is fraction and is $8 / 3$. We note that $2<8 / 3<3$.

So we from two subproblems with additional constraints respectively as $x_{1} \leq 2$ and $x_{1} \geq 3$.

Thus two problems are


$$
\begin{array}{ll}
\text { Maximize } z= & x_{1}+x_{2} \\
\text { subject to } & 3 x_{1}+2 x_{1} \leq 12 \\
& x_{2} \leq 2  \tag{LPP1.1}\\
& x_{1} \leq 2 \\
& x_{1}, x_{2} \geq 0
\end{array}
$$

and

$$
\begin{array}{ll}
\text { Maximize } z= & x_{1}+x_{2} \\
\text { subject to } & 3 x_{1}+2 x_{1} \leq 12 \\
& x \leq 2 \\
& x_{1} \leq 3 \\
& x_{1}, x_{2} \geq 0
\end{array}
$$

By graphical method, the optimal solution of the LPP1.1 is $x_{1}=2, x_{2}=2$ with $z=4$ and that of the LPP1. 2 is $x_{1}=3, x_{2}=3 / 2$ with $z=9 / 2$


LPPI.I


LPP1. 2

Since the optimal solution of the LPPI. 1 are integers there is no need to branch this problem further. On the other hand the optimal value of $x_{2}$ is fraction for the LPPI.2. So branching of the LPPI. 2 is to be done. Let the two subproblems obtained by branching by LPP1.2.1 and LPP1.2.2.

They are obtained to follows.
The optimal value of $x_{2}$ for LPP1.2 is $3 / 2$ and $1<3 / 2<2$.
$\therefore$ The additional constraints to be introduced are $x_{2} \leq 1$ and $x_{2} \geq 2$ respectively. Thus LPP1.2.1 and LPP1.2.2 are given by

$$
\begin{array}{ll}
\text { Maximize } z= & x_{1}+x_{2} \\
\text { subject to } & 3 x_{1}+2 x_{1} \leq 12 \\
& x_{2} \leq 2 \\
& x_{1} \geq 3 \\
& x_{2} \leq 1 \\
& x_{1}, x_{2} \geq 0
\end{array}
$$

$$
x_{1} \geq 3 \quad \text {.... } \quad . . \quad \text { (LPP1,2,1) }
$$

and

$$
\begin{array}{ll}
\text { Maximize } z= & x_{1}+x_{2} \\
\text { subject to } & 3 x_{1}+2 x_{1} \leq 12 \\
& x_{2} \leq 2
\end{array}
$$

$$
\begin{aligned}
& x_{1} \geq 3 \\
& x_{2} \geq 2 \\
& x_{1}, x_{2} \geq 0
\end{aligned}
$$



LPP1.2.1


LPP1.2.2

Using graphical method the optimal solution of the LPP1.2.1 is $x_{1}=10 / 3, x_{2}$ $=1$ with the value of $z=13 / 3=4 \frac{1}{3}$. As $x_{1}$ is not an integer and $z=13 / 3$ which is greater that the optimal value $z=4$ of the LPP1.1, we need branching of this LPP toget LPP1.2.1.1, and LPP1.2.1.2. (Here we note that instead of $z=13 / 3$ if the value of $z$ would be less than 4 then no branching is nedded)

The LPP1.2.2. has no feasible, so no question of branching.
To get branching of LPP1.2.1. we note that $3<10 / 3<4$. So that additional constraints to the LPP.1.2.1 to get sub problem are respectively $x_{1} \leq 3$ and $x_{1} \geq 4$.

Thus the subproblems are given by

$$
\begin{array}{ll}
\text { Maximize } z= & x_{1}+x_{2} \\
\text { subject to } & 3 x_{1}+2 x_{1} \leq 12 \\
& x_{2} \leq 2 \\
& x_{1} \geq 3  \tag{LPPI.2.1.1}\\
& x_{2} \leq 1 \\
& x_{1} \leq 3 \\
& x_{1}, x_{2} \geq 0
\end{array}
$$

and

$$
\begin{array}{ll}
\text { Maximize } z= & x_{1}+x_{2} \\
\text { subject to } & 3 x_{1}+2 x_{1} \leq 12 \\
& x_{2} \leq 2 \\
& x_{1} \geq 3  \tag{LPP1.2.1.2}\\
& x_{2} \leq 1 \\
& x_{1} \geq 4 \\
& x_{1}, x_{2} \geq 0
\end{array}
$$



LPP1.2.1.1


LPP1.2.1.2

Graphical we get the optimal solution of the LPP1.2.11 as $x_{1}=3, x_{2}=1$ with $z=4$ which is some as the optimal value of $z$ of the LPP1.1. The optimal solution of the LPP 1.2.1.2. is $x_{1}=4, x_{2}=0$ with $z=4$. No further branching is necessary.

The over all maximum value of the objective function is $z=4$ and the integer valued solution are $x_{1}=2 ; x_{1}=3, x_{2}=1 ; x_{1}=4, x_{2}=0$.

### 6.5. Summary

Gomory cutting plane method for all IPP and Branch and bound method for general IPP have been considered and explained with examples. Need forIPP has been explained in detail with examples.

### 6.6 Self Assessment Questions

1. Solve the following IPP using 'Gomory's cutting plane method.
(i) Maximize $z=2 x_{1}+2 x_{2}$ subject to

$$
5 x_{1}+3 x_{2} \leq 8
$$

$$
x_{1}+2 x_{2} \leq 4
$$

$$
x_{1}, x_{2} \geq 0
$$

$x_{1}, x_{2}$ are integers
[Ans: $x_{1}=1, x_{2}=1, \max z=4$ ]
(ii) Maximize $z=4 x_{1}+3 x_{2}$
subject to
$3 x_{1}+4 x_{2} \leq 12$
$4 x_{1}+2 x_{2} \leq 9$
$x_{1}, x_{2} \geq 0$
$x_{1}, x_{2}$ are integers
[ Ans : $x_{1}=1, x_{2}=2$, max $z=10$ ]
(iii) Maximize $z=x_{1}-2 x_{2}$
subject to $\quad 4 x_{1}+2 x_{2} \leq 15$
$x_{1}, x_{2} \geq 0$
$x_{1}, x_{2}$ are integers
[ Ans : $x_{1}=3, x_{2}=0$, max $z=3$ ]
2. Using Branch and Bound method solve the following IPP
(i) Maximize $z=3 x_{1}+4 x_{2}$
subject to $\quad 3 x_{1}+2 x_{2} \leq 8$
$x_{1}+4 x_{2} \leq 0$
$x_{1}, x_{2} \geq 0$
$x_{1}, x_{2}$ are integers
[ Ans : $x_{1}=1, x_{2}=1, \max z=11$ ]
(ii) Maximize $z=7 x_{1}+9 x_{2}$

$$
\text { subject to } \quad-x_{1}+3 x_{2} \leq 6
$$

$7 x_{1}+x_{2} \leq 35$
$0 \leq x_{1} \leq 7$
$0 \leq x_{2} \leq 7$
$x_{1}, x_{2}$ are integets
[ Ans : $x_{1}=4, x_{2}=3$, $\max z=55$ ]

## Unit $7 \square$ One dimensional minimization method

## Structure

### 7.1 Introduction

### 7.2 Unimodal Function

### 7.2.1 Definition

### 7.3 Fibonacci Method

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7.5 Golden Section
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### 7.1 Introduction

Numerical method ofoptimization are used to solve the problems involving objective function and/or constraints which are two complicated or cannot be expressed as explicit function.

One dimensional minimization method plays an important role to solve the problems using numerical technique. In numerical methods we are to minimize $f\left(x_{i}+\lambda_{i} S_{i}\right)$ with respect to $l_{i}$ for known values of $x_{i}$ and $s_{i}$.

This is nothing but aone dimensional minimization problem. Among many onedimensional minimization methods Fibonacci method and golden section method are simple and important. They are discussed in this unit. These two methods are used for unimodal functions.

### 7.2 Unimodal Function

In the process of finding optimal point often it becomes necessary that the function has only one optimum point in the domain of search. As in many methods we need only the values of the function at various points, the functionmay not be continuous and differentiable. What we need is that it should be unimodal. Unimodality of a function of one variable is defined as follows

### 7.2.1. Definition

A real valued function $\mathrm{f}(\mathrm{x})$ is said to be unimodal (minimum) is $[a, b]$ if there is a point $x^{*} \in[a, b]$ such that
(i) it a $<x_{1}<x_{2} x^{*}$ then $f\left(x_{1}\right)>f\left(x_{2}\right)$
(ii) it a $<x_{1}<x_{2}<b$ then $f\left(x_{2}\right)>f\left(x_{1}\right)$

### 7.3 Fibonacci Method

Fibonacci method is based on Fibonacci sequence ( Fn ) defined by

$$
\begin{aligned}
& F_{0}=F_{1}=1 \\
& F_{n}+F_{n-1}+F_{n-2}, n=2,3,4,
\end{aligned}
$$

Thus
$F_{0}=1, F_{1}=1, F_{2}=2, F_{3}=3, F_{4}=5, F_{5}=8, F_{6}=13, F_{7}=21, F_{8}=$ $34, F_{9}=55, F_{10}=89, F_{11}=144$,

Fibonacci method can be used to find the optimum of a function of one variable. The function must be unimodal, it may ormay not be continuous or differtiable. This method has the following limitations :
(i) The initial interval of uncertainty $[a, b]$, in which the optimum lies, has to be known
(ii) The function to be optimized has to be unimodal in the initial interval of uncertainty.
(iii) The exact optimum point cannot be located by this method. Only an interval, known as the final interval of uncertainty can be obtained.

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(iv) The number of function evaluations to be used in the search has to be specified beforehand.

The final interval of uncertainly can be made as small as we desire by making the number of function evaluations more.

Procedure : Let $L$ be the length of the initial interval of uncertainty $[a, b]$ be the initial interval of uncertanity. Therefore $\mathrm{L}_{0}=b-a$.

