Important snaps by Team PIS
Class- Xllth

SUBJECT: MATHEMATICS
BOOK : NCERT
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## Chapter 1 RELATIONS AND FUNCTIONS

## POINTS TO REMEMBER

$\square$ Reflexive relation $R$ in $X$ is a relation with (a, a) $\in R \forall a \in X$.
$\square$ Symmetric relation $R$ in $X$ is a relation satisfying ( $a, b$ ) $\in R$ implies $(b, a) \in R$.

- $\square$ Transitive relation $R$ in $X$ is a relation satisfying ( $a, b$ ) $\in R$ and $(b, c) \in R$
- implies that $(a, c) \in R$.
- $\square \quad$ Equivalence relation $R$ in $X$ is a relation which is reflexive, symmetric and
- transitive.
- $\square$ Equivalence class [a] containing a $\in X$ for an equivalence relation $R$ in $X$ is
- the subset of $X$ containing all elements brelated to a.
- $\square \quad$ A function $f: X \rightarrow Y$ is one-one (or injective) if
- $f(x 1)=f(x 2) \Rightarrow x 1=x 2 \forall x 1, x 2 \in X$.


## Chapter 1 POINTS TO REMEMBER

- A function $f: X \rightarrow Y$ is onto (or surjective) if given any $y \in Y, \exists x \in X$ such
- that $f(x)=y$.
- A function $f: X \rightarrow Y$ is one-one and onto (or bijective), if $f$ is both one-one
- and onto.
- $\square$ The composition of functions $f: A \rightarrow B$ and $g: B \rightarrow C$ is the function
- gof : $A \rightarrow C$ given by gof $(x)=g(f(x)) \forall x \in A$.
- A function $f: X \rightarrow Y$ is invertible if $\exists g: Y \rightarrow X$ such that $g \circ f=I X$ and
- $f \circ g=I Y$.
- A function $f: X \rightarrow Y$ is invertible if and only if $f$ is one-one and onto.


## Chapter-1 POINTS TO REMIEMIBER

- $\square$ Given a finite set $X$, a function $f: X \rightarrow X$ is one-one (respectively onto) if and
- only if $f$ is onto (respectively one-one). This is the characteristic property of a
- finite set. This is not true for infinite set
$\rightarrow \quad \square \quad$ A binary operation $*$ on a set $A$ is a function $*$ from $A \times A$ to $A$.
$\rightarrow \square \quad$ An element $e \in X$ is the identity element for binary operation ${ }^{*}: X \times X \rightarrow X$,
- if $a * e=a=e * a \forall a \in X$.
$\rightarrow \square$ An element $a \in X$ is invertible for binary operation $*: X \times X \rightarrow X$, if
- there exists $b \in X$ such that $a * b=e=b * a$ where, $e$ is the identity for the
- binary operation *. The element $b$ is called inverse of a and is denoted by a-1.
$\rightarrow \quad \square \quad$ An operation $*$ on $X$ is commutative if $a * b=b * a \forall a, b$ in $X$.
$\rightarrow \quad$ An operation $*$ on $X$ is associative if $(a * b) * c=a *(b * c) \forall a, b, c$ in $X$.


## Chapter 2 INVERSE TRIGONOMETRIC FUNCTIONS

The domains and ranges (primcipal value branches) of inverse trigonometric functions are given in the following table:

| Functions | Domain | Range <br> $y=\sin ^{-1} x$ |
| :--- | :---: | :---: |
| $y=\cos ^{-1} x$ | $[-1,1]$ | $\left[\frac{-\pi}{2}, \frac{\pi}{2}\right]$ |
| $y=\operatorname{cosec}^{-1} x$ | $[-1,1]$ | $[0, \pi]$ |
| $y=\sec ^{-1} x$ | $\mathbf{R}-(-1,1)$ | $\left[\frac{-\pi}{2}, \frac{\pi}{2}\right]-\{0\}$ |
| $y=\tan ^{-1} x$ | $\mathbf{R}-(-1,1)$ | $\left.[0, \pi]-\frac{\pi}{2}\right\}$ |
| $y=\cot ^{-1} x$ | $\mathbf{R}$ | $\left(-\frac{\pi}{2}, \frac{\pi}{2}\right)$ |

## Chapter-2 POINTS TO REMEMBER

$$
\begin{aligned}
& y=\sin ^{-1} x \Rightarrow x=\sin y \\
& \sin \left(\sin ^{-1} x\right)=x
\end{aligned}
$$

- $\sin ^{-1} \frac{1}{x}=\operatorname{cosec}^{-1} x$
- $\cos ^{-1} \frac{1}{x}=\sec ^{-1} x$
- $\tan ^{-1} \frac{1}{x}=\cot ^{-1} x$
- $x=\sin y \Rightarrow y=\sin ^{-1} x$
- $\sin ^{-1}(\sin x)=x$
- $\cos ^{-1}(-x)=\pi-\cos ^{-1} x$
- $\cot ^{-1}(-x)=\pi-\cot ^{-1} x$

$$
\sec ^{-1}(-x)=\pi-\sec ^{-1} x
$$

## Chapter 2 POINTS TO REMEMBER

$$
\begin{aligned}
& \sin ^{-1}(-x)=-\sin ^{-1} x \\
& \tan ^{-1} x+\cot ^{-1} x=\frac{\pi}{2} \\
& \sin ^{-1} x+\cos ^{-1} x=\frac{\pi}{2} \\
& \tan ^{-1} x+\tan ^{-1} y=\tan ^{-1} \frac{x+y}{1-x y} \\
& \tan ^{-1} x-\tan ^{-1} y=\tan ^{-1} \frac{x-y}{1+x y} \\
& 2 \tan ^{-1} x=\sin ^{-1} \frac{2 x}{1+x^{2}}=\cos ^{-1} \frac{1-x^{2}}{1+x^{2}}
\end{aligned}
$$

## Chapter- 3 MATRICES

Matrix: It is an ordered rectangular arrangement of numbers (or functions). The numbers (or functions) are called the elements of the matrix. Horizontal line of elements is row of matrix. Vertical line of elements is column of matrix.

Numbers written in the horizontal line form a row of the matrix. Number written in the vertical line form a column of the matrix.

Order of Matrix with ' $m$ ' rows and ' $n$ ' columns is $m \times n$ (read as $m$ by $n$ ).

## Chapter- 3 POINTS TO REMEMBER

## Types of Matrices

A row matrix has only one now (order: $1 \times n$ )

A column matrix has only one column (order: m>1)

A square matrix has number of nows equal to number of columnis (onder: $m \times m$ or $n \times n_{-}$)

A diagonal matrix is a square matrix with all non-diagonal elements equal to zero and diagonal elements not all zeroes.

A scalar matrix is a diagonal matrix in which all diagonal elements ane equal.

An ildentity matrix is a scalar matrix in which each diagonal elenent is 1 (unity).

A zero matrix or mul matrix is the matrix having all elements zero.

## Chapter- 3 POINTS TO REMEMBER

Equal matrices: two matrices $A=\left[a_{i j l m m}\right.$ and $B=\left[b_{i j}\right] m \times n$ are equal if
(a) Both have same onder
(b) $a_{i j}=b_{i j} V$ i and j

## Dperations on matrices

- Two matrices can be added or subtracted, if both have same order.
- If $A=\left[a_{i j}\right]_{m \times n}$ and $B=\left[b_{i j}\right]_{m \times n}$, then
$\mathrm{A} \pm \mathrm{B}=\left[a_{i j} \pm b_{i j}\right]_{m \times m}$
- $\quad \lambda A=\left[\lambda, a_{i j}\right]_{m \times n}$ where $\lambda$ is a scalar
- Two matrices $A$ and $B$ can be multiplied if number of columns in $A$ is equal to number of rows in $B$.

