Delhi CET Polytechnic 2018 Question Paper With Solutions

Time Allowed : 2 Hours	**Maximum Marks :** 600	**Total questions :** 150

General Instructions

Read the following instructions very carefully and strictly follow them:

1. **Mode of Examination:** Online (Computer Based exam)
2. **Number of Questions:** 150
3. **Type of Questions:** MCQ (Multiple Choice Questions)
4. **Duration:** 2 hours
5. **Negative Marking:** - 1 mark

Mathematics

1. Every point on a number line represents :

(1) An Irrational number

(2) A Rational number

(3) A Unique real number

(4) A Natural number

Correct Answer: (3) A Unique real number

Solution: Concept: The number line is a visual representation of numbers. The types of numbers that can be represented on it are important.

Step 1: Understanding the Number Line A number line is a straight line with points that are used to represent numbers. Typically, zero is placed at the center, positive numbers extend to the right, and negative numbers extend to the left.

Step 2: Types of Numbers

- **Natural Numbers:** $1, 2, 3, \dots$
- **Whole Numbers:** $0, 1, 2, 3, \dots$
- **Integers:** $\dots, -3, -2, -1, 0, 1, 2, 3, \dots$
- **Rational Numbers:** Numbers that can be expressed as a fraction p/q, where p and q are integers and $q \neq 0$. This includes integers, terminating decimals, and repeating decimals. Examples: $1/2, -3/4, 5, 0.25$.
- **Irrational Numbers:** Numbers that cannot be expressed as a simple fraction p/q. Their decimal representations are non-terminating and non-repeating. Examples: $\pi, \sqrt{2}, e$.
- **Real Numbers:** The set of all rational numbers and all irrational numbers combined. Real numbers fill the entire number line.

Step 3: What each point on the number line represents There is a one-to-one correspondence between the set of real numbers and the points on a number line. This means:

1. Every real number corresponds to exactly one unique point on the number line.
2. Every point on the number line corresponds to exactly one unique real number.

The number line includes points for integers, fractions (rational numbers), and numbers like $\sqrt{2}$ or π (irrational numbers).

Step 4: Analyzing the options

- **(1) An Irrational number:** Incorrect. Points like 2 or 0.5 are on the number line but are rational.
- **(2) A Rational number:** Incorrect. Points like $\sqrt{2}$ or π are on the number line but are irrational.
- **(3) A Unique real number:** Correct. Every point corresponds to one specific real number (which could be rational or irrational), and this real number is unique to that point.
- **(4) A Natural number:** Incorrect. Points like -1, 0.5, or $\sqrt{2}$ are on the number line but are not natural numbers.

Therefore, every point on a number line represents a unique real number.

Quick Tip

The number line is a complete representation of all **real numbers**. Real numbers include:

- Rational numbers (integers, fractions, terminating/repeating decimals)
- Irrational numbers (non-terminating, non-repeating decimals like π, $\sqrt{2}$)

Each point on the line maps to one specific real number, and vice versa.

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2. If a,b,c are positive real numbers, then $\sqrt{a^{-1}b} \times \sqrt{b^{-1}c} \times \sqrt{c^{-1}a}$ is equal to :

(1) abc

(2) $\sqrt{abc}$

(3) $\frac{1}{abc}$

(4) 1

Correct Answer: (4) 1

Solution: Concept: This problem involves simplifying an expression with square roots and negative exponents using the laws of exponents. Key laws of exponents:

- $x^{-n} = \frac{1}{x^n}$
- $\sqrt{x} = x^{1/2}$
- $\sqrt{xy} = \sqrt{x}\sqrt{y}$ or $\sqrt{x}\sqrt{y}\sqrt{z} = \sqrt{xyz}$
- $x^m \times x^n = x^{m+n}$
- $\frac{x^m}{x^n} = x^{m-n}$
- $x^0 = 1$ (for $x \neq 0$)

Step 1: Rewrite terms with negative exponents We are given the expression: $\sqrt{a^{-1}b} \times \sqrt{b^{-1}c} \times \sqrt{c^{-1}a}$ Using $x^{-1} = \frac{1}{x}$:

- $a^{-1}b = \frac{1}{a} \cdot b = \frac{b}{a}$
- $b^{-1}c = \frac{1}{b} \cdot c = \frac{c}{b}$
- $c^{-1}a = \frac{1}{c} \cdot a = \frac{a}{c}$

So the expression becomes:

$$\sqrt{\frac{b}{a}} \times \sqrt{\frac{c}{b}} \times \sqrt{\frac{a}{c}}$$

Step 2: Combine the square roots Using the property $\sqrt{x}\sqrt{y}\sqrt{z} = \sqrt{xyz}$, we can combine the terms under a single square root:

$$\sqrt{\frac{b}{a} \times \frac{c}{b} \times \frac{a}{c}}$$

Step 3: Simplify the expression inside the square root Multiply the fractions inside the square root:

$$\frac{b \times c \times a}{a \times b \times c}$$

We can cancel out common terms in the numerator and the denominator:

- The 'a' in the numerator cancels with the 'a' in the denominator.
- The 'b' in the numerator cancels with the 'b' in the denominator.
- The 'c' in the numerator cancels with the 'c' in the denominator.