Let $n$ be the total number of experiments to be conducted. We define $\mathrm{L}_{2}^{+}=\frac{\mathrm{F}_{n-2}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}$

The first two experiments are placed at the points $x_{1}$ and $x_{2}$ which are located at a distance $L_{2}^{*}$ from each end of $L_{0}$. The values of the function $f$ at $x_{1}, x_{2}$ are evaluated as $f_{1}=f\left(x_{1}\right)$ at $f_{2}=f\left(x_{2}\right)$. Using unimodality assumption one of the intervals [ $\left.a, x_{1}\right]$ and $\left[x_{2}, b\right]$ is to be discarded. The remaning interval of uncertainty is denoted by $\mathrm{L}_{2}$.

$$
\text { Then } \begin{aligned}
\mathrm{L}_{2} & =\mathrm{L}_{0}-\mathrm{L}_{2}^{*} \\
& =\mathrm{L}_{0}-\frac{\mathrm{F}_{n-2}}{\mathrm{~F}_{n}} \cdot \mathrm{~L}_{0} \\
& =\mathrm{L}_{0}\left(\frac{\mathrm{~F}_{n}-\mathrm{F}_{n-2}}{\mathrm{~F}_{n}}\right) \\
& =\frac{\mathrm{F}_{n-1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}
\end{aligned}
$$

Now $\mathrm{L}_{2}-\mathrm{L}_{2}^{*}$

$$
\begin{aligned}
& =\frac{\mathrm{F}_{n-1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}-\frac{\mathrm{F}_{n-2}}{\mathrm{~F}_{n}} \mathrm{~L}_{0} \\
& =\frac{\mathrm{L}_{0}}{\mathrm{~F}_{n}}\left(\mathrm{~F}_{n-1}-\mathrm{F}_{n-2}\right) \\
& =\frac{\mathrm{L}_{0}}{\mathrm{~F}_{n}}\left(\mathrm{~F}_{n-2}+\mathrm{F}_{n-3}-\mathrm{F}_{n-2}\right)
\end{aligned}
$$

$$
\begin{aligned}
& =\frac{\mathrm{F}_{n-3}}{\mathrm{~F}_{n}} \mathrm{~L}_{0} \\
& \therefore \quad \frac{\mathrm{~L}_{2}-\mathrm{L}_{2}}{\mathrm{~L}_{2}^{*}}=\frac{\mathrm{F}_{n-3}}{\mathrm{~F}_{n-2}}<1
\end{aligned}
$$

i.e. $\mathrm{L}_{2}-\mathrm{L}_{2}^{*}<\mathrm{L}_{2}^{*}$

Thus in the interval of uncertainty L2 there is one point, either $x_{1}$ or $x_{2}$, whose distance from the two ends of $\mathrm{L}_{2}$ are $\mathrm{L}_{2}^{*}$ and $\mathrm{L}_{2}-\mathrm{L}_{2}^{*}$. The smaller of the two $\mathrm{L}_{2}-\mathrm{L}_{2}^{*} \& \mathrm{~L}_{2}^{*}$ is denoted) by. i.e $\mathrm{L}_{3}^{*}, \mathrm{~L}_{3}^{*}=\mathrm{L} 2-\mathrm{L}_{2}^{*}$

Now, $\mathrm{L}_{3}^{*}=\mathrm{L}_{2}-\mathrm{L}_{2}^{*}=\frac{\mathrm{F}_{n-3}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}$.
We now place the third experiment $x_{3}$ and $\mathrm{L}_{2}$ so that the current two experiment are located at a distance $\mathrm{L}_{3}$ from each end of $\mathrm{L}_{2}$. Again by the unimodal property we can reduce the interval of uncertainty from $L_{2}$ to $L_{3}$ given by $L_{3}=L_{2}-L_{3}^{*}$ $=\frac{F_{n-2}}{F_{n}} L_{0}$.
$\therefore$ The interval of uncertainty at the end of 3rd experiment is given by

$$
\mathrm{L}_{3}=\frac{\mathrm{F}_{n-2}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}
$$

and this obtained by discarding $L_{3}^{*}=\frac{\mathrm{F}_{n-3}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}$ continuing in this manner we have the following result in general.

The $j$ th experiment is to be placed at a distance $L_{j}^{*}=\frac{F_{n-j}}{F_{n}} L_{0}$ from one end of $L_{j-1}$ and the interval of uncertainty at the end of $j$ th experiment is given by $\mathrm{L}_{\mathrm{j}}=\frac{\mathrm{F}_{n-j+1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}$

Taking $\mathrm{j}=\mathrm{n}$ we see that the nth experiment is to be placed at a distance $\mathrm{L}_{n}^{*}$
$=\frac{F_{0}}{F_{n}} L_{0}=\frac{L_{0}}{F_{n}}$ from one end of $L_{n-1}$ and the interval of uncertainty at the end of n th experiment is given by $\mathrm{L}_{\mathrm{n}}=\frac{\mathrm{F}_{1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}=\frac{\mathrm{L}_{0}}{\mathrm{~F}_{n}}$

$$
\begin{aligned}
& \text { Now } L_{n-1}=\frac{\mathrm{F}_{n-(n-1)+1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}=\frac{\mathrm{F}_{2}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}=\frac{2 \mathrm{~L}_{0}}{\mathrm{~F}_{n}} \\
& \therefore \mathrm{~L}_{n}=\frac{1}{2} \mathrm{~L}_{n-1}
\end{aligned}
$$

Therefore, the last two experiments are located at a distance $\mathrm{L}_{n}^{*}=\frac{1}{2} \mathrm{~L}_{n-1}$ from each end of $L_{n-1}$. So they have the same location. To remove this difficulty we place the n th experiment very close to the remaining valid experiment in $\mathrm{L}_{\mathrm{n}-1}$. This enables us to obtain the final interest of uncertainty of length $\frac{1}{2} L_{n-1}=L_{n}=\frac{L_{0}}{F_{n}}$

From $L_{n}=\frac{L_{0}}{F_{n}}$ we note that we can determine $n$ for given $L_{n}$

### 7.4 Illustrative Examples

Example 7.4.1 : Maximize $f(x)=\left\{\begin{array}{l}2 x / 3, x \leq 3 \\ 5-x, x>3\end{array}\right.$
in the interest [1, 4] by Fibonacci method using $n=6$
Solution : Here number it experiment to be performed is $n:=6$.
From Fibonacei sequence we have
$\mathrm{F}_{0}=\mathrm{F}_{1}=1$
$F_{2}=2, F_{3}=3, F_{4}=5, F_{5}=8, F_{6}=13, F_{7}=21 \mathrm{etc}$.
Here $L_{0}=4-1=3$.
$\therefore L_{2}=\frac{\mathrm{F}_{4}}{\mathrm{~F}_{6}} \mathrm{~L}_{0}=\frac{5}{13} \times 3=1.1538$

The first two experiments are placed at the positions $x_{1}$ and $x_{2}$ such that

$$
x_{1}=1+L_{i}^{0}=1+1.1538=2.1538
$$

$$
\text { \& } x_{2}=4-L_{1}=4-1.1538=2.8462
$$

Now $f_{1}=f\left(x_{1}\right)=\frac{2 x_{1}}{3}=\frac{2 \times 2.1538}{3}=1.4359$
and $f_{2}=f\left(x_{2}\right)=\frac{2 x_{2}}{3}=\frac{2 \times 2.8462}{3}=1.8975$
Since $f_{1}<f_{2}$, using uniondal property we delete the interval $\left[1, x_{1}\right]$. Thus the reduced interval of uncertainly is $\left[x_{1}, 4\right]$ i.e., $[1.4359,4]$ with $x_{2}$ inside it and near to $x_{1}$.

The third experiment is placed at the position $x_{3}$ given by

$$
\begin{aligned}
4-x_{3} & =x_{2}-x_{1} \\
x_{3} & =4-x_{2}+x 1 \\
& =4-2.8462+2 \cdot 1538 \\
& =3.3076
\end{aligned}
$$

or,

Now $f_{3}=f\left(x_{3}\right)=5-x_{3}=5-3 \cdot 3076=1.6924$
Here $f_{3}<f_{2}$. So by unimodaily we delete the interval $\left[x_{3}, 4\right]$. The remaining interval of uncertainly becomes $\left[x_{1}, x_{3}\right]$ with $x_{2}$ inside it and near to the point $x_{3}$.

The fourth experiment is placed at $x_{4}$ given by

$$
\begin{aligned}
& x_{4}-x_{1}=x_{3}-x_{2} \\
& \therefore x_{4}=x_{1}+x_{3}-x_{2}=2.1538+3.3076-2.8462=2.6152 \\
& \text { Now, } f_{4}=\left(x_{4}\right)=\frac{2 x_{4}}{3}=\frac{2 \times 2.1652}{3}=1.7435
\end{aligned}
$$

Since $f_{4}<f_{2}$ we delite the interval $\left[x_{1}, x_{4}\right]$. The remaining interval of uncertainly is $\left[x_{4}, x_{3}\right]$ with $x_{2}$ inside it and near to $x_{4}$.

The fifth experiment is placed of $x_{5}$ given by

$$
x_{3}-x_{5}=x_{2}-x_{4}
$$

or, $x_{5}=x_{3}-x_{2}+x_{4}=3.3076-2.8462+2.6152=3.0766$
Now $f_{5}=f\left(x_{5}\right)=5-x_{5}=5-3.0766=1.9234$
Since $f_{5}<f_{2}$, using unimodal property we delete the interval $\left[x_{4}, x_{2}\right]$. The remaining interval of uncertainty is $\left[x_{2}, x_{3}\right]$ with $x_{5}$ inside it and near $x_{2}$.

The sixth experiment is placed at $x_{6}$ given by

$$
\begin{aligned}
& \quad x_{3}-x_{6}=x_{5}-x_{2} \\
& \text { or, } x_{6}=x_{3}-x_{5}+x_{2}=3.3076-3.0766+2.8462=3.0772 \\
& \text { Now } f_{6}=f\left(x_{6}\right)=5-x_{6}=5-3.0772=1.9228
\end{aligned}
$$

since $f_{6}<f_{5}$, using unimodality we delete the interval $\left[x_{6}, x_{3}\right]$. The final interval of uncertainty is $\left[x_{2}, x_{6}\right]=[2.8462,3.0772]$

Here we note that if the exact calculation be carried out then we would get $x_{5}=x_{6}$. In that situation $x_{6}$ should be selected very close to $x_{5}$. But here we see $x_{5} \neq x_{6}$. This is due to round off error involved in the calcution.

### 7.5 Golden Section

Ancient Greek architects believed that a building having sides $b$ and $c$ satisfying | $\square c$. | $\begin{array}{l}\text { the relation } \frac{b+c}{b}=\frac{b}{c}=\gamma \text { will be having the most } \\ \text { pleasing properties. This ratio is called Golden ration. } \\ \text { tt is also found in Euclid's geometry that the division } \\ \text { of a line segment into unequal parts so that the ration } \\ \text { of the whole to the largest part is equal to the ratio } \\ \text { of the large part to the smaller part: This section is } \\ \text { known as the golden section }\end{array}$ |
| :---: | :--- |

Thus the Golden section

$$
\frac{A C}{A B}=\frac{A B}{B C}=\gamma \quad \text { i.e., } \frac{A B+B C}{A B}=\frac{A B}{B C}=\gamma
$$

From this we have

$$
\frac{A B}{A B}+\frac{B C}{A B}=\frac{A B}{B C}=\gamma
$$

$$
\begin{aligned}
& \text { or, } 1+\frac{1}{\gamma}=\gamma \\
& \text { or, } \gamma^{2}-\gamma-1=0 \\
& \therefore \quad \gamma=\frac{-(-1) \pm \sqrt{(-1)^{2}-4.1 \cdot(-1)}}{2.1} \\
& \quad=\frac{1 \pm \sqrt{5}}{2}
\end{aligned}
$$

Since $\gamma$ is a positive number we have

$$
\gamma=\frac{\sqrt{5}+1}{2}=1.618
$$

### 7.6 Golden Section Method

Golden section method is similar to the Fibonacei method except for one difference. The difference is that in Fibonacei method the total number of experiments to be performed has to be specified before beginning the calculation, whereas, this is not required in golden section method. In fact when $n$ is very large then Fibonacei method reduces to golden section method. In Fibonacei method the number of experiments to be performed is decided at the begining but in golden section method the total number of experiments are to be decided during the computations.