If $A=\left[a_{i j}\right]_{m \times n}$ and $\left[b_{j k} \|_{n \times p}\right.$
Then $A B=\left[c_{i k}\right]_{m \times p}$ where $c_{i k}=\sum_{j=1}^{n} a_{i j} b_{j k}$

## Chapter- 3 POINTS TO REMEMBER


 $\operatorname{by} A^{*} / A^{T}$.


Properties of tramspose matrices A and B are:
(i)

$$
\begin{array}{ll}
\text { (i) } & \left(A^{\prime}\right)^{n}=A \\
\text { (ii) } & (K A)^{n}=k A^{*}(k=\text { constant }) \\
\text { (iii) } & (A+B)^{n}=A^{n}+B^{r} \\
\text { (iv) } & (A B)^{n}=B^{*} A^{r}
\end{array}
$$

Symmmetric Milatrix and Skewu-Symmenetric matrix

 (All diagonal elements are zero in skevw-symmmetric matrix)

## Chapter- 3 POINTS TO REMEMBER

A is a symmetric matrix if $\mathrm{A}^{\prime}=\mathrm{A}$.
A is a skew symmetric matrix if $\mathrm{A}^{\prime}=-\mathrm{A}$.
Any square matrix can be represented as the sum of a symmetric and a skew symmetric matrix.

- Elementary operations of a matrix are as follows:
(i) $\mathrm{R}_{i} \leftrightarrow \mathrm{R}_{j}$ or $\mathrm{C}_{i} \leftrightarrow \mathrm{C}_{j}$
(ii) $\mathrm{R}_{i} \rightarrow k \mathrm{R}_{i}$ or $\mathrm{C}_{i} \rightarrow k \mathrm{C}_{i}$
(iii) $\mathrm{R}_{i} \rightarrow \mathrm{R}_{i}+k \mathrm{R}_{j}$ or $\mathrm{C}_{i} \rightarrow \mathrm{C}_{i}+k \mathrm{C}_{j}$
- If $A$ and $B$ are two square matrices such that $A B=B A=I$, then $B$ is the inverse matrix of $A$ and is denoted by $A^{-1}$ and $A$ is the inverse of $B$.
- Inverse of a square matrix, if it exists, is unique.


## Chapter- 4 DETERMINANTS

Determinant of a matrix $\mathrm{A}=\left[a_{11}\right]_{1 \times 1}$ is given by $\left|a_{11}\right|=a_{11}$
Determinant of a matrix $\mathrm{A}=\left[\begin{array}{ll}a_{11} & a_{12} \\ a_{21} & a_{22}\end{array}\right]$ is given by

$$
|\mathrm{A}|=\left|\begin{array}{ll}
a_{11} & a_{12} \\
a_{21} & a_{22}
\end{array}\right|=a_{11} a_{22}-a_{12} a_{21}
$$

Determinant of a matrix $A=\left[\begin{array}{lll}a_{1} & b_{1} & c_{1} \\ a_{2} & b_{2} & c_{2} \\ a_{3} & b_{3} & c_{3}\end{array}\right]$ is given by (expanding along $R_{1}$ )

$$
|A|=\left|\begin{array}{lll}
a_{1} & b_{1} & c_{1} \\
a_{2} & b_{2} & c_{2} \\
a_{3} & b_{3} & c_{3}
\end{array}\right|=a_{1}\left|\begin{array}{ll}
b_{2} & c_{2} \\
b_{3} & c_{3}
\end{array}\right|-b_{1}\left|\begin{array}{ll}
a_{2} & c_{2} \\
a_{3} & c_{3}
\end{array}\right|+c_{1}\left|\begin{array}{ll}
a_{2} & b_{2} \\
a_{3} & b_{3}
\end{array}\right|
$$

## Chapter- 4 POINTS TO REMEMBER

For any square matrix $A$, the $|A|$ satisfy following properties.

- $\left|A^{\prime}\right|=\| A \mid$, where $A^{\prime}=$ transpose of $A$.
- If we interchange any two rows (or columns), then sign of determinant changes.
- If any two rows or any two columns are identical or proportional, then value of determinant is zero.
- If we multiply each element of a row or a column of a deteminant by constant $\hbar$, then value of determinant is multiplied by $k$.
- Multiplying a determinant by $k$ means multiply elements of only one row (or one column) by $k$.
- If $A=\left[a_{i j}\right]_{3 \times 3}$, then $|k \cdot A|=k^{3}|A|$
- If elements of a row or a column in a determinant can be expressed as sum of two or more elements, then the given determinant can be expressed as sum of two or more determinants.
- If to each element of a row or a column of a determinant the equimultiples of corresponding elements of other rows or columns are added, then value of determinant remains same.


## Chapter- 4 POINTS TO REMEMBER

Area of a triangle with vertices $\left(x_{1}, y_{1}\right),\left(x_{2}, y_{2}\right)$ and $\left(x_{3}, y_{3}\right)$ is given by

$$
\Delta=\frac{1}{2}\left|\begin{array}{lll}
x_{1} & y_{1} & 1 \\
x_{2} & y_{2} & 1 \\
x_{3} & y_{3} & 1
\end{array}\right|
$$

Minor of an element $a_{i j}$ of the determinant of matrix A is the determinant obtained by deleting $i^{\text {th }}$ row and $j^{\text {th }}$ column and denoted by $\mathrm{M}_{i j}$.
Cofactor of $a_{i j}$ of given by $\mathrm{A}_{i j}=(-1)^{i+j} \mathrm{M}_{i j}$
Value of determinant of a matrix $A$ is obtained by sum of product of elements of a row (or a column) with corresponding cofactors. For example,

$$
|\mathrm{A}|=a_{11} \mathrm{~A}_{11}+a_{12} \mathrm{~A}_{12}+a_{13} \mathrm{~A}_{13}
$$

If elements of one row (or column) are multiplied with cofactors of elements of any other row (or column), then their sum is zero. For example, $a_{11} \mathrm{~A}_{21}+a_{12}$ $\mathrm{A}_{22}+a_{13} \mathrm{~A}_{23}=0$

## Chapter-4 POINTS TO REMEMBER

- If $A=\left[\begin{array}{lll}a_{11} & a_{12} & a_{13} \\ a_{21} & a_{22} & a_{23} \\ a_{31} & a_{32} & a_{33}\end{array}\right]$, then $a d j A=\left[\begin{array}{lll}A_{11} & A_{21} & A_{31} \\ A_{12} & A_{22} & A_{32} \\ A_{13} & A_{23} & A_{33}\end{array}\right]$, where $A_{i j}$ is cofactor of $a_{i j}$
- $A(a d j A)=(a d j A) A=|A| I$, where $A$ is square matrix of order $n$.
- A square matrix $A$ is said to be singular or mon-singular according as $|A|=0$ or $|A| \neq 0$.
If $A B=B A=I$, where $B$ is square nomatrix, then B is called inverse of $A$. Also $A^{-1}=B$ or $B^{-1}=A$ and hence $\left(A^{-1}\right)^{-1}=A$.
A square matrix $A$ has inverse if and only if $A$ is non-singular.
$A^{-1}=\frac{1}{|A|}(\operatorname{adj} A)$
If $\quad a_{1} x+b_{1} y+c_{1}==d_{1}$ $a_{2} x+b_{2} y+c_{2}==d_{2}$
$a_{3} x+b_{3} y+c_{3}==d_{3}$,
then these equations can be written as $A X=B$, where
$A=\left[\begin{array}{lll}a_{1} & b_{1} & c_{1} \\ a_{2} & b_{2} & c_{2} \\ d_{3} & b_{3} & c_{3}\end{array}\right], X=\left[\begin{array}{l}x \\ y \\ z\end{array}\right]$ and $B=\left[\begin{array}{l}d_{1} \\ d_{2} \\ d_{3}\end{array}\right]$