In the Fibonacei method, the interval of uncertanity at the end of two experiments is given by $L_{2}=\frac{F_{n-1}}{F_{n}} L_{0}$

In Golden Section method is $n$ is very large this $L_{2}$ becomes

$$
\mathrm{L}_{2}=\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}=\mathrm{L}_{0}\left(\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}}\right)
$$

Also in Fibona method $\mathrm{L}_{3}$ is given by

$$
\mathrm{L}_{3}=\frac{\mathrm{F}_{n-2}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}
$$

$\therefore$ In Golden section method $\mathrm{L}_{3}$ will be given by

$$
\begin{aligned}
& \mathbf{L}_{3}=\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0} \\
& =\lim _{n \rightarrow \infty}\left(\frac{F_{n-2}}{F_{n-1}} \cdot \frac{F_{n-1}}{F_{n}} L_{0}\right) \\
& =L_{0} \cdot\left(\lim _{n \rightarrow \infty} \frac{F_{n-2}}{F_{n-1}}\right)\left(\lim _{n \rightarrow \infty} \frac{F_{n-1}}{F_{n}}\right) \\
& =L_{0} \cdot\left(\lim _{n \rightarrow \infty} \frac{F_{n-1}}{F_{n}}\right)\left(\lim _{n \rightarrow \infty} \frac{F_{n-1}}{F_{n}}\right) \\
& =L_{0} \cdot\left(\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}}\right)^{2}
\end{aligned}
$$

Similary, we get $L_{4}=L_{0}\left(\lim _{n \rightarrow \infty} \frac{F_{n-1}}{F_{n}}\right)^{3}$
Generalizing these results we have

$$
L_{k}=\left(\lim _{n \rightarrow \infty} \frac{F_{n-1}}{F_{n}}\right)^{k-1} \cdot L_{0}
$$

We have the relation

$$
\begin{aligned}
& \quad \mathrm{F}_{\mathrm{n}}=\mathrm{F}_{n-1}+\mathrm{F}_{n-2} \\
& \therefore \frac{\mathrm{~F}_{n}}{\mathrm{~F}_{n-1}}=1+\frac{\mathrm{F}_{n-2}}{\mathrm{~F}_{n-1}} \\
& \text { or, } \lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n}}{\mathrm{~F}_{n-1}}=1+\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-2}}{\mathrm{~F}_{n-1}} \\
& \\
& =1+\lim _{n \rightarrow \infty} \frac{1}{\mathrm{~F}_{n-1}} \mathrm{~F}_{n-2}
\end{aligned}
$$

$$
\begin{aligned}
& =1+\frac{1}{\lim _{n \rightarrow \infty} \frac{F_{n-1}}{F_{n-2}}} \\
& =1+\frac{1}{\lim _{n \rightarrow \infty} \frac{F_{n}}{F_{n-1}}}
\end{aligned}
$$

Let $\gamma=\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n}}{\mathrm{~F}_{n-1}}$
$\therefore$ We have $\gamma=1+\frac{1}{\gamma}$

$$
\begin{aligned}
& \text { or, } \gamma^{2}=g+1 \\
& \text { or, } \gamma^{2}-g-1=0 \\
& \text { or, } \gamma=\frac{-(-1) \pm \sqrt{(-1)^{2}-4.1(-1)}}{2.1} \\
& \quad=\frac{1 \pm \sqrt{5}}{2}
\end{aligned}
$$

Since $\gamma$ is a positive real number, we have $\gamma=\frac{\sqrt{5}+1}{2}=1 \cdot 618$, which is nothing but golden ratio or golden section.

Hence we have in general,

$$
L_{k}=\left(\frac{1}{\gamma}\right)^{k-1} L_{0}=(0.618)^{k-1} L_{0}
$$

$\therefore$ In the Golden section method the interval of uncertainty at the end of $k$ th experiment is given by

$$
L_{k}=(0.618)^{k-1} L_{0}
$$

### 7.7 Procedure of Golden Section Method

In the Fibonacci method, the location of the first two experiments are the points situated at a distance $L_{2}^{*}$ from the two ends of the initial interval of uncertainty, where $L_{2}^{*}$ is given by

$$
L_{2}^{*}=\frac{\mathrm{F}_{n-2}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}
$$

In Golden section method n is very large. Therefore $\mathrm{L}_{2}^{*}$ is given by

$$
\begin{aligned}
L_{2}^{*} & =\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-2}}{\mathrm{~F}_{n}} L_{0} \\
& =\lim _{n \rightarrow \infty} \cdot\left(\frac{\mathrm{~F}_{n-2}}{\mathrm{~F}_{n-1}} \cdot \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}} \mathrm{~L}_{0}\right) \\
& =\mathrm{L}_{0} \cdot \lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-2}}{\mathrm{~F}_{n-1}} \cdot \lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}} \\
& =\mathrm{L}_{0} \cdot \lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}} \cdot \lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}} \\
& =\mathrm{L}_{0} \cdot\left(\lim _{n \rightarrow \infty} \frac{\mathrm{~F}_{n-1}}{\mathrm{~F}_{n}}\right)^{2} \\
& =\mathrm{L}_{0} \cdot\left(\frac{1}{\gamma}\right)^{2} \\
& =\mathrm{L}_{0} \cdot(\cdot 613)^{2}=0.382 \mathrm{~L}_{0} .
\end{aligned}
$$

$\therefore$ In the Golden section method, the first two experiments are placed at the points $x_{1}$ and $x_{2}$ which are located at a distance $L_{2}^{*}=0.382 \mathrm{~L}_{0}$ from each end of $\mathrm{L}_{0}$. The values of the functions $f$ at $x_{1}, x_{2}$ are evaluated as $f_{1}=f\left(x_{1}\right)$ and $f_{2}=$ $f\left(x_{2}\right)$. Using the assumption of unimodality, one of the two intervals $[a, x]$ and $\left[x_{2}\right.$, $b]$ can be discarded. The remaining interval of uncertainty will be $\mathrm{L}_{2}=0.618 \mathrm{~L}_{0}$. The interval will contain one experiment point. The smaller distance of this experiment point from the ends of $L_{2}$ is denoted by $L_{3}^{*}$. The third experiment $x_{3}$ is placed in $L_{2}$ so that the current two experiments are located at a distance $L_{3}^{0}$ from each end of $L_{2}$. Again using unimodelity we can discard one of the end intervals and the
reduced internal of uncertainty at the end of 3rd experiment becomes $\mathrm{L}_{3}=(0.618)^{2}$ $\mathrm{L}_{0}$. This process is continued until the desired length of the interval of uncertainty is obtained.

### 7.8 Illustrative Examples

Example 7.8.1 Maximize $f(x)= \begin{cases}2 x / 3, & x \leq 3 \\ 5-x, & x>3\end{cases}$
in the interval $[1,4]$ by Golden selection method up to six experiments.
Solution : We have $\mathrm{L}_{0}=4-1=3$
Now $L_{2}^{*}=382 L_{0}=.382 \times 3=1.146$
The first two experiments are placed at the positions $x_{1}$ and $x_{2}$ such that

$$
\begin{aligned}
& x_{1}=1+L_{2}^{*}=1+1 \cdot 164=2 \cdot 146 \\
& x_{2}=4-L_{2}^{*}=1 \cdot 146=2 \cdot 854
\end{aligned}
$$

Now $f_{1}=f\left(x_{1}\right)=\frac{2 x_{1}}{3}=\frac{2 \times 2 \cdot 146}{3}=1.43066$

$$
f_{2}=f\left(x_{2}\right)=\frac{2 x_{2}}{3}=\frac{2 \times 2 \cdot 854}{3}=1.90266
$$

As $f_{1}<f_{2}$ and the problem is of maximization, using unimodal property we felete the interval $\left[1, x_{1}\right]$. Thus the reduced interval of uncertainly is $[x, 4]$ with $x_{2}$ inside it and near to the point $x_{1}$.

The third experiment is to be placed at $x_{3}$ given by

$$
\begin{aligned}
& \quad 4-x_{3}=x_{2}-x_{1} \\
& \text { or, } x_{3}=4-x_{2}+x_{1}=4-2.854+2.146=3.292 \\
& \text { Now, } f_{3}=f\left(x_{3}\right)=5-3.292=1.708
\end{aligned}
$$

Here $f_{3}<f_{2}$. So by unimodality we delete the interval $\left[x_{3}, 4\right]$. The remaning 139
interval of uncertainty becomes $\left[x_{1}, x_{3}\right]$ with $x_{2}$ inside it and near to the point $x_{3}$. The fourth experiment is placed at $x_{4}$ given by

$$
x_{4}-x_{1}=x_{3}-x_{2}
$$

or, $x_{4}=x_{1}+x_{3}-x_{2}=2.146+3.292-2.854=2.584$
Now $f_{4}=f\left(x_{4}\right)=\frac{2 x_{4}}{3}=\frac{2 \times 2.584}{3}=1.7226$
Hence, $f_{4}<f_{2}$. Using unimodality we delete the interval $\left[x_{1}, x_{4}\right]$. The remaning interval of uncertainty is $\left[x_{4}, x_{3}\right]$ with $x_{2}$ inside it and near to $x_{4}$.

The fifth experiment is placed at $x_{5}$ given by

$$
\begin{aligned}
& x_{3}-x_{5}=x_{2}-x_{4} \\
\text { or, } & x_{5}=x_{3}-x_{2}+x_{4}=3.292-2.854+2.584=3.022
\end{aligned}
$$

Now $f_{5}=f\left(x_{5}\right)=5-3.022=1.978$
Since $f_{5}>f_{2}$, using unimodal properly we delete the interval $\left[x_{4}, x_{2}\right]$. The remaining interval of uncertainty is $\left[x_{2}, x_{3}\right]$ with $x_{5}$ inside it and near to $x_{2}$.

The sixth experiment is placed at $x_{6}$ given by

$$
x_{3}-x_{6}=x_{5}-x_{2}
$$

or, $x_{6}=x_{3}-x_{5}+x_{2}=3.292-3.022+2.854=3.124$
Now $f_{6}=f\left(x_{6}\right)=5-3.124=1.876$
Since $f_{6}<f_{5}$, using unimodality we delete the interval $\left[x_{6}, x_{3}\right]$. The final interval of uncertainty is given by $\left[x_{2}, x_{6}\right]$ is $[2.854,3 \cdot 124]$

### 7.9. Summary

The necessity of numerical methods of optimization is discussed. The inportance of one-dimensional minimization methods is solving multivariable optimization problems in described. The concept of unimodal function and its role in the elimination
methods is presented. Fibonacei method and Golden section methods are discussed in detail through examples.

### 7.10 Self Assesment Questions

1. Minimize $f(x)= \begin{cases}8-x, & x \leq 4 \\ x, & x \leq 4\end{cases}$
in the interval $[1,7]$ by Fibonacei method using $n=6$
2. Minimize $f(x)=|x-1|$ in the interval $[-1,5]$ by Fibonacei method using $n=5$.
3. Manimize $f(x)= \begin{cases}4 x / 3, & x \leq 3 \\ 7-x, & x \leq 3\end{cases}$
in the interval $[1,5]$ by Golden section method upto six experiments.
4. Minimize $=f(x)= \begin{cases}6-x, & x \leq 5 \\ 2 x-9, & x \leq 5\end{cases}$
in the interval $[2,8]$ by Golden section method upto five experiments.
5. Minimize $=f(x)= \begin{cases}2 \sqrt{x}, & x \leq 1 \\ 3-x, & x \geq 1\end{cases}$
in the interval $[0,5]$ by Golden section method upto six experiments.
6. Minimize $f(x)=|x|$ in the interval $[-2,2]$ by Golden section method upto six experiments.

## Unit 8 - Unconstrained Optimization Technique

## Structure

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8.2 General Iterative Scheme of Optimization
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8.11 Self Assessment Questions
8.1 IntroductionThe solution ofunconstrained optimization problem need not satisfy anyconstraints, Unconstrained optimization technique is important because of thefollowing reasons
(i) Some of the most powerful and convenient methods of solving constrained optimization problems involve the transformation of the problem into one of unconstrained optimization.
(ii) The study of the unconstrained optimization methods provides the basic understanding necessary for the study of the constrained optimization methods.

Several methods are available for solving an unconstrained optimization problem. These methods are classified into two broad categories viz direct search methods and descent methods. The different methods of these two categories are shown below.

### 8.2 General Iterative scheme of optimization

All the unconstrained optimization methods are iterative in nature. Hence they start from an initial trial solution and proceed towards the optimum point in a sequential manner. It is importat tc ote that all the unconstrained optimization methods requires an initial point $x_{1}$ to start the iterative procedure. One method differs from another only in the method of generation the new point $x_{i+1}$ from $x_{\mathrm{i}}$ and in testing the point $x_{i+1}$ for optimality.

If the search a ion frem $x_{i}$ be $s_{i}$ and the step length for movement along the search direction $s_{i}$ be $\lambda_{i}^{*}$ then the next point to $x_{i}$ is obtained as $x_{i+1}=x_{i}+\lambda_{i}^{*} s_{i}$. Thus the terative scheme becomes.
(i) Start with an initial trial point $x_{1}$.
(ii) Find a suitable direction $s_{i}$ ( $i=1$ to start with) which points is general direction of the minimum.
(iii) Find and appropriate step length $\lambda_{i}^{*}$ for moverient along the direction si.
(iv) Obtain the new approximation $x_{i+1}$ as $x_{i+1}=x_{i}+\lambda_{i} s_{i}$
(v) Test whether $x_{i+1}$ is optimurs. If $x_{i+1}$ is optimum then stop the procedure, otherwise set new $i=i+1$ ard repeat step (ii) onward.
Thus as mentioned before, the efficiency of an optimization method depends an the efficiency with which the quantities $\lambda_{i}^{*}$ and si are determined to generate the new point $x_{i+1}$ as $x_{i}+\lambda_{i}^{*} s_{i}$. To find we are to minimize $f\left(x_{i}+\lambda_{i} s_{i}\right)$ regarding it as a function of $\lambda_{i}$ only.

$$
\mathrm{f}\left(\mathrm{x}_{i}+\lambda_{i}^{0} \mathrm{~s}_{\mathrm{i}}\right)=\min _{\lambda_{i}}\left(f\left(x_{i}+\lambda_{i} s_{i}\right)\right.
$$

The flow chart for the iterative scheme may thus be shown as follows


### 8.3 Steepest Descent Method

In the steepest descent method of minimize a function f of n variables $x_{1}, x_{2}$, ....., xn we use the gradient of the the function f defined by

$$
\nabla f=\left[\frac{\partial f}{\partial x_{1}} \frac{\partial f}{\partial x_{2}} \ldots \ldots \frac{\partial f}{\partial x_{n}}\right]^{\gamma}
$$

The gradient of f is a n -companent vector and has a very important property viz if we move along the gradient direction from any point in the n-dimensional space, then the function value increases at the fastest rate. To prove this properly we first define directional derivalive.

Definition 8.3.1 Directional Devivative : The directional devivative of $f(x)$ in the direction of the unit vector y is defined as the following limit

$$
\lim _{t \rightarrow 0} f(x+t y)-f(x)
$$

The directional derivative of $\mathrm{f}(\mathrm{x})$ in the direction y is thus given by using Taylor's theorem

$$
\begin{aligned}
& \lim _{t \rightarrow 0} \frac{\{f(x)+(t y) \nabla f(x)+\text { terms of higher degree in } t\}-f(x)}{t} \\
& =y^{\prime} \nabla f(x)
\end{aligned}
$$

$\therefore$ The directional derivative of $f(x)$ in the direction of unit vector $y$
$=y^{\prime} \nabla f(x)$
$=$ rate of change of $f(x)$ in the direction of $y$.
Theorem 8.3.1 Prove that $f(x)$ increases at the fastest rate in the direction of $\nabla f$.
Proof: We have that the rate of change of $\mathrm{f}(\mathrm{x})$ in the direction of the unit vector y is $\mathrm{y}^{\prime} \nabla f(\mathrm{x})$ - (1)

Now the unit vector in the direction of the gradient vector $\nabla f$ is $\nabla f f|\nabla f|$. Therefore, the rate of change of $f(x)$ in the direction of the gradient vector

$$
\begin{equation*}
=\left(\frac{\nabla f}{|\nabla f|}\right)^{\prime} \nabla f=\frac{(\nabla f)^{\prime}(\nabla f)}{|\nabla f|}=\frac{|\nabla f|^{2}}{|\nabla f|}=|\nabla f| \tag{2}
\end{equation*}
$$

Since $|\nabla f|>0$, it follows the $f(x)$ increases in the direction of $\nabla f$.
Using cauchy schwarz inequality we have

$$
\begin{equation*}
\left|y^{\prime} \nabla f\right| \leq|y||\nabla f|=|\nabla f|[\because|y|=1] . \tag{3}
\end{equation*}
$$

From (1), (2) and (3) it follows that the rate of change of $f(x)$ in the direction of $\nabla f$ is greater than that in the direction of any unit vector $y$. In other words $f(x)$ increases at the fastest rate in the direction of $\nabla f$.

Note : Since $f(x)$ increases at the fastest rate in the direction of $\nabla f$, it follows that $f(x)$ decreases at the fastest rate in the direction of $-\nabla f$. Thus the direction of $\nabla f$ and $-\nabla f$ are respectively the directions of the steepest ascent and steepest descent.

### 8.4 Iterative Scheme of Steepest Descent Method

The steepest descent method uses the properly that a function $f(x)$ decreases a the fastest rate in the direction of $-\nabla f$. Thus at $x_{i}$ the function decreases at the fastes rate along the direction si given by $s_{i}=[-\nabla f]_{x_{i}}=-\nabla f_{i}$.

The iterative scheme of steepest descent method is given below.
(i) Start with an initial point $x_{1}$.
(ii) Take the search direction $s_{\mathrm{i}}$ at $x_{\mathrm{i}}\left(\mathrm{i}=1\right.$ to start with) as $s_{\mathrm{i}}=[-\nabla f]_{x_{i}}$ anc denote it by $-\nabla f_{\mathrm{i}}$.
(iii) Find the step length $\lambda_{i}^{*}$ for movement along si which minimizes $f\left(x_{\mathrm{i}}+\lambda_{i}^{*} s_{\mathrm{i}}\right.$
(iv) Obtain the new appromimation point $x_{i+1}$ as $x_{\mathrm{i}+1}=x_{\mathrm{i}}+\lambda_{i}^{*} s_{\mathrm{i}}$
(v) Test whether $x_{i+1}$ is optimum. If $x_{i+1}$ is optimum then stop the procedure Otherwise set new $i=i+1$ and repeat setp (ii) onward.

### 8.5 Illustrative Examples

Example 8.5.1 Using steepest descent method minimize $f=$ $x_{1}^{2}+x_{2}^{2}+2 g x_{1}+2 f y_{1}+c$ starting from the point

Solution : Here $f\left(x_{1}, x_{2}\right)=x_{1}^{2}+x_{2}^{2}+2 g x_{1}+2 f y_{1}+c$
$\therefore$ The gradient of f is given by

$$
\nabla f=\left[\begin{array}{l}
\frac{\partial f}{\partial x_{1}} \\
\frac{\partial f}{\partial x_{2}}
\end{array}\right]=\left[\begin{array}{l}
2 x_{1}+2 g \\
2 x_{2}+2 f
\end{array}\right]
$$

The starting point is $x_{1}=\left[\begin{array}{l}\alpha \\ \beta\end{array}\right]$. Using steepest desant method the search direction at $x_{1}$ is given by

$$
s_{i}=[-\nabla f]_{x_{i}}=\left[\begin{array}{l}
-2 \alpha-2 g \\
-2 \beta-2 f
\end{array}\right]
$$

The step length $\lambda_{1}$ is obtained by minimising $f\left(x_{1}+\lambda_{1} s_{1}\right)$ with respect to $\lambda_{1}$.
Now $\left(x_{1}+\lambda_{1} s_{1}\right)=\left[\begin{array}{l}\alpha \\ \beta\end{array}\right]+\lambda_{1}\left[\begin{array}{l}-2 \alpha-2 g \\ -2 \beta-2 f\end{array}\right]=\left[\begin{array}{l}\alpha-2 \lambda_{1} \alpha-2 \lambda_{1} g \\ \beta-2 \lambda_{1} \beta-2 \lambda_{1} f\end{array}\right]=\left[\begin{array}{l}\gamma \\ \delta\end{array}\right]$
Where $\gamma=\alpha+\lambda_{1}(-2 \alpha-2 \mathrm{~g})$
and $\delta=\beta+\lambda_{1}(-2 \beta-2 \mathrm{f})$
$\therefore f\left(x_{1}+\lambda_{1} s_{1}\right)=\gamma^{2}+\delta^{2}+2 g \gamma+2 f \delta+c$
For minimum value of f we have $\frac{d f}{d \lambda_{1}}=0$. This gives $\frac{\partial f}{\partial \gamma} \frac{\partial \lambda}{\partial \lambda_{1}}+\frac{\partial f}{\partial \delta} \cdot \frac{\partial \delta}{\partial \lambda_{1}}=0$.
or, $(2 \gamma+2 \mathrm{~g})(-2 \alpha-2 \mathrm{~g})+(2 \delta+2 f)(-2 \beta-2 f)=0$
or $\quad$ ry $-\underline{\underline{g}} \cdot \cdots+\underline{q}+(\delta+f)(\beta+f)=0$
or, $\left\{\alpha+\lambda_{1}(-2 \alpha-2 \mathrm{~g})\right\}(\alpha+\mathrm{g})+\left\{\beta+\lambda_{1}(-2 \beta-2 f)+f \cap(\beta-f)=0\right.$
or, $(\alpha+g)^{2}-2 \lambda_{1}(\alpha+g)^{2}+(\beta+f)^{2}-2 \lambda_{1}(\beta+f)^{2}=0$
or, $\left(1-2 \lambda_{1}\right)\left[(\alpha+\mathrm{g})^{2}+(\beta+f)^{2}\right]=0$
or, $1-2 \lambda_{1}=0$
or, $\lambda_{1}=\frac{1}{2}$
$\therefore \lambda_{1}=\frac{1}{2}$
Now, $x_{1}$ is given by $x_{2}=x_{1}+\lambda_{1}^{*} s_{1}=\left[\begin{array}{l}\alpha \\ \beta\end{array}\right]+\frac{1}{2}\left[\begin{array}{l}-2 \alpha-2 g \\ -2 \beta-2 f\end{array}\right]=\left[\begin{array}{l}\alpha-\alpha-g \\ \beta-\beta-f\end{array}\right]=\left[\begin{array}{l}-g \\ -f\end{array}\right]$
The gradient of f at $x_{2}$ is given by

$$
[\nabla f]_{x_{2}}=\left[\begin{array}{l}
2(-g)+2 g \\
2(-f)+2 f
\end{array}\right]=\left[\begin{array}{l}
0 \\
0
\end{array}\right]
$$

This shows that $x_{2}$ is the optimum point
$\therefore \quad x_{\mathrm{opt}}=x_{2}=\left[\begin{array}{c}-g \\ -f\end{array}\right]$
Example 8.5.2 Using steepest descent method minimize $f\left(x_{1}, x_{2}\right)=x_{1}-x_{2}+$ $2 x_{1}^{2}+2 x_{1} x_{2}+x_{2}^{2}$ starting from the point $\left[\begin{array}{l}0 \\ 0\end{array}\right]$.