## Chapter-5 CONTINUITY AND DIFFERENTIABILITY

A real valued function is continuous at a point in its domain if the limit of the function at that point equals the value of the function at that point. A function is continuous if it is continuous on the whole of its domain.
Sum, difference, product and quotient of continuous functions are continuous. i.e., if $f$ and $g$ are continuous functions, then
$(f \pm g)(x)=f(x) \pm g(x)$ is continuous.
$(f \cdot g)(x)=f(x) \cdot g(x)$ is continuous.
$\left(\frac{f}{g}\right)(x)=\frac{f(x)}{g(x)}$ (wherever $g(x) \neq 0$ ) is continuous.
Every differentiable function is continuous, but the converse is not true.

## Chapter-5 POINTS TO REMEMBER

Chain rule is rule to differentiate composites of functions. If $f=v=u, t=u(x)$ and if both $\frac{d t}{d x}$ and $\frac{d v}{d t}$ exist then

$$
\frac{d f}{d x}=\frac{d v}{d t} \cdot \frac{d t}{d x}
$$

- Following are some of the standard derivatives (in appropriate domains):

$$
\begin{array}{ll}
\frac{d}{d x}\left(\sin ^{-1} x\right)=\frac{1}{\sqrt{1-x^{2}}} & \frac{d}{d x}\left(\cos ^{-1} x\right)=-\frac{1}{\sqrt{1-x^{2}}} \\
\frac{d}{d x}\left(\tan ^{-1} x\right)=\frac{1}{1+x^{2}} & \frac{d}{d x}\left(\cot ^{-1} x\right)=\frac{-1}{1+x^{2}} \\
\frac{d}{d x}\left(\sec ^{-1} x\right)=\frac{1}{x \sqrt{1-x^{2}}} & \frac{d}{d x}\left(\operatorname{cosec}^{-1} x\right)=\frac{-1}{x \sqrt{1-x^{2}}} \\
\frac{d}{d x}\left(e^{x}\right)=e^{x} & \frac{d}{d x}(\log x)=\frac{1}{x}
\end{array}
$$

## Chapter-5 POINTS TO REMEMBER

Logarithmic differentiation is a powerful technique to differentiate functions of the form $f(x)=[u(x)]^{v(x)}$. Here both $f(x)$ and $u(x)$ need to be positive for this technique to make sense.
$\bullet$ Rolle's Theorem: If $f:[a, b] \rightarrow \mathbf{R}$ is continuous on $[a, b]$ and differentiable on $(a, b)$ such that $f(a)=f(b)$, then there exists some $c$ in $(a, b)$ such that $f^{\prime}(c)=0$.
Mean Value Theorem: If $f:[a, b] \rightarrow \mathbf{R}$ is continuous on $[a, b]$ and differentiable on $(a, b)$. Then there exists some $c$ in $(a, b)$ such that

$$
f^{\prime}(c)=\frac{f(b)-f(a)}{b-a}
$$

## Chapter-6 APPLICATION OF DERIVATIVES

If a quantity $y$ varies with another quantity $x$, satisfying some rule $y=f(x)$, then $\frac{d y}{d x}$ (or $\left.f^{\prime}(x)\right)$ represents the rate of change of $y$ with respect to $x$ and $\left.\frac{d y}{d x}\right]_{x=x_{0}}$ (or $f^{\prime}\left(x_{0}\right)$ ) represents the rate of change of $y$ with respect to $x$ at $x=x_{0}$.

- If two variables $x$ and $y$ are varying with respect to another variable $t$, i.e., if $x=f(t)$ and $y=g(t)$, then by Chain Rule

$$
\frac{d y}{d x}=\frac{d y}{d t} / \frac{d x}{d t}, \text { if } \frac{d x}{d t} \neq 0
$$

- A function $f$ is said to be
(a) increasing on an interval $(a, b)$ if

$$
x_{1}<x_{2} \text { in }(a, b) \Longrightarrow f\left(x_{1}\right) \leq f\left(x_{2}\right) \text { for all } x_{1}, x_{2} \in(a, b)
$$

## Chapter-6 POINTS TO REMEMBER

Alternatively, if $f^{\prime}(x) \geq 0$ for each $x$ in $(a, b)$
(b) decreasing on $(a, b)$ if

$$
x_{1}<x_{2} \text { in }(a, b) \Longrightarrow f\left(x_{1}\right) \geq f\left(x_{2}\right) \text { for all } x_{1}, x_{2} \in(a, b)
$$

Alternatively, if $f^{\prime}(x) \leq 0$ for each $x$ in $(a, b)$
The equation of the tangent at $\left(x_{0}, y_{0}\right)$ to the curve $y=f(x)$ is given by

$$
\left.y-y_{0}=\frac{d y}{d x}\right]_{\left(x_{0}, y_{0}\right)}\left(x-x_{0}\right)
$$

- If $\frac{d y}{d x}$ does not exist at the point $\left(x_{0}, y_{0}\right)$, then the tangent at this point is parallel to the y -axis and its equation is $x=x_{\mathrm{o}}$.
- If tangent to a curve $y=f(x)$ at $x=x_{0}$ is parallel to $x$-axis, then $\left.\frac{d y}{d x}\right]_{x=x_{0}}=0$.


## Chapter-6 POINTS TO REMEMBER

Equation of the nominal to the curve $y=f(x)$ at a point $\left(x_{0}, y_{0}\right)$ is given by

$$
y-y_{0}=\frac{-1}{\left.\frac{d y}{d x}\right]_{\left(x_{0}-y_{0}\right)}}\left(x-x_{0}\right)
$$

- If $\frac{d y}{d x}$ at the point $\left(x_{0}, y_{0}\right)$ is zero, then equation of the nommal is $x=x_{0}$.

If $\frac{d y}{d x}$ at the point $\left(x_{0}, y_{0}\right)$ does not exist, then the nommal is parallel to $x$-axis and its equation is $y=y_{0}$ -
Let $y=f(x)$, $\Delta x$ be a small incrennent in $x$ and $\Delta y$ be the incrennent in $y$ corresponding to the incrennent in $x$, i.e. $\Delta y=\mathscr{f}(x+\Delta x)-\mathscr{f}(x)$. Then $d y$ given by

$$
d y=f^{\prime}(x) d x \text { or } d y=\left(\frac{d y}{d x}\right) \Delta x \text {. }
$$

is a goad approximation of $\Delta y$ when $d x=\Delta x$ is melatively snnall and we denote it by $d y=\Delta y$.
A point $c$ in the donnain of a function $f^{\prime}$ at which either $f^{\prime}(C)=0$ or $\mathcal{f}$ is not differentiable is called a critical point off.