Solution : Here $f=f\left(x_{1}, x_{2}\right)=x_{1}-x_{2}+2 x_{1}^{2}+2 x_{1} x_{2}+x_{2}^{2}$ and the starting point is $x_{1}=\left[\begin{array}{l}0 \\ 0\end{array}\right]$.

The gradient of f is given by

$$
\begin{aligned}
& \nabla f=\left[\begin{array}{c}
\frac{\partial f}{\partial x_{1}} \\
\frac{\partial f}{\partial x_{2}}
\end{array}\right]=\left[\begin{array}{c}
1+4 x_{1}+2 x_{2} \\
-1+2 x_{1}+2 x_{2}
\end{array}\right] \\
\therefore & \nabla f_{1}=[\nabla f]_{x_{1}}=\left[\begin{array}{c}
1 \\
-1
\end{array}\right]
\end{aligned}
$$

The search direction at $x_{1}$ is given by $s_{1}=-\nabla f_{1}=\left[\begin{array}{c}-1 \\ 1\end{array}\right]$.
To find $x_{2}$ we are to find the optimal step length $\lambda_{1}$. For this we are to minimize $f\left(x_{1}+\lambda_{1} s_{1}\right)$ with respect to $\lambda_{1}$.

Now $x_{1}+\lambda_{1} s_{1}=-\lambda_{1}-\lambda_{1}+2 \lambda_{1}^{2}-2 \lambda_{1}^{2}+\lambda_{1}^{2}=\lambda_{1}^{2}-2 \lambda_{1}$
For minimum value of f we have $\frac{d f}{d \lambda_{1}}=0$.
From this we have $2 \lambda_{1}-2=0$
or, $\lambda_{1}=1$
$\therefore \lambda_{1}^{*}=1$

Thus we obtain $x_{2}$

$$
x_{2}=x_{1}+\lambda_{1}^{*} s_{1}=\left[\begin{array}{l}
0 \\
0
\end{array}\right]+1\left[\begin{array}{c}
-1 \\
1
\end{array}\right]=\left[\begin{array}{c}
-1 \\
1
\end{array}\right]
$$

The gradient of f at $x_{2}$ is given by

$$
\nabla f_{2}=[\nabla f]_{x_{2}}=\left[\begin{array}{c}
1-4+2 \\
-1-1+2
\end{array}\right]=\left[\begin{array}{l}
-1 \\
-1
\end{array}\right] \neq\binom{ 0}{0}
$$

$\therefore x_{2}$ is not an optimum point. So we proceed to the next iteration.
The search direction at $x_{2}$ is given by

$$
s_{2}=-\nabla f_{2}=\left[\begin{array}{l}
1 \\
1
\end{array}\right]
$$

To find $x_{3}$ we find the step length $\lambda_{3}^{*}$ by minimizing $\mathrm{f}\left(x_{2}+\lambda_{2} s_{2}\right)$ with respect to $\lambda_{2}$.
Now $x_{2}+\lambda_{2} s_{2}=\left[\begin{array}{l}-1 \\ 1\end{array}\right]+\lambda_{2}\left[\begin{array}{l}1 \\ 1\end{array}\right]=\left[\begin{array}{c}-1+\lambda_{2} \\ 1+\lambda_{2}\end{array}\right]$
$\therefore f\left(x_{2}+\lambda_{2} s_{2}\right)=\left(-1+\lambda_{2}\right)\left(1+\lambda_{2}\right)+2\left(-1+\lambda_{2}\right)^{2}+2\left(-1+\lambda_{2}\right)\left(1+\lambda_{2}\right)$

$$
+\left(1+\lambda_{2}\right)^{2}
$$

$$
=-1+\lambda_{2}-1-\lambda_{2}+2-4 \lambda_{2}+2 \lambda_{2}^{2}-2+2 \lambda_{2}^{2}+1+2 \lambda_{2}+\lambda_{2}^{2}
$$

$$
=-1-2 \lambda_{2}+5 \lambda_{2}^{2}
$$

To minimize $f$ we set $\frac{d f}{d \lambda_{2}}=0$
Form this we have $-2+10 \lambda 2=0$
or, $\lambda_{2}=+\frac{1}{5}$
$\therefore \lambda_{2}^{\circ}=\frac{1}{5}$
Hence $x_{3}=x_{2}+\lambda_{2}^{*} s_{2}=\left[\begin{array}{c}-1 \\ 1\end{array}\right]+\frac{1}{5}\left[\begin{array}{l}1 \\ 1\end{array}\right]=\left[\begin{array}{c}-0.8 \\ 1 \cdot 2\end{array}\right]$

The gradient of f at $x_{3}$ is given by

$$
\nabla f_{3}=[\nabla f]_{x 2}=\left[\begin{array}{c}
1+4(-0 \cdot 8)+2(1 \cdot 2) \\
-1+2(-0 \cdot 8)+2(1 \cdot 2)
\end{array}\right]=\left[\begin{array}{c}
0 \cdot 2 \\
-0 \cdot 2
\end{array}\right] \neq\left[\begin{array}{l}
0 \\
0
\end{array}\right]
$$

$\therefore x_{3}$ is not optimum and we proceed to the next iteration.
The search dipection at $x_{3}$ is given by

$$
s_{3}=-\nabla f_{3}=\left[\begin{array}{c}
-0.2 \\
0.2
\end{array}\right]
$$

To find $x_{4}$ we are to find the step length $\lambda_{4}^{*}$ by minimizing $f\left(x_{3}+\lambda_{3} s_{3}\right)$ with respect to $\lambda_{3}$.

$$
\begin{aligned}
& \text { Now } x_{3}+\lambda_{3} s_{3}+\left[\begin{array}{c}
-0.8 \\
1.2
\end{array}\right]+\lambda_{3}\left[\begin{array}{c}
-0.2 \\
0.2
\end{array}\right]=\left[\begin{array}{c}
-0.8-\lambda_{3} 0.2 \\
1.2+\lambda_{3} 0.2
\end{array}\right] \\
& \begin{aligned}
& \therefore f\left(x_{3}+\lambda_{3} s_{3}\right)=\left(-0.8-0.2 \lambda_{3}\right)-\left(1.2+0.2 \lambda_{3}\right)+2\left(-0.8-0.2 \lambda_{3}\right)^{2} \\
&+2\left(-0.8-0.2 \lambda_{3}\right)\left(1.2+0.2 \lambda_{3}\right)+\left(1.2+0.2 \lambda_{3}\right)^{2} \\
& \quad=0.04 \lambda_{3}^{2}-0.08 \lambda_{3}-1.20
\end{aligned}
\end{aligned}
$$

To minimize $f$ we set $\frac{d f}{d \lambda_{3}}=0$
To gives $2 \times 0.04 \lambda_{3}-0.08=0$
or, $\lambda_{3}=1$
$\therefore \quad \lambda_{3}^{*}=1$
Hence $x_{4}=x_{3}+\lambda_{3}^{*} s_{3}=\left[\begin{array}{c}-0.8 \\ 1 \cdot 2\end{array}\right]+1\left[\begin{array}{c}-0 \cdot 2 \\ 0 \cdot 2\end{array}\right]=\left[\begin{array}{c}-1 \cdot 0 \\ 1 \cdot 4\end{array}\right]$
The gradient of f at $x_{4}$ is given by

$$
\nabla f_{4}=[\nabla f]_{x 4}=\left[\begin{array}{c}
1+4(-1 \cdot 0)+2(1 \cdot 4) \\
-1+2(-1 \cdot 0)+2(1 \cdot 4)
\end{array}\right]=\left[\begin{array}{c}
-0 \cdot 20 \\
-0 \cdot 20
\end{array}\right] \neq\left[\begin{array}{l}
0 \\
0
\end{array}\right]
$$

So $x_{4}$ is also not optinum and we are to continue the iterations until we have $\nabla f_{n} \simeq\left[\begin{array}{l}0 \\ 0\end{array}\right]$ and then $x_{\mathrm{n}}$ is taken has the optimum point.

Convergence Creteria: The following criteria can be used to terminate the iteratice process.
(i) $\left|\frac{f\left(x_{i+1}\right)-f\left(x_{i}\right)}{f\left(x_{i}\right)}\right| \leq \epsilon$
(ii) $\left|\frac{\partial f}{\partial x_{i}}\right|<\in$ for all $i=1,2$, n
(iii) $\left|x_{i+1}-x_{i}\right| \leq \epsilon$

### 8.6 Quadratically Convergent Method

Example 8.6.1 A minimization method is called quadratically convergent method if it locates the minimum of general function in no more than a pre-determined number of operations and if the limitting number of operations is directly related to the number of variates.

Definition 8.6.2 Let A be an nxn pymmetric matrix. A set of $n$ vectors $s_{1}$, $s_{2}, \ldots \ldots ., \mathrm{sn}$ is said to be A conjugate directions if $s_{\mathrm{i}}^{T} \mathrm{~A} s_{\mathrm{i}}=0$ for all $\mathrm{i} \neq \mathrm{j}, \mathrm{i}, \mathrm{j}$ $=1,2,3, \ldots . .$. ,

Example 8.6.1 Find the conjugate direction for the symmetric matrix $\left[\begin{array}{cc}2 & -3 \\ -3 & 2\end{array}\right]$
Solution : Let $A=\left[\begin{array}{cc}2 & -3 \\ -3 & 2\end{array}\right]$ and A-conjugate direction be $s_{1}=\left[\begin{array}{l}\alpha \\ \beta\end{array}\right]$ and $s_{2}$ $=\left[\begin{array}{l}\gamma \\ \delta\end{array}\right]$
$\therefore s_{1}^{T} \mathrm{~A} s_{2}=0$
or, $\left[\begin{array}{ll}\alpha & \beta\end{array}\right]\left[\begin{array}{cc}2 & -3 \\ -3 & 2\end{array}\right]\left[\begin{array}{l}\gamma \\ \delta\end{array}\right]=0$
or, $\gamma(2 \alpha-3 \beta)+\delta(-3 \alpha+2 \beta)=0$
Let, $\alpha=1, \beta=2, \gamma=1$

$$
\begin{aligned}
& \therefore-1(2.1-3.2)+\delta(-3.1+2.2)=0 \\
& \text { or, } 4+\delta(+1)=0 \\
& \text { or, } \delta=-4
\end{aligned}
$$

Thus the conjugate direction are $\left[\begin{array}{l}1 \\ 2\end{array}\right]$ and $\left[\begin{array}{l}-1 \\ -4\end{array}\right]$
We note that for a given matrix there are many conjugate directions.
Matrix representation of quadratic expression :
A.ny quadratic expression can be expressed with the help of matrices as

$$
\frac{1}{2} x^{T} \mathrm{~A} x+\mathrm{B}^{T} x+c
$$

Where A is asymmetric matrix
eg. $3 x_{1}^{2}+2 x_{2}^{2}+4 x_{3}^{2}+4 x_{1} x_{2}-x_{2} x_{3}+3 x_{3} x_{1}+3 x_{1}-2 x+x_{3}+7$
can be written as $\frac{1}{2} x^{T} \mathrm{~A} x+\mathrm{B}^{T} x+c$
Where $A=\left[\begin{array}{ccc}6 & 4 & 3 \\ 4 & 4 & -1 \\ 3 & -1 & 8\end{array}\right], B=\left[\begin{array}{c}3 \\ -2 \\ 1\end{array}\right], C-7, X=\left[\begin{array}{c}x_{1} \\ x_{2} \\ x_{3}\end{array}\right]$
We state the following important theorem,
Theorem 8.6.1 If quadratic function $\mathrm{Q}(\mathrm{x})=\frac{1}{2} x^{T} \mathrm{~A} x+\mathrm{B}^{T} x+c$ is minimized sequentially once along each direction of a set of $n$ A-conjugate directions them the global minimum of $\mathrm{Q}(\mathrm{x})$ will be located at a before the n th setp regareless of the starting point and the order in which the directions are used.

### 8.7 Newton's Method

If the function $f(x)$ is continuously differentiable then the local minimum point $x^{*}$ is given by $[\nabla f]_{x^{*}}=0$. Solving the set of $n$ nonlinear equations $\nabla f=0$ we get the optimal point $x^{*}$.

Newton's method : To get the minimum point $x^{*}$ of the continuously differentiable function $f(x)$ we are to solve the n nontinear equation $\nabla f=0$. To solve these n nonlinear equations by the Newton's method, we first linearize the set of equation about the i th appromimations $x_{\mathrm{i}}$ to the minimum point $x^{*}$ of f .

Let $x^{*}=x_{\mathrm{i}}+s$ and $\nabla f=\mathrm{g}$
From $[\nabla f]_{x^{*}}=0$ we have $g\left(x^{*}\right)=0$ or, $g\left(x_{\mathrm{i}}+\mathrm{s}\right)=0$
By Taylor's series expansion we get
$\mathrm{g}\left(\mathrm{x}_{\mathrm{i}}\right)+[\mathrm{J}]_{x_{i}} \mathrm{~s}+\ldots . .=0$ where $[\mathrm{J}]_{x i}$ is the matrix of second partial devivatives of $f$ evaluated at the point Negleeting the higher order terms we get

$$
\mathrm{g}\left(x_{\mathrm{i}}\right)+[\mathrm{J}]_{x_{i}} \mathrm{~s}=0
$$

or, $\mathrm{g}_{\mathrm{i}}+\mathrm{J}_{\mathrm{i}} s=0$ where $\mathrm{g}\left(x_{\mathrm{i}}\right)=\mathrm{gi}$ and $[\mathrm{J}]_{s_{\mathrm{i}}}=\mathrm{J}_{\mathrm{i}}$. If $\mathrm{J}_{\mathrm{i}}$ is non singular, then we have

$$
\mathrm{S}=-\mathrm{J}_{i}^{-1} g_{i}
$$

But the higher order terms are not negligible in general. Hence an iterative procedure has to be used to find the improved approximations. The iterative scheme is given by

$$
x_{i+1}=x_{i}+s_{i}=x_{i}-\mathrm{J}_{i}^{-1} g_{i}
$$

If J is nonsingular then it can be shown that the sequence of points $x_{1}, x_{2}, \ldots \ldots$
$\qquad$ converges to the actual solution $x^{*}$ from any initial point $x_{1}$ sufficiently close to the solution $x^{*}$.

Theorem 8.7.1 If $f(x)$ is a quadratics then the minimum point can be obtained in a single step by Newton's method.

Proof : Let $f(x)=\frac{1}{2} x^{T} \mathrm{~A} x+\mathrm{B}^{T} x+c$ \& the minimum point be $x^{*}$. Then $[\nabla f]_{x^{*}}=0$
or, $[A x+B]_{x^{*}}=0$
or, $A x^{*}+B=0$
or, $x^{*}=-A^{-1} B$.
From $f(x)=\frac{1}{2} x^{T} \mathrm{~A} x ; \mathrm{B}^{T} x+c$ we have $\nabla f=\mathrm{A} x+\mathrm{B}$ and $\mathrm{J}=$ matrix of second partial derivatives of $f=\mathrm{A}$. By Newton's method we have

$$
\begin{aligned}
x_{i+1} & =x_{i}-\mathrm{J}_{i}^{-1} g_{i} \\
& =x_{i}-\mathrm{A}^{-1}\left(\mathrm{~A} x_{i}+\mathrm{B}\right) \\
& \left.=x_{i}-\mathrm{A}^{-1} \mathrm{~A} x_{i}+\mathrm{A}^{-1} \mathrm{~B}\right) \\
& =x_{i}-x_{i}-\mathrm{A}^{-1} \mathrm{~B} \\
& =-\mathrm{A}^{-1} \mathrm{~B}=x^{*}
\end{aligned}
$$

$\therefore x_{2}=-\mathrm{A}^{-1} \mathrm{~B}=x^{*}$ for any starting point $x_{1}$.
Thus the answer is obtained in a single step.
Example 8.7.1 Using Newton's method
minimize $f\left(x_{1}, x_{2}\right)=x_{1}-x_{2}+2 x_{1}^{2}+2 x_{1} x_{2}+2 x_{2}^{2}$ with $\left[\begin{array}{l}0 \\ 0\end{array}\right]$ as starting point.
Solution : Here $f=x_{1}-x_{2}+2 x_{1}^{2}+2 x_{1} x_{2}+2 x_{2}^{2}$

$$
\begin{aligned}
\therefore & \frac{\partial f}{\partial x_{1}}=1+4 x_{1}+2 x_{2}, \quad \frac{\partial f}{\partial x_{2}}=-1+2 x_{1}+2 x_{2} \\
& \frac{\partial^{2} f}{\partial x_{1}^{2}}=4, \frac{\partial^{2} f}{\partial x_{1} \partial x_{2}}=2, \frac{\partial^{2} f}{\partial x_{2}^{2}}=2
\end{aligned}
$$

The starting point is $x_{1}=\left[\begin{array}{l}0 \\ 0\end{array}\right]$

$$
\nabla f_{1}=[\nabla f]_{x_{1}}=\left[\begin{array}{c}
1+0+0 \\
-1+0+0
\end{array}\right]=\left[\begin{array}{c}
1 \\
-1
\end{array}\right]
$$

\& $\mathrm{JI}=\left[\begin{array}{ll}4 & 2 \\ 2 & 2\end{array}\right]$
$\therefore \quad J_{1}^{-1}=\frac{1}{4}\left[\begin{array}{cc}2 & -2 \\ -2 & 4\end{array}\right]=\left[\begin{array}{cc}\frac{1}{2} & -\frac{1}{2} \\ -\frac{1}{2} & 1\end{array}\right]$
We have $x_{1}=x_{2}-J_{1}^{-1} \nabla f_{1}$

$$
=\left[\begin{array}{l}
0 \\
0
\end{array}\right]-\left[\begin{array}{cc}
\frac{1}{2} & -\frac{1}{2} \\
-\frac{1}{2} & 1
\end{array}\right]\left[\begin{array}{c}
1 \\
-1
\end{array}\right]\left[\begin{array}{c}
\frac{1}{2}+\frac{1}{2} \\
-\frac{1}{2}-1
\end{array}\right]=\left[\begin{array}{c}
-1 \\
3 / 2
\end{array}\right]
$$

Now $\nabla f_{2}=[\nabla f]_{s 2}=\left[\begin{array}{c}1+4(-1)+2(3 / 2) \\ -1+2(-1)+2(3 / 2)\end{array}\right]=\left[\begin{array}{l}0 \\ 0\end{array}\right]$
As $\nabla f_{2}=\left[\begin{array}{l}0 \\ 0\end{array}\right], x_{2}=\left[\begin{array}{c}-1 \\ 3 / 2\end{array}\right]$ is the optimum point.

### 8.8 Davidon-Fletcher-Powell Method (Variable Metric Method)

Davidon-Fletcher-Powell method is an important quasi-Newton method. This method is the best general purpose unconstrained optimization technique making use of the derivatioes.

The iterative procedure of this method is as follows :
(i) Start with an initial point $x_{1}$ and a $n \times n$ positive definite symmetric matrix $\mathrm{H}_{1}$. Usually $\mathrm{H}_{1}$ is taken as the identely matrix 1 . Set iteratio number is $\mathrm{i}=1$.
(ii) Compute the gradient of the function $f$ at the point $x_{\mathrm{f}}$ i.e., compute $\nabla f_{1}=[\nabla f]_{x_{1}}$

Take $s_{i}=H_{i} \nabla f_{i}$ as the search direction at $x_{i}$.
(iii) Find the optimal step length $\lambda_{i}^{*}$ in the direction $s_{1}$ and set $x_{i+1}=x_{i}+\lambda_{i} s_{i}$
(iv) Test the new point $x_{i+1}$ for optimality. If $x_{i+1}$ is optimal, terminate the iterative process. Otherwise go to setp (v).
(v) Update $\mathrm{H}_{\mathrm{i}}$ to $\mathrm{H}_{\mathrm{i}+1}$ as

$$
H_{i+1}=H_{i}+M_{i}+N_{i}
$$

Where $\mathrm{M}_{\mathrm{i}}=\left(\lambda_{i} s_{i} s_{i}^{T}\right) /\left(s_{i}^{T} Q_{i}\right)$
$\mathrm{Ni}=-\left(\mathrm{H}_{\mathrm{i}} \mathrm{Q}_{\mathrm{i}}\right)\left(\mathrm{H}_{\mathrm{i}} \mathrm{Q}_{\mathrm{i}}\right)^{\mathrm{T} /}\left(\mathrm{Q}_{i}^{\top} \mathrm{H}_{\mathrm{i}} \mathrm{Q}_{i}\right)$
$Q_{i}=\nabla f_{i+1}-\nabla f_{i}$
(vi) Set the new iteration number $\mathrm{i}=\mathrm{i}+1$ and go to step (ii).

### 8.9 Illustrative Examples

Example 8.9.1 Using Davidon Fletcher-Powell method minimize $f\left(x_{1}, x_{2}\right)=$ $2 x_{1}^{2}+4 x_{2}^{2}-12 x_{1}+16 x_{2}+41$ with $x_{1}=\left[\begin{array}{l}1 \\ 1\end{array}\right]$ as starting point.

Solution : Here $\mathrm{f}=2 x_{1}^{2}+4 x_{2}^{2}-12 x_{1}+16 x_{2}+41$
$\therefore \nabla f=\left[\begin{array}{l}\partial f / \partial x_{1} \\ \partial f / \partial x_{2}\end{array}\right]=\left[\begin{array}{l}4 x_{1}-12 \\ 8 x_{2}+16\end{array}\right]$
Thus $\nabla f_{1}=[\nabla f]_{x_{1}}=\left[\begin{array}{l}4-12 \\ 8+16\end{array}\right]=\left[\begin{array}{l}-8 \\ 24\end{array}\right]$
We take $H_{1}=\left[\begin{array}{ll}1 & 0 \\ 0 & 1\end{array}\right]$
$\therefore s_{1}=-H_{1} \nabla f_{1}=-\left[\begin{array}{ll}1 & 0 \\ 0 & 1\end{array}\right]\left[\begin{array}{l}-8 \\ 24\end{array}\right]=\left[\begin{array}{l}-8 \\ 24\end{array}\right]$
To find the minimizing step length $\lambda_{1}^{*}$ along $s_{1}$, we minimize

$$
\begin{aligned}
f\left(x_{1}\right. & \left.+\lambda_{1} s_{1}\right)=f\left(1+8 \lambda_{1}, 1-24 \lambda_{1}\right) \\
& =2\left(1+8 \lambda_{1}\right)^{2}+4\left(1-24 \lambda_{1}\right)^{2}-12\left(1+8 \lambda_{1}\right)+16\left(1-24 \lambda_{1}\right)+41 \\
& =2+32 \lambda_{1}+128 \lambda_{1}^{2}+4-192 \lambda_{1}+2304 \lambda_{1}^{2}-12-96 \lambda_{1}+16-384 \lambda_{1}+41 \\
& =2432 \lambda_{1}^{2}-640 \lambda_{1}+51
\end{aligned}
$$