## Chapter-6 POINTS TO REMEMBER

First Derivative Test Let $f$ be a function defined on an open interval I. Let $f$ be continuous at a critical point $c$ in I. Then
(i) If $f^{\prime}(x)$ changes sign from positive to negative as $x$ increases through c , i.e., if $f^{\prime}(x)>0$ at every point sufficiently close to and to the left of $c$, and $f^{\prime}(x)<0$ at every point sufficiently close to and to the right of $c$, then $c$ is a point of local maxima.
(ii) If $f^{\prime}(x)$ changes sign from negative to positive as $x$ increases through $c$, i.e., if $f^{\prime}(x)<0$ at every point sufficiently close to and to the left of $c$, and $f^{\prime}(x)>0$ at every point sufficiently close to and to the right of $c$, then $c$ is a point of local minima.
(iii) If $f^{\prime}(x)$ does not change sign as $x$ increases through $c$, then $c$ is neither a point of local maxima nor a point of local minima. Infact, such a point is called point of inflexion.

## Chapter-6POINTS TO REMEMBER

Second Derivative Test Let $f$ be a function defined on an interval I and $c \in$ I. Let $f$ be twice differentiable at $c$. Then
(i) $x=c$ is a point of local maxima if $f^{\prime}(c)=0$ and $f^{\prime \prime}(c)<0$ The values $f(c)$ is local maximum value of $f$.
(ii) $x=c$ is a point of local minima if $f^{\prime}(c)=0$ and $f^{\prime \prime}(c)>0$

In this case, $f(c)$ is local minimum value of $f$.
(iii) The test fails if $f^{\prime}(c)=0$ and $f^{\prime \prime}(c)=0$.

In this case, we go back to the first derivative test and find whether $c$ is a point of maxima, minima or a point of inflexion.
Working rule for finding absolute maxima and/or absolute minima
Step 1: Find all critical points of $f$ in the interval, i.e., find points $x$ where either $f^{\prime}(x)=0$ or $f$ is not differentiable.
Step 2:Take the end points of the interval.
Step 3: At all these points (listed in Step 1 and 2), calculate the values of $f$.
Step 4: Identify the maximum and minimum values of $f$ out of the values calculated in Step 3. This maximum value will be the absolute maximum value of $f$ and the minimum value will be the absolute minimum value of $f$.

## Chapter-7 INTEGRALS

Integration is the inverse process of differentiation. In the differential calculus, we are given a function and we have to find the derivative or differential of this function, but in the integral calculus, we are to find a function whose differential is given. Thus, integration is a process which is the inverse of differentiation.

Let $\frac{d}{d x} \mathrm{~F}(x)=f(x)$. Then we write $\int f(x) d x=\mathrm{F}(x)+\mathrm{C}$. These integrals are called indefinite integrals or general integrals, C is called constant of integration. All these integrals differ by a constant.
From the geometric point of view, an indefinite integral is collection of family of curves, each of which is obtained by translating one of the curves parallel to itself upwards or downwards along the $y$-axis.

## Chapter-7 POINTS TO REMEMBER

Some properties of indefinite integrals are as follows:

1. $\int[f(x)+g(x)] d x=\int f(x) d x+\int g(x) d x$
2. For any real number $k, \int k f(x) d x=k \int f(x) d x$

More generally, if $f_{1}, f_{2}, f_{3}, \ldots, f_{n}$ are functions and $k_{1}, k_{2}, \ldots, k_{n}$ are real numbers. Then

$$
\begin{aligned}
\int\left[k_{1} f_{1}(x)+k_{2} f_{2}(x)\right. & \left.+\ldots+k_{n} f_{n}(x)\right] d x \\
& =k_{1} \int f_{1}(x) d x+k_{2} \int f_{2}(x) d x+\ldots+k_{n} \int f_{n}(x) d x
\end{aligned}
$$

## Chapter-7 POINTS TO REMEMBER

Somme stamclanil integrals
(i) $\int x^{n} d x=\frac{x^{n+1}}{n+1}+C, n \neq-1$. Particularly; $\int d x=x+C$
(ii) $\int \cos x d x=\sin x+C \quad$ (iii) $\int \sin x d x=-\cos x+C$
(iv) $\int \sec ^{2} x d x=\tan x+C \quad$ (v) $\int \operatorname{cosec}^{2} x d x=-\cot x+C$
(vi) $\int \sec x \tan x d x=\sec x+C$
(vii) $\int \operatorname{cosec} x \cot x d x=-\operatorname{cosec} x+C$ (viii) $\int \frac{d x}{\sqrt{1-x^{2}}}=\sin ^{-1} x+C$
(ix) $\int \frac{d x}{\sqrt{1-x^{2}}}=-\cos ^{-1} x+C$
(x) $\int \frac{d x}{1+x^{2}}=\tan ^{-1} x+C$
(xi) $\int \frac{d x}{1+x^{2}}=-\cot ^{-1} x+C$
(xii) $\int e^{x} d x=e^{x}+C$
(xiii) $\int a^{x} d x=\frac{a^{x}}{\log a}+C$
(xiv) $\int \frac{d x}{x \sqrt{x^{2}-1}}=\sec ^{-1} x+C$
(xv) $\int \frac{d x}{x \sqrt{x^{2}-1}}=-\operatorname{cosec}^{-1} x+C$
(xvi) $\int \frac{1}{x} d x=\log |x|+C$

## Chapter-7 POINTS TO REMEMBER

1. 

$\frac{p x+a}{(x-a)(x-b)}$
2. $\quad \frac{p x+a}{(x-a)^{2}}$
$=$
$\frac{A}{x-a}+\frac{B}{x-b}=a \neq b$

$$
\frac{p x^{2}+g x+x}{(x-a)(x-b)(x-c)}
$$

$=$
$\frac{A}{x-a}+\frac{B}{(x-a)^{2}}$

3
4. $\quad \frac{p x^{2}+a x+1}{(x-a)^{2}(x-b)}$
5. $\quad \frac{p x^{2}+x+x+1}{(x-a)\left(x^{2}+b x+c\right)}$

$$
=\quad \frac{A}{x-a}+\frac{B}{(x-a)^{2}}+\frac{C}{x-b}
$$

where $x^{2}+b x+c$ can mot be factorised further.
Indegriation byy sublastituntion
A change in the varialole of integration often meduces am intepral to one of the fumdammental imteprals. The mofthorl in which wre champe the vaniable to somme other vamiable is called the mnethocl of subastitution. Whem the inteprand involves some trigomometric fumctioms, whe use somme well kmown idemitites to find the inteprals. Usinp subistitution teclumique, we obitiv the followvimp stancland imiteprals.
(i) $\int \tan x d x=10 g|\sec x|+c \quad$ (ii) $\int \cot x d x=10 \sin x \mid+c$
(iii) $\int \sec x d x=1 \log |\sec x+\tan x|+C$
(iv) $\int \operatorname{cosec} x d x=10 \log |\operatorname{cosec} x-\cot x|+c$

## Chapter-7 POINTS TO REMEMBER

$$
\begin{aligned}
& \int \frac{1}{x^{2}-a^{2}} d x=\frac{1}{2 a} \log \left|\frac{x-a}{x+a}\right|+c \\
& \int \frac{1}{a^{2}+x^{2}} d x=\frac{1}{a} \tan -1 \frac{x}{a}+c \\
& \int \frac{1}{\sqrt{a^{2}-x^{2}}} d x=\sin -1 \frac{x}{a}+c=-\cos -1 \frac{x}{a}+c \\
& \int \frac{1}{\sqrt{a^{2}+x^{2}}} d x=\log \left|x+\sqrt{a^{2}+x^{2}}\right|+c \\
& \int \frac{1}{\sqrt{x^{2}-a^{2}}} d x=\log \left|x+\sqrt{x^{2}-a^{2}}\right|+c \\
& \int \sqrt{a^{2}-x^{2}} d x=\frac{x}{2} \sqrt{a^{2}-x^{2}}+\frac{a^{2}}{2} \sin -1 \frac{x}{a}+c \\
& \int \sqrt{a^{2}+x^{2}} d x=\frac{x}{2} \sqrt{a^{2}+x^{2}}+\frac{a^{2}}{2} \log \left|x+\sqrt{a^{2}+x^{2}}\right|+c \\
& \left.\int \sqrt{x^{2}-a^{2}} d x=\frac{x}{2} \sqrt{x^{2}-a^{2}}-\frac{a^{2}}{2} \log \right\rvert\, x+\sqrt{x^{2}-a^{2}}+c
\end{aligned}
$$

## Chapter-7 POINTS TO REMEMBER

## First fundamental theorem of integral calculus

Let the area function be defined by $\mathrm{A}(x)=\int_{a}^{x} f(x) d x$ for all $x \geq a$, where the function $f$ is assumed to be continuous on $[a, b]$. Then $\mathrm{A}^{\prime}(x)=f(x)$ for all $x \in[a, b]$.

- Second fundamental theorem of integral calculus

Let $f$ be a continuous function of $x$ defined on the closed interval $[a, b]$ and let F be another function such that $\frac{d}{d x} \mathrm{~F}(x)=f(x)$ for all $x$ in the domain of
$f$, then $\int_{a}^{b} f(x) d x=[\mathrm{F}(x)+\mathrm{C}]_{a}^{b}=\mathrm{F}(b)-\mathrm{F}(a)$.
This is called the definite integral of $f$ over the range $[a, b]$, where $a$ and $b$ are called the limits of integration, $a$ being the lower limit and $b$ the upper limit.

## Chapter-8 APPLICATION OF INTEGRALS

- The area of the region bounded by the curve $y=f(x), x$-xxis and the lines $x=a$ and $x=b(b>a)$ is given by the formula: Area $=\int_{a}^{b} y d x=\int_{a}^{b} f(x) d x$.
- The area of the region bounded by the curve $x=\phi(y), y$-axis and the lines
$y=c, y=d$ is given by the formula: Area $=\int_{c}^{d} x d y=\int_{c}^{d} \phi(y) d y$.


## Chapter-8 POINTS TO REMEMBER

The area of the region enclosed between two curves $y=f(x), y=g(x)$ and the lines $x=a, x=b$ is given by the formula,

Area $=\int_{a}^{b}[f(x)-g(x)] d x$, where, $f(x) \geq g(x)$ in $[a, b]$

- If $f(x) \geq g(x)$ in $[a, c]$ and $f(x) \leq g(x)$ in $[c, b], a<c<b$, then Area $=\int_{a}^{c}[f(x)-g(x)] d x+\int_{c}^{b}[g(x)-f(x)] d x$.


## Chapter-9 DIFFERENTIAL EQUATIONS

- Differential Equation: Equation containing derivatives of a dependent variable with respect to an independent variable is called differential equation.
- Order of a Differential Equation: The order of a differential equation is defined to be the order of the highest order derivative occurring in the differential equation.
- Degree of a Differential Equation: Highest power of highest order derivative involved in the equation is called degree of differential equation where equation is a polynomial equation in differential coefficients.
- Formation of a Differential Equation: We differentiate the family of curves as many times as the number of arbitrary constant in the given family of curves. Now eliminate the arbitrary constants from these equations.

After elimination, the equation obtained is differential equation.

## Chapter-9 POINTS TO REMEMBER

## Solution of Differential Equation

(i) Variable Separable Method

$$
\frac{d y}{d x}=f(x, y) .
$$

We Separate the variables and get

$$
f(x) d x=g(y) d y
$$

Then $\int f(x) d x=\int g(y) d y+c$ is the required solutions.

## Chapter-9 POINTS TO REMEMBER

(ii) Homogeneous Differential Equation: A differential equation of the form $\frac{d y}{d x}=\frac{f(x, y)}{g(x, y)}$ where $f(x, y)$ and $g(x, y)$ are both homogeneous functions of the same degree in $x$ and $y$ i.e., of the form $\frac{d y}{d x}=F\left(\frac{y}{x}\right)$ is called a homogeneous differential equation.

For solving this type of equations we substitute $y=v x$ and then $\frac{d y}{d x}=v+x \frac{d v}{d x}$. The equation can be solved by variables separable method.

A homogeneous differential equation can be of the form $\frac{d x}{d y}=F\left(\frac{x}{y}\right)$
To solve this equation, we substitute $x=v y$ and them $\frac{d x}{d y}=v+y \frac{d v}{d y}$ then the equation can be solved by variable separate method.

## Chapter-9 POINTS TO REMEMBER

(iii)

Linear Differential Equation: An equation of the from $\frac{d y}{d x}+P y=Q$
where $P$ and $Q$ are constant or functions of $x$ only is called a linear differential equation. For finding solution of this type of equations, we find integrating factor $\left(I . F_{.}\right)=e^{\int \mathrm{Pdx}}$

Solution is $y($ I.F. $)=\int Q$. (I. F.) $d x+c$
Similarly, differential equations of the type $\frac{d x}{d y}+P x=Q$ where P and Q are constants or functions of $y$ only can be solved.

Here, I.F. $=\mathrm{e}^{\int \text { Pdy }}$ and the solution is $\times($ I.F. $)=\int \mathrm{Q} \times(\mathrm{I}$. F. $) \mathrm{dy}+\mathrm{C}$

## Chapter-10 VECTOR ALGEBRA

Position vector of a point $\mathrm{P}(x, y, z)$ is given as $\overrightarrow{\mathrm{OP}}(=\vec{r})=x \hat{i}+y \hat{j}+z \hat{k}$, and its magnitude by $\sqrt{x^{2}+y^{2}+z^{2}}$.

- The scalar components of a vector are its direction ratios, and represent its projections along the respective axes.
The magnitude $(r)$, direction ratios $(a, b, c)$ and direction cosines $(l, m, n)$ of any vector are related as:

$$
l=\frac{a}{r}, \quad m=\frac{b}{r}, \quad n=\frac{c}{r}
$$

The vector sum of the three sides of a triangle taken in order is $\overrightarrow{0}$.

## Chapter-10 POINTS TO REMEMBER

The vector sum of two coinitial vectors is given by the diagonal of the parallelogram whose adjacent sides are the given vectors.
The multiplication of a given vector by a scalar $\lambda$, changes the magnitude of the vector by the multiple $|\lambda|$, and keeps the direction same (or makes it opposite) according as the value of $\lambda$ is positive (or negative).

For a given vector $\vec{a}$, the vector $\hat{a}=\frac{\vec{a}}{|\vec{a}|}$ gives the unit vector in the direction of $\vec{a}$.
The position vector of a point $R$ dividing a line segment joining the points $P$ and $Q$ whose position vectors are $\vec{a}$ and $\vec{b}$ respectively, in the ratio $m: n$
(i) internally, is given by $\frac{n \vec{a}+m \vec{b}}{m+n}$.
(ii) externally, is given by $\frac{m \vec{b}-n \vec{a}}{m-n}$.