We set $\frac{d f}{d \lambda_{i}}=0$
$\therefore 2432 \times 2 \lambda 1-640=0$
or, $\lambda 1=\frac{640}{2 \times 2432}=\frac{10}{76}=0.1316$
$\therefore \lambda_{1}^{*}=0.1316$
$\therefore$ The second approximation is given by

$$
x_{2}=x_{1}+\lambda_{1}^{*} s_{1}=\left[\begin{array}{l}
1 \\
1
\end{array}\right]+0 \cdot 1316\left[\begin{array}{c}
8 \\
-24
\end{array}\right]=\left[\begin{array}{c}
2 \cdot 0528 \\
-2 \cdot 1584
\end{array}\right]
$$

Now $\nabla f_{2}=[\nabla f]_{x 2}=\left[\begin{array}{c}4 \times 2.0528-12 \\ 8 \times(-2.1584)+16\end{array}\right]=\left[\begin{array}{l}-3.7888 \\ -1.2672\end{array}\right] \neq\left[\begin{array}{l}0 \\ 0\end{array}\right]$
$\therefore x_{2}$ is not optimum point
To update the matrix $\mathrm{H}_{1}$ we compute

$$
Q_{1}=\nabla f_{2}-\nabla f_{1}=\left[\begin{array}{l}
-3 \cdot 7888 \\
-1 \cdot 2672
\end{array}\right]-\left[\begin{array}{l}
-8 \\
24
\end{array}\right]=\left[\begin{array}{c}
4 \cdot 2112 \\
-25 \cdot 2672
\end{array}\right]
$$

$$
\begin{aligned}
& \therefore \quad \mathrm{S}_{1}^{T} \mathrm{Q}_{1}=[8-24]\left[\begin{array}{c}
4 \cdot 2112 \\
-25 \cdot 2672
\end{array}\right]=640 \cdot 1024 \\
& S_{1} S_{1}^{r}=\left[\begin{array}{c}
8 \\
-24
\end{array}\right]\left[\begin{array}{ll}
8 & -24
\end{array}\right]=\left[\begin{array}{cc}
64 & -192 \\
-192 & 576
\end{array}\right] \\
& \mathrm{H}_{1} \mathrm{Q}_{1}=\mathrm{Q}_{1}=\left[\begin{array}{c}
4.2112 \\
-25 \cdot 2672
\end{array}\right] \\
& \therefore\left(H_{1} Q_{1}\right)\left(H_{1} \mathrm{Q}_{1}\right)^{\mathrm{T}}=\left[\begin{array}{c}
4 \cdot 2112 \\
-25 \cdot 2672
\end{array}\right]=\{4 \cdot 2112-25 \cdot 2672] \\
& =\left[\begin{array}{cc}
17.7242 & -106.4052 \\
-106.4052 & 638.4314
\end{array}\right] \\
& \text { Also } \mathrm{Q}_{1}^{\gamma}\left(\mathrm{H}_{1} \mathrm{Q}_{1}\right)=\left[\begin{array}{ll}
4 \cdot 2112 & -25 \cdot 2672
\end{array}\right]\left[\begin{array}{c}
4 \cdot 2112 \\
-25 \cdot 2672
\end{array}\right]=656 \cdot 1656 \\
& \therefore N_{1}=-\frac{\left(H_{1} Q_{1}\right)\left(H_{1} \mathrm{Q}_{1}\right)^{T}}{\mathrm{Q}_{1}^{T}\left(\mathrm{H}_{1} \mathrm{Q}_{1}\right)}=-\frac{1}{656 \cdot 1656}\left[\begin{array}{cc}
17.7242 & -106.4052 \\
-106.4052 & 638.4314
\end{array}\right] \\
& =-\left[\begin{array}{cc}
0.027 & -0.1625 \\
-0.1625^{\circ} & 0.973
\end{array}\right] \\
& \mathrm{M}_{1}=\frac{\lambda_{1}^{T} \mathrm{~S}_{1} \mathrm{~S}_{1}^{\tau}}{\mathrm{S}_{1}^{\top} \mathrm{Q}_{1}}=\frac{0.1316}{640 \cdot 1024}\left[\begin{array}{cc}
64 & -192 \\
-192 & 576
\end{array}\right]=\left[\begin{array}{cc}
0.0132 & -0.0395 \\
-0.0395 & 0.1184
\end{array}\right] \\
& \therefore \mathrm{H}_{2}+\mathrm{H}_{1}+\mathrm{M}_{1}+\mathrm{N}_{1} \\
& =\left[\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right]+\left[\begin{array}{cc}
.0 .0132 & -0.0395 \\
-0.0395 & 0.1184
\end{array}\right]+\left[\begin{array}{cc}
-0.027 & 0.1625 \\
0.1625 & -0.973
\end{array}\right]
\end{aligned}
$$

$$
=\left[\begin{array}{cc}
0.8062 & 0.123 \\
0.123 & 0.1454
\end{array}\right]
$$

Hence $\mathrm{S}_{2}=-\mathrm{H}_{2} \nabla f_{2}=-\left[\begin{array}{cc}0.8062 & 0.123 \\ 0.123 & 0.1454\end{array}\right]\left[\begin{array}{l}-3.7888 \\ -1.2672\end{array}\right]=\left[\begin{array}{c}3.21 \\ 0.622\end{array}\right]$
To find the minimizing step length along $S_{2}$ we are to minimize $f\left(x_{2}+\lambda_{2} S_{2}\right)$ $=f\left(3.21 \lambda_{2},-0.9472,0.622 \lambda_{2}-0.1584\right)=2\left(3.21 \lambda_{2}-0.9472\right)^{2}+4\left(0.622 \lambda_{2}-0.1584\right)$
$-12\left(3.21 \lambda_{2}-0.9474\right)+16\left(0.622 \lambda_{2}-0.1584\right)+41$
We set $\frac{d f}{d \lambda_{2}}=0$
This gives $\lambda 2=0.292$

$$
\therefore \quad \lambda_{2}^{*}=0.292
$$

The third approximation is given by

$$
x_{3}=x_{2}+\lambda_{2}^{0} s_{2}=\left[\begin{array}{c}
2 \cdot 0528 \\
-2 \cdot 1584
\end{array}\right]+0 \cdot 292\left[\begin{array}{c}
3 \cdot 21 \\
0 \cdot 622
\end{array}\right]=\left[\begin{array}{c}
2.99 \\
-1 \cdot 98
\end{array}\right]
$$

Now, $\nabla f_{3}=[\nabla f]_{x_{3}}=\left[\begin{array}{c}4 \times 2.99-12 \\ 8 \times(-1.98)+16\end{array}\right]=\left[\begin{array}{c}-0.04 \\ 0.16\end{array}\right] \simeq\left[\begin{array}{l}0 \\ 0\end{array}\right]$
$\therefore x_{3}=\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right]=\left[\begin{array}{c}2.99 \\ -1.98\end{array}\right]$ i.e., $x_{1}=2.99, x_{2}=-1.98$ is the optimum point.

### 8.10 Summary

The unit is devoted to some unconstrained method of optimization viz. steepest descent method, Quadralically convergent method, Newton's method and Dairlon-Fletches-Powell method. These methods are explained with examples.

### 8.11 Self Assessment Questions

1. Using steepest descent method minimize the function $f\left(x_{1}, x_{2}, x_{3}\right)=$ $x_{1}^{2}+x_{2}^{2}+x_{3}^{2}-6 x_{1}-4 x_{1}+3 x_{3}+9$ starting from the point $(1,2,30)$.
2. Using steepest descent method minimize $f\left(x_{1}, x_{2}\right)=2 x_{1}-x_{2}+8 x_{1}^{2}+4 x_{1} x_{2}+$ $x_{2}^{2}$ starting from the point $(0,0)$.
3. Find the conjugate directions for the matrix $\left[\begin{array}{ll}4 & 5 \\ 5 & 4\end{array}\right]$
4. Using Davidon Fletcher and Powell method minimize $f\left(x_{1}, x_{2}\right)=x_{1}-2 x_{2}+2 x_{1}^{2}$ $+4 x_{1} x_{2}+4 x_{2}^{2}$ starting from the point $\left[\begin{array}{l}0 \\ 0\end{array}\right]$
5. Using Davidon-Fletcher Powell method minimize $f\left(x_{1}, x_{2}\right)=8 x_{1}^{2}+4 x_{2}^{2}-24 x_{1}$ $+16 x_{2}+35$ with $\left[\begin{array}{c}1 / 2 \\ 1\end{array}\right]$ as the starting point.
6. Using Davidon-Fletcher Powell method minimize $f\left(x_{1}, x_{2}\right)=2 x_{1}+3 x_{2}+8 x_{1}^{2}$ $+12 x_{1} x_{2}+9 x_{2}^{2}$ with $\left[\begin{array}{l}1 / 2 \\ 1 / 3\end{array}\right]$ as the starting point.

## Unit 9 Constrained Optimization Techniques

## Structure

### 9.1 Introduction

9.2 Cutting Plane Method
9.3 Algorithm of Cutting Plane Method
9.4 Illustrative Examples
9.5 Summary
9.6 Self Assessment Questions

### 9.1 Introduction

The constrained optimization problem is
Minimize $f(x)$
subject to $\mathrm{g}_{\mathrm{j}},(x) \leq 0, \mathrm{j}=1,2, \ldots . ., m$
There are many techniques to solve a constrained non linear programming problem. All these methods canbe classified as follows.

(i) Heuristic search methods
(ii) Methods of feasible directions
(a) Zoutendijlis method
(b) Gradient projection method
(iii) Cutting plane

In the direct methods, the constraints are handled in an explicit manner whereas in most of the indirect methods, the constrained problem is solved as a sequence of unconstrained minimization problems.

In this unit we discuss only cutting plane method.

### 9.2 Cutting Plane Method

In the cutting plane method, the nonlinear constraints are linearized by using Taylor's series expansion thereby approximating the feasible region by linearized envelopes. Assuming that the objective function is linear, we can solve the approximating LPP by this simplex method. If the solution of the LPP is not sufficiently accurate, we relinearize the binding constraints about the current point and formulate a new approximating LPP as solve it using the simplex method. We repeat this procedure until asufficiently accurate solution is found. We note that the approximating linear constraint cut off a portion of the existing feasible region. Hence the method is called cutting plane method.

To apply cutting plane method it is necessary that the objective function is linear. If the objective function is non-linear then we can formulate an equivalent optimization problem with linear objective function as follows.