## Chapter-10 POINTS TO REMEMBER

The scalar product of two given vectors $\vec{a}$ and $\vec{b}$ having angle $\theta$ between them is defined as

$$
\vec{a} \cdot \vec{b}=|\vec{a}||\vec{b}| \cos \theta
$$

Also, when $\vec{a} \cdot \vec{b}$ is given, the angle ' $\theta$ ' between the vectors $\vec{a}$ and $\vec{b}$ nay be determined by

$$
\cos \theta=\frac{\vec{a} \cdot \vec{b}}{|\vec{a}||\vec{b}|}
$$

- If $\Theta$ is the angle between two vectors $\vec{a}$ and $\vec{b}$, then their cross product is given as

$$
\vec{a} \times \vec{b}=|\vec{a} \| \vec{b}| \sin \theta \hat{n}
$$

Where $\hat{n}$ is a unit vector perpendicular to the plane containing $\vec{a}$ and $\vec{b}$. Such that $\vec{a}, \vec{b}, \hat{n}$ form right handed system of coordinate axes.

- If we have two vectors $\vec{a}$ and $\vec{b}$, given in component form as $\bar{a}=a_{1} \hat{i}+a_{2} \hat{j}+a_{3} \hat{k}$ and $\hat{b}=b_{1} \hat{i}+b_{2} \hat{j}+b_{3} \hat{k}$ and $\lambda$ any scalar,


## Chapter-10 POINTS TO REMEMBER

$$
\begin{aligned}
\vec{a}+\vec{b} & =\left(a_{1}+b_{1}\right) \hat{i}+\left(a_{2}+b_{2}\right) \hat{j}+\left(a_{3}+b_{3}\right) \hat{k} \\
\lambda \vec{a} & =\left(\lambda a_{1}\right) \hat{i}+\left(\lambda a_{2}\right) \hat{j}+\left(\lambda a_{3}\right) \hat{k} \\
\vec{a} \cdot \vec{b} & =a_{1} b_{1}+a_{2} b_{2}+a_{3} b_{3} \\
\vec{a} \times \vec{b} & =\left|\begin{array}{lll}
\hat{i} & \hat{j} & \hat{k} \\
a_{1} & b_{1} & c_{1} \\
a_{2} & b_{2} & c_{2}
\end{array}\right|
\end{aligned}
$$

## Chapter-1 1THREE DIMENSIONAL GEOMETRY

Direction cosines of a line are the cosines of the angles made by the line with the positive directions of the coordinate axes.

- If $l, m, n$ are the direction cosines of a line, then $l^{2}+m^{2}+n^{2}=1$.
- Direction cosines of a line joining two points $\mathrm{P}\left(x_{1}, y_{1}, z_{1}\right)$ and $\mathrm{Q}\left(x_{2}, y_{2}, z_{2}\right)$ are

$$
\frac{x_{2}-x_{1}}{\mathrm{PQ}}, \frac{y_{2}-y_{1}}{\mathrm{PQ}}, \frac{z_{2}-z_{1}}{\mathrm{PQ}}
$$

$$
\text { where PQ }=\sqrt{\left(x_{2}-x_{1}\right)^{2}+\left(y_{2}-y_{1}\right)^{2}+\left(z_{2}-z_{1}\right)^{2}}
$$

- Direction ratios of a line are the numbers which are proportional to the direction cosines of a line.
If $l, m, n$ are the direction cosines and $a, b, c$ are the direction ratios of a line


## Chapter-11 POINTS TO REMEMBER

$$
l=\frac{a}{\sqrt{a^{2}+b^{2}+c^{2}}} ; m=\frac{b}{\sqrt{a^{2}+b^{2}+c^{2}}} ; n=\frac{c}{\sqrt{a^{2}+b^{2}+c^{2}}}
$$

- Skew lines are lines in space which are neither parallel nor intersecting. They lie in different planes.
- Angle between skew lines is the angle between two intersecting lines drawn from any point (preferably through the origin) parallel to each of the skew lines.
- If $l_{1}, m_{1}, n_{1}$ and $l_{2}, m_{2}, m_{2}$ are the direction cosines of two lines; and $\theta$ is the acute angle between the two lines; then

$$
\cos \theta=\left|l_{1} l_{2}+m_{1} m_{2}+n_{1} n_{2}\right|
$$

- If $a_{1}, b_{1}, c_{1}$ and $a_{2}, b_{2}, c_{2}$ are the direction ratios of two lines and $\theta$ is the acute angle between the two lines; then

$$
\cos \theta=\left|\frac{a_{1} a_{2}+b_{1} b_{2}+c_{1} c_{2}}{\sqrt{a_{1}^{2}+b_{1}^{2}+c_{1}^{2}} \sqrt{a_{2}^{2}+b_{2}^{2}+c_{2}^{2}}}\right|
$$

- Vector equation of a line that passes through the given point whose position vector is $\vec{a}$ and parallel to a given vector $\vec{b}$ is $\vec{r}=\vec{a}+\lambda \vec{b}$.
- Equation of a line through a point $\left(x_{1}, y_{1}, z_{1}\right)$ and having direction cosines $h, m, n$ is $\frac{x-x_{1}}{\eta}=\frac{y-y_{1}}{\eta}=\frac{z-z_{1}}{\eta}$


## Chapter-11 POINTS TO REMEMBER

The vector equation of a line which passes through two points whose position vectors are $\vec{a}$ and $\vec{b}$ is $\vec{r}=\vec{a}+\lambda(\vec{b}-\vec{a})$.
Cartesian equation of a line that passes through two points ( $x_{1}, y_{1}, z_{1}$ ) and

$$
\left(x_{2}, y_{2}, z_{2}\right) \text { is } \frac{x-x_{1}}{x_{2}-x_{1}}=\frac{y-y_{1}}{y_{2}-y_{1}}=\frac{z-z_{1}}{z_{2}-z_{1}} .
$$

If $\theta$ is the acute angle between $\vec{r}=\vec{a}_{1}+\lambda \vec{b}_{1}$ and $\vec{r}=\vec{a}_{2}+\lambda \vec{b}_{2}$, then $\cos \theta=\left|\frac{\vec{b}_{1} \cdot \vec{b}_{2}}{\left|\vec{b}_{1}\right|\left|\vec{b}_{2}\right|}\right|$

- If $\frac{x-x_{1}}{l_{1}}=\frac{y-y_{1}}{m_{1}}=\frac{z-z_{1}}{n_{1}}$ and $\frac{x-x_{2}}{l_{2}}=\frac{y-y_{2}}{m_{2}}=\frac{z-z_{2}}{n_{2}}$
are the equations of two lines, then the acute angle between the two lines is given by $\cos \theta=\left|l_{1} l_{2}+m_{1} m_{2}+n_{1} n_{2}\right|$.


## Chapter-11 POINTS TO REMEMBER

Shortest distance between two skew lines is the line segment perpendicular to both the lines.

- Shortest distance between $\bar{r}=\bar{a}_{1}+\lambda \bar{b}_{1}$ and $\bar{\gamma}=\bar{a}_{2}+\mu \bar{b}_{2}$ is

$$
\left|\frac{\left(\bar{b}_{1} \times \bar{b}_{2}\right) \cdot\left(\bar{a}_{2}-\vec{a}_{1}\right)}{\left|\vec{b}_{1} \times \bar{b}_{2}\right|}\right|
$$

- Shortest distance between the lines: $\frac{x-x_{1}}{a_{1}}=\frac{y-y_{1}}{b_{1}}=\frac{z-z_{1}}{c_{1}}$ and

$$
\begin{aligned}
& \frac{x-x_{2}}{a_{2}}=\frac{y-y_{2}}{b_{2}}=\frac{z-z_{2}}{c_{2}} \text { is } \\
& \qquad \left.\begin{array}{cccc}
x_{2}-x_{1} & y_{2}-y_{1} & z_{2}-z_{1} \\
a_{1} & b_{1} & c_{1} \\
a_{2} & b_{2} & c_{2}
\end{array} \right\rvert\, \\
& \frac{\left.\mid c_{1} a_{2}-c_{2} a_{1}\right)^{2}+\left(a_{1} b_{2}-a_{2} b_{1}\right)^{2}}{\sqrt{\left(b_{1} c_{2}-b_{2} c_{1}\right)^{2}+\left(c_{1}\right.}}
\end{aligned}
$$

- Distance between parallel lines $\bar{r}=\bar{a}_{1}+\lambda \bar{b}$ and $\vec{r}=\bar{a}_{2}+\mu \bar{b}$ is

$$
\left|\frac{\bar{b} \times\left(\bar{a}_{2}-\bar{a}_{1}\right)}{\| \bar{b} \mid}\right|
$$

## Chapter-11 POINTS TO REMEMBER

In the vector form, equation of a plane which is at a distance $d$ from the origin, and $\hat{n}$ is the unit vector normal to the plane through the origin is $\vec{r} \cdot \hat{n}=d$.

- Equation of a plane which is at a distance of $d$ from the origin and the direction cosines of the normal to the plane as $l, m, n$ is $l x+m y+n z=d$.
- The equation of a plane through a point whose position vector is $\vec{a}$ and perpendicular to the vector $\overrightarrow{\mathrm{N}}$ is $(\vec{r}-\vec{a}) . \overrightarrow{\mathrm{N}}=0$.
- Equation of a plane perpendicular to a given line with direction ratios A, B, C and passing through a given point $\left(x_{1}, y_{1}, z_{1}\right)$ is

$$
\mathrm{A}\left(x-x_{1}\right)+\mathrm{B}\left(y-y_{1}\right)+\mathrm{C}\left(z-z_{1}\right)=0
$$

Equation of a plane passing through three non collinear points $\left(x_{1}, y_{1}, z_{1}\right)$,

## Chapter-11 POINTS TO REMEMBER

$\left(x_{2}, y_{2}, z_{2}\right)$ and $\left(x_{3}, y, z_{3}\right)$ is

$$
\left|\begin{array}{ccc}
x-x_{1} & y-y_{1} & z-z_{1} \\
x_{2}-x_{1} & y_{2}-y_{1} & z_{2}-z_{1} \\
x_{3}-x_{1} & y_{3}-y_{1} & z_{3}-z_{1}
\end{array}\right|=0
$$

- Vector equation of a plame that contains three non collinear points having position vectors $\bar{a}, \bar{b}$ and $\vec{c}$ is $(\vec{p}-\vec{a})-[(\bar{b}-\vec{a}) \times(\vec{c}-\bar{a})]=0$
- Equation of a plame that cuts the coondinates axes at $(a, O, O),(O, b, O)$ and $(\mathrm{CO}, \mathrm{O}, \mathrm{C})$ is

$$
\frac{x}{a}+\frac{y}{b}+\frac{z}{c}=1
$$

- Vector equation of a plane that passes through the intersection of plames $\bar{\gamma}-\bar{n}_{1}=d_{1}$ and $\overline{\gamma^{\prime}} \cdot \bar{n}_{2}=d_{2}$ is $\overline{p^{\prime}} \cdot\left(\bar{n}_{1}+\lambda \bar{n}_{2}\right)=d_{1}+\lambda d_{2}$, whene $\lambda$ is any nonzero constant.
- Vector equation of a plane that passes through the intersection of two given plames $A_{1} x+B_{1} y+C_{1} z+D_{1}=0$ and $A_{2} x+B_{2} y+C_{2} z+D_{2}=0$ is $\left(A_{1} x+B_{1} y+C_{1} z+D_{1}\right)+\lambda\left(A_{2} x+B_{2} y+C_{2} z+D_{2}\right)=0$.
- Two planes $\bar{\gamma}=\vec{a}_{1}+\lambda \bar{b}_{1}$ and $\vec{\gamma}=\vec{a}_{2}+\mu \bar{b}_{2}$ are coplamar if

$$
\left(\bar{a}_{2}-\bar{a}_{1}\right) \cdot\left(\bar{b}_{1} \times \bar{b}_{2}\right)=0
$$

## Chapter-11 POINTS TO REMEMBER

Two planes $a_{1} x+b_{1} y+c_{1} z+d_{1}=0$ and $a_{2} x+b_{2} y+c_{2} z+d_{2}=0$ are coplanar if $\left|\begin{array}{ccc}x_{2}-x_{1} & y_{2}-y_{1} & z_{2}-z_{1} \\ a_{1} & b_{1} & c_{1} \\ a_{2} & b_{2} & c_{2}\end{array}\right|=0$.

In the vector form, if $\theta$ is the angle between the two planes, $\vec{r} \cdot \vec{n}_{1}=d_{1}$ and $\vec{r} \cdot \vec{n}_{2}=d_{2}$, then $\theta=\cos ^{-1} \frac{\left|\vec{n}_{1} \cdot \vec{n}_{2}\right|}{\left|\vec{n}_{1}\right|\left|\vec{n}_{2}\right|}$.

- The angle $\phi$ between the line $\vec{r}=\vec{a}+\lambda \vec{b}$ and the plane $\vec{r} \cdot \hat{n}=d$ is

$$
\sin \phi=\left|\frac{\vec{b} \cdot \hat{n}}{|\vec{b}||\hat{n}|}\right|
$$

## Chapter-11 POINTS TO REMEMBER

The angle $\theta$ between the planes $\mathrm{A}_{1} x+\mathrm{B}_{1} y+\mathrm{C}_{1} z+\mathrm{D}_{1}=0$ and $\mathrm{A}_{2} x+\mathrm{B}_{2} y+\mathrm{C}_{2} z+\mathrm{D}_{2}=0$ is given by
$\cos \theta=\left|\frac{A_{1} A_{2}+B_{1} B_{2}+C_{1} C_{2}}{\sqrt{\mathrm{~A}_{1}^{2}+\mathrm{B}_{1}^{2}+\mathrm{C}_{1}^{2}} \sqrt{\mathrm{~A}_{2}^{2}+\mathrm{B}_{2}^{2}+\mathrm{C}_{2}^{2}}}\right|$

- The distance of a point whose position vector is $\vec{a}$ from the plane $\vec{r} \cdot \hat{n}=d$ is $|d-\vec{a} \cdot \hat{n}|$
- The distance from a point $\left(x_{1}, y_{1}, z_{1}\right)$ to the plane $\mathrm{A} x+\mathrm{B} y+\mathrm{C} z+\mathrm{D}=0$ is

$$
\left|\frac{\mathrm{A} x_{1}+\mathrm{B} y_{1}+\mathrm{C} z_{1}+\mathrm{D}}{\sqrt{\mathrm{~A}^{2}+\mathrm{B}^{2}+\mathrm{C}^{2}}}\right|
$$

## Chapter-12LINEAR PROGRAMMING

A linear programming problem is one that is concerned with finding the optimal value (maximum or minimum) of a linear function of several variables (called objective function) subject to the conditions that the variables are non-negative and satisfy a set of linear inequalities (called linear constraints). Variables are sometimes called decision variables and are non-negative.