Let the given problem be
Find $\left(x_{1}, x_{2}, \ldots . . ., x_{n}\right)$ which minimize $f\left(x_{1}, x_{2}, \ldots \ldots, x_{n}\right)$
subject to the constraints $g j\left(x_{1}, x_{2}, \ldots \ldots, x_{n}\right) \leq 0, \mathrm{j}=1,2, \ldots \ldots ., m$.
We introduced a new variable $x_{\mathrm{n}+1}$ änd transform this problem into an equivalent problem as follows

Find $\left(x_{1}, x_{2}, \ldots \ldots ., x_{n}, x_{n+1}\right)$ which minimize $0 x_{1}, 0 x_{2},+\ldots . .+0 x_{n},+x_{n+1}$ subject to the cosntraints $g j\left(x_{1}, x_{2}, \ldots \ldots, x_{n}\right) \leq 0 \mathrm{j}=1,2, \ldots \ldots ., m$ and $g_{m+1}\left(x_{1}, x_{2}, \ldots \ldots .\right.$. , $\left.x_{\mathrm{n}+1}\right)=f\left(x_{1}, x_{2}, \ldots \ldots ., x_{\mathrm{n}}\right)-x_{\mathrm{n}+1} \leq 0$

Thus, without loss of generally, we can assume that the given problem is Minimize $f(x)=f\left(x_{1}, x_{2}, \ldots \ldots, x_{n}\right)=c^{\mathrm{T}} x=c_{1} x_{1}+c_{2} x_{2}+\ldots . .+c_{\mathrm{nxn}}$
subject to the constraints $\mathrm{g}_{\mathrm{j}}(x)=\mathrm{g}_{\mathrm{j}}\left(x_{1}, x_{2}, \ldots ., x_{\mathrm{n}}\right) \leq 0 \mathrm{j}=1,2, \ldots \ldots, \mathrm{~m}$
The iterative procedure of cutting plane method can be stated as follows :

### 9.3 Algorithm of Cutting Plane Method

(i) Start with an initial point $x_{1}$ and set the iteration number as $\mathrm{i}=1$. The point $x_{1}$ need not be feasible
(ii) Linearize the nonlinear constraint functions $\mathrm{g}_{\mathrm{j}}(x)$ about the point $x_{\mathrm{i}}$ as $g_{j}(x) \simeq g_{j}\left(x_{j}\right)+\left[\nabla g_{j}\left(x_{i}\right)\right]^{T}\left(x-x_{\mathrm{i}}\right), \mathrm{j}=1,2, \ldots . ., m$
(iii) Formulate the approximating linear programming problem as Minimize $f(x)=c^{T} x$ subject to $\mathrm{g}_{\mathrm{j}}\left(x_{\mathrm{i}}\right)+\left[\nabla \mathrm{g}_{\mathrm{j}}\left(x_{\mathrm{j}}\right)\right]^{\mathrm{T}}\left(x-x_{\mathrm{i}}\right) \leq 0, \mathrm{j}=1,2, \ldots ., m$
(iv) Solve the approximating LPP to obtain the solution vector $x_{i+1}$.
(v) Evaluate the original constraints at $x_{i+1}$ i.e., find gj $\left(x_{i+1}\right)$ for all $\mathrm{j}=1,2$, ....., $m$.
(vi) If $\mathrm{g}_{\mathrm{j}}\left(x_{\mathrm{i}+1}\right) \leq \in$ for all $\mathrm{j}=1,2, \ldots \ldots, m$ where $\in$ is a prescribed small positive tolerance then all the original constaints can be assumed to have been satisfied.

Hence stop the procedure and take $\mathrm{x}_{\mathrm{opt}}=\mathrm{x}_{\mathrm{i}+1}$
It $. \mathrm{g}_{\mathrm{i}}\left(x_{\mathrm{i}+1}\right)>$ for some value of j , find the most violated constranit as
$g_{k}\left(x_{i+1}\right)=\max \left[g_{j}\left(x_{i+1}\right)\right]$
Relinearize the constrant $\mathrm{g}_{\mathrm{k}}(x) 0$ about the point $x_{\mathrm{i}+1}$ as
$g_{k}(x) \propto g_{k}\left(x_{i+1}\right)+\left[g_{k}\left(x_{i+1}\right)\right]^{\mathrm{T}}\left(x-x_{i+1}\right) \leq 0$
and add this linear constraint to the previous approximating LPP.
(vii) Set the new iteration number $\mathbf{i}=\mathrm{i}+1$ and increase the total number of constraints in thenew approximationg LPP by one and go to step (iv).
Note : To avoid the unbounded solution of the first approximating LPP we may take the first approximating LPP as

Minimize $f(x) c^{\mathrm{T}} x$
subject to $l_{\mathrm{i}} x_{\mathrm{i}} l_{\mathrm{i}}, \mathrm{i}=1,2, \ldots \ldots, \mathrm{n}$
Where $l_{\mathrm{i}}$ and $l_{\mathrm{i}}$ are chosen as lower and upper bounds of $x_{\mathrm{i}}$ take theoptimum solution of this first approximating LPP as $x_{1}$ in this first step.

### 9.4 Illustrative Examples

Example 9.4.1 Using cutting plane method
Maximize $\mathrm{f}\left(x_{1}, x_{2}\right)=7-2 x_{1}-4 x_{2}$
subject to $\left(x_{1}-4\right)^{2}+2\left(x_{2}-3\right)^{2} \leq 12$ taking $\in=0.03$

$$
\begin{aligned}
& x_{1}+2 x_{2} \leq 6 \\
& 1 \leq x_{1} \leq 6 \\
& 1 \leq x_{2} \leq 6
\end{aligned}
$$

Solution : We first consider the LPP
Maximize $f\left(x_{1}, x_{2}\right)=7-2 x_{1}-4 x_{2}$
subject to $x_{1}+2 x_{2} \leq 6$
$1 \leq x_{1} \leq 6$
$1 \leq x_{2} \leq 6$
The extreme point of the feasible region are $\mathrm{A}(1,1), \mathrm{B}(4,1)$ and $\mathrm{C}(1,5 / 2)$.
The value of the objective functions are

$$
(1,1)=1, \quad(4,1)=-5, \quad(1,5 / 2)=-5
$$

$\therefore$ The optimal solution of the LPP is $(1,1)$
$\therefore \quad$ The first approximationg point is $x_{1}=\left[\begin{array}{l}1 \\ 1\end{array}\right]$
Let $\mathrm{g}\left(x_{1}, x_{2}\right)=\left(x_{1}-4\right)^{2}+2\left(x_{2}-3\right)^{2}-12$
$\therefore$ The given non-linear constraint is $\mathrm{g}\left(x_{1}, x_{2}\right) \leq 0$
We gave $g(x)=\left[\begin{array}{l}2\left(x_{1}-4\right) \\ 4\left(x_{2}-3\right)\end{array}\right]$
Now $g\left(x_{1}\right)=g(1,1)(1-4)^{2}+2(1-3)^{2}-12=5>\epsilon=0.03$.

Hence we linearize $\mathrm{g}(x)$ about $x_{1}$ as follows to replace

$$
\begin{aligned}
& \mathrm{g}(x) 0 \text { as } \mathrm{g}\left(x_{1}\right)+\left[\nabla \mathrm{g}\left(x_{1}\right)\right]^{\mathrm{T}}\left[\begin{array}{l}
x_{1}-1 \\
x_{2}-1
\end{array}\right] \leq 0 \\
& \text { or, } 5+[-6,-8]\left[\begin{array}{l}
x_{1}-1 \\
x_{2}-1
\end{array}\right] \leq 0 \\
& \text { or, } 5+(-6)\left(x_{1}-1\right)+(-8)\left(x_{2}-1\right) \leq 0 \\
& \text { or, }-6 x_{1}-8 x_{2}+19 \leq 0 \\
& \text { or, } 6 x_{1}+8 x_{2} \geq 19
\end{aligned}
$$

We now consider the following. LPP by adding the constraint $6 x_{1}+8 x_{2} \geq 19$ as Maximize $f=7-2 x_{1}-4 x_{2}$
subject to $x_{1}+2 x_{2} \leq 6$

$$
\begin{aligned}
& 6 x_{1}+8 x_{2} \geq 19 \\
& 1 \leq x_{1} \leq 6 \\
& 1 \leq x_{2} \leq 6
\end{aligned}
$$

The extreme points of the feasible region are
$A_{1}(1,13 / 8), A_{2}(11 / 6,1), B(4,1)$ and $C(1,5 / 2)$
The values of the objective function are

$$
f(1,13 / 8)=-3 / 2, f(11 / 6,1)=-2 / 3, f(4,1)=-5, f(1,5 / 2)=-5
$$

$\therefore$ The optimal solution of the LPP is

$$
x_{1}=11 / 6, x_{2}=1
$$

$\therefore \quad$ We take the next appromimality point as $x_{2}=\left[\begin{array}{c}11 / 6 \\ 1\end{array}\right]$
Now $g\left(x_{2}\right)=g(11 / 6,1)=\left(\frac{11}{6}-4\right)^{2}+2(1-3)^{2}-12=\frac{25}{36}=0.69>E=0.03$
We relinearize $g(x)$ about $x_{2}$ as follows and consider

$$
\mathrm{g}(x) \leq 0 \text { as } \mathrm{g}\left(x_{2}\right)+\left[\nabla \mathrm{g}\left(x_{2}\right)\right]^{\mathrm{r}}\left[\begin{array}{c}
x_{1}-11 / 6 \\
x_{2}-1
\end{array}\right] \leq 0
$$

$$
\text { or, } \frac{25}{36}+\left[-\frac{13}{3}-8\right]\left[\begin{array}{c}
x_{1}-11 / 6 \\
x_{2}-1
\end{array}\right] \leq 0
$$

$$
\text { or, } \quad 165 x_{1}+288 x_{2} \geq 599
$$

We add this constraint to the previous LPP to get the following LPP

$$
\begin{array}{ll}
\text { Maximize } & f=7-2 x_{1}-4 x_{2} \\
\text { subject to } & x_{1}+2 x_{2} \leq 6 \\
& 6 x_{1}+8 x_{2} \geq 19 \\
& 156 x_{1}+288 x_{2} \geq 599 \\
& 1 \leq x_{1} \leq 6 \\
& 1 \leq x_{2} \leq 6
\end{array}
$$

The extreme points of the feasible region are

$$
\mathrm{A}_{1}(1,13 / 8), \mathrm{B}(17 / 12,21 / 16) \text { and } \mathrm{C}_{1}(311 / 156,1)
$$

The values of the objective function are
$f(1,13 / 8)=-3 / 2, f(17 / 12,21 / 16)=-13 / 12, f\left({ }^{311} / 156,1\right)=-77 / 78$
$\therefore$ The optimum solution is $(311 / 156,1)$
We take $x_{3}=\left[\begin{array}{c}311 / 156 \\ 1\end{array}\right]=\left[\begin{array}{c}1.994 \\ 1\end{array}\right]$
Now $g\left(x_{3}\right)=g(1.994,1)=(1.994-4)^{2}+2(1-3)^{2}-12=0.027<0.03 \in$ Hence, the optimum solution is given by $x_{1}=1.994, x_{2}=1$

### 9.5 Summary

Among all the methods of constrained optimization here we have considered only the cutting plane method. The method is explained with the help of an example.

### 9.6 Self Assessment Questions

Using cutting plane method

$$
\begin{array}{ll}
\text { Maximize } & f=7-2 x_{1}-4 x_{2} \\
\text { subject to } & \left(x_{1}-4\right)^{2}+2\left(x_{2}-3\right)^{2}-12 \geq 0 \\
& x_{1}+2 x_{2}-6 \leq 0 \\
& 1 \leq x_{1}, x_{2} \leq 6
\end{array}
$$

with the tolerance as $\in=0.3$
Using cutting plane method

$$
\begin{array}{ll}
\text { Maximize } & f=1-4 x_{1}-2 x_{2} \\
\text { subject to } & 2\left(x_{1}-2\right)^{2}+\left(x_{2}-3\right)^{2}-12 \geq 0 \\
& 2 x_{1}+x_{2}-3 \leq 0 \\
& 0 \leq x_{1}, x_{2} \leq 5 \\
\text { with } \in=0.2
\end{array}
$$

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> Published by Netaji Subhas Open University, DD-26, Sector-I, Salt Lake, Kolkata - 700064 \& Printed at Gita Printers, 51A, Jhamapukur Lane, Kolkata-700 009.