- A few important linear programming problems are:
(i) Diet problems
(ii) Manufacturing problems
(iii) Transportation problems
- The common region determined by all the constraints including the non-negative constraints $x \geq 0, y \geq 0$ of a linear programming problem is called the feasible region (or solution region) for the problem.
- Points within and on the boundary of the feasible region represent feasible solutions of the constraints.

Any point outside the feasible region is an infeasible solution.

## Chapter-12 POINTS TO REMEMBER

Any point in the feasible region that gives the optimal value (maximum or minimum) of the objective function is called an optimal solution.

The following Theorems are fundamental in solving linear programming problems:
Theorem 1 Let $R$ be the feasible region (convex polygon) for a linear programming problem and let $Z=a x+b y$ be the objective function. When $Z$ has an optimal value (maximum or minimum), where the variables $x$ and $y$ are subject to constraints described by linear inequalities, this optimal value must occur at a corner point (vertex) of the feasible region.
Theorem 2 Let $R$ be the feasible region for a linear programming problem, and let $Z=a x+b y$ be the objective function. If $R$ is bounded, then the objective function $Z$ has both a maximum and minimum value on $R$ and each of these occurs at a corner point (vertex) of $R$.
If the feasible region is unbounded, then a maximum or a minimum may not exist. However, if it exists, it must occur at a corner point of $R$.

## Chapter-12 POINTS TO REMEMBER

Corner point method for solving a linear programming problem. The method comprises of the following steps:
(i) Find the feasible region of the linear programming problem and detemmine its corner points (vertices).
(ii) Evaluate the objective function $Z=a x+b y$ at each comer point. Let $M$ and $m$ respectively be the largest and smallest values at these points.
(iii) If the feasible region is bounded, $M$ and $m$ respectively are the maximum and minimum values of the objective function.
If the feasible region is umbounded, then
(i) $M$ is the maximum value of the objective function, if the open halfplane determined by $a x+b y>M$ has no point in common with the feasible region. Otherwise, the objective function has no maximum value.
(ii) $m$ is the minimum value of the objective function, if the open half plane detenmined by $a x+b y<m$ has no point in common with the feasible region. Otherwise, the objective function has no minimum value.
If two conner points of the feasible region are both optimal solutions of the same type, i.e., both produce the same maxinnum or minimum, then any point on the line segment joining these two points is also an optinal solution of the same type.

## Chapter-13PROBABILITY

The conditional probability of an event $E$, given the occurrence of the event $F$ is given by $P(E \mid F)=\frac{P(E \cap F)}{P(F)}, P(F) \neq 0$
$0 \leq \mathrm{P}(\mathrm{E} \mid \mathrm{F}) \leq 1, \quad \mathrm{P}\left(\mathrm{E}^{\prime} \mid \mathrm{F}\right)=1-\mathrm{P}(\mathrm{E} \mid \mathrm{F})$
$P((E \cup F) \mid G)=P(E \mid G)+P(F \mid G)-P((E \cap F) \mid G)$
$P(E \cap F)=P(E) P(F \mid E), P(E) \neq 0$
$P(E \cap F)=P(F) P(E \mid F), P(F) \neq 0$
If $E$ and $F$ are independent, then
$P(E \cap F)=P(E) P(F)$
$\mathrm{P}(\mathrm{E} \mid \mathrm{F})=\mathrm{P}(\mathrm{E}), \mathrm{P}(\mathrm{F}) \neq 0$
$P(F \mid E)=P(F), P(E) \neq 0$

## Chapter-13 POINTS TO REMEMBER

## Theorem of total probability

Let $\left\{E_{1}, E_{2}, \ldots, E_{n}\right)$ be a partition of a sample space and suppose that each of $E_{1}, E_{2}, \ldots, E_{n}$ has nonzero probability. Let A be any event associated with $S$, then
$P(A)=P\left(E_{1}\right) P\left(A \mid E_{1}\right)+P\left(E_{2}\right) P\left(A \mid E_{2}\right)+\ldots+P\left(E_{n}\right) P\left(A \mid E_{m}\right)$
Bayes" theorem If $E_{1}, E_{2}, \ldots, E_{p,}$ are events which constitute a partition of sample space $S$, i.e. $E_{1}, E_{2}, \ldots, E_{n}$ are pairwise disjoint and $E_{1} \cup E_{2} \cup \ldots \cup E_{n}=S$ and A be any event with nonzero probability, then

$$
P\left(E_{i} \mid A\right)=\frac{P\left(E_{i}\right) P\left(A \mid E_{i}\right)}{\sum_{j=1}^{m} P\left(E_{j}\right) P\left(A \mid E_{j}\right)}
$$

A randon variable is a real valued function whose domain is the sample space of a randon experiment.
The probability distribution of a random variable $X$ is the system of numbers

| $X$ | $:$ | $x_{1}$ | $x_{2}$ | $\cdots$ | $x_{n}$ |
| :--- | :--- | :--- | :--- | :--- | :--- |
| $P(X)$ | $:$ | $p_{1}$ | $p_{2}$ | $\cdots$ | $p_{n}$ |

where,

$$
p_{i}>0, \sum_{i=1}^{n} p_{i}=1, i=1,2, \ldots, n
$$

## Chapter-13 POINTS TO REMEMBER

Let $X$ be a random variable whose possible values $x_{1}, x_{2}, x_{3}, \ldots, x_{n}$ occur with probabilities $p_{1}, p_{2}, p_{3}, \ldots p_{n}$ respectively. The mean of $X$, denoted by $\mu$, is the number $\sum_{i=1}^{n} x_{i} p_{i}$.
The mean of a random variable $X$ is also called the expectation of $X$, denoted by $E$ (X).

- Let $X$ be a random variable whose possible values $x_{1}, x_{2}, \ldots, x_{n}$ occur with probabilities $p\left(x_{1}\right), p\left(x_{2}\right), \ldots, p\left(x_{n}\right)$ respectively.
Let $\mu=E(X)$ be the mean of $X$. The variance of $X$, denoted by Var ( $X$ ) or
$\sigma_{x}^{2}$, is defined as $\sigma_{x}^{2}=\operatorname{Var}(X)=\sum_{i=1}^{n}\left(x_{i}-\mu\right)^{2} p\left(x_{i}\right)$
or equivalently $\sigma_{x}^{2}=\mathrm{E}(X-\mu)^{2}$
The non-negative number

$$
\sigma_{x}=\sqrt{\operatorname{Var}(X)}=\sqrt{\sum_{i=1}^{n}\left(x_{i}-\mu\right)^{2} p\left(x_{i}\right)}
$$

is called the standard deviation of the randon variable $X$.

## Chapter-13 POINTS TO REMEMBER

$\operatorname{Var}(\mathrm{X})=\mathrm{E}\left(\mathrm{X}^{2}\right)-[\mathrm{E}(\mathrm{X})]^{2}$

- Trials of a random experiment are called Bernoulli trials, if they satisfy the following conditions:
(i) There should be a finite number of trials.
(ii) The trials should be independent.
(iii) Each trial has exactly two outcomes: success or failure.
(iv) The probability of success remains the same in each trial.

For Binomial distribution $\mathrm{B}(n, p), \mathrm{P}(\mathrm{X}=x)={ }^{n} \mathrm{C}_{x} q^{n-x} p^{x}, x=0,1, \ldots, n$ $(q=1-p)$