So, the expression inside the square root simplifies to:

$$\frac{abc}{abc} = 1$$

Step 4: Calculate the final value The expression becomes:

$$\sqrt{1}$$

And the square root of 1 is 1.

$$\sqrt{1} = 1$$

Therefore, $\sqrt{a^{-1}b} \times \sqrt{b^{-1}c} \times \sqrt{c^{-1}a} = 1$.

Quick Tip

1. Convert negative exponents: $x^{-1} = 1/x$. So, $\sqrt{a^{-1}b} = \sqrt{b/a}$, etc. 2. Combine under one square root: $\sqrt{b/a \cdot c/b \cdot a/c}$ 3. Multiply the fractions inside: $\sqrt{(bca)/(abc)}$ 4. Cancel terms: Since multiplication is commutative ($bca = abc$), the fraction simplifies to 1. $\sqrt{1} = 1$.

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3. If $x + \frac{1}{x} = 2$, then $x^3 + \frac{1}{x^3} =$

(1) 64

(2) 8

(3) 2

(4) 1

Correct Answer: (3) 2

Solution: Concept: This problem can be solved in two ways: by finding the value of x first, or by using an algebraic identity.

Method 1: Finding the value of x Given: $x + \frac{1}{x} = 2$ Multiply the entire equation by x (assuming $x \neq 0$):

$$x(x) + x\left(\frac{1}{x}\right) = 2(x)$$

$$x^2 + 1 = 2x$$

Rearrange into a quadratic equation:

$$x^2 - 2x + 1 = 0$$

This is a perfect square trinomial: $(x-1)^2 = 0$. Taking the square root of both sides:

$$x - 1 = 0$$

$$x = 1$$

Now that we have $x = 1$, we can find the value of $x^3 + \frac{1}{x^3}$:

$$x^3 + \frac{1}{x^3} = (1)^3 + \frac{1}{(1)^3} = 1 + \frac{1}{1} = 1 + 1 = 2$$

Method 2: Using the algebraic identity for $(a+b)^3$ We know the identity: $(a+b)^3 = a^3 + b^3 + 3ab(a+b)$. Let $a = x$ and $b = \frac{1}{x}$. Then $a + b = x + \frac{1}{x}$. And $a^3 + b^3 = x^3 + \left(\frac{1}{x}\right)^3 = x^3 + \frac{1}{x^3}$. Also, $ab = x \cdot \frac{1}{x} = 1$.

Substitute these into the identity:

$$\left(x + \frac{1}{x}\right)^3 = \left(x^3 + \frac{1}{x^3}\right) + 3\left(x \cdot \frac{1}{x}\right)\left(x + \frac{1}{x}\right)$$

$$\left(x + \frac{1}{x}\right)^3 = \left(x^3 + \frac{1}{x^3}\right) + 3(1)\left(x + \frac{1}{x}\right)$$

We are given $x + \frac{1}{x} = 2$. Substitute this value:

$$(2)^3 = \left(x^3 + \frac{1}{x^3}\right) + 3(2)$$

$$8 = \left(x^3 + \frac{1}{x^3}\right) + 6$$

Now, solve for $x^3 + \frac{1}{x^3}$:

$$x^3 + \frac{1}{x^3} = 8 - 6$$

$$x^3 + \frac{1}{x^3} = 2$$

Both methods yield the same result.

Quick Tip

A useful trick: If $x + \frac{1}{x} = 2$, then x must be 1. You can check this: $1 + \frac{1}{1} = 1 + 1 = 2$. Once you know $x = 1$, then $x^n + \frac{1}{x^n}$ will always be $1^n + \frac{1}{1^n} = 1 + 1 = 2$ for any integer n. So, $x^3 + \frac{1}{x^3} = 2$.

4. If $(x-a)$ is a factor of $x^3 - 3x^2a + 2a^2x + b$, then the value of b is :

(1) 1

(2) 2

(3) 3

(4) 0

Correct Answer: (4) 0

Solution: Concept: This problem uses the Factor Theorem. The Factor Theorem states that if $(x-k)$ is a factor of a polynomial $P(x)$, then $P(k) = 0$. Conversely, if $P(k) = 0$, then $(x-k)$ is a factor of $P(x)$.

Step 1: Apply the Factor Theorem We are given that $(x-a)$ is a factor of the polynomial $P(x) = x^3 - 3x^2a + 2a^2x + b$. According to the Factor Theorem, if $(x-a)$ is a factor, then $P(a) = 0$.

Step 2: Substitute $x = a$ into the polynomial $P(x)$ Replace every x in $P(x)$ with a:

$$P(a) = (a)^3 - 3(a)^2a + 2a^2(a) + b$$

Step 3: Simplify the expression for $P(a)$

$$P(a) = a^3 - 3(a^2 \cdot a) + 2(a^2 \cdot a) + b$$

$$P(a) = a^3 - 3a^3 + 2a^3 + b$$

Combine the terms with a^3:

$$P(a) = (1 - 3 + 2)a^3 + b$$

$$P(a) = (-2 + 2)a^3 + b$$

$$P(a) = (0)a^3 + b$$

$$P(a) = 0 + b$$

$$P(a) = b$$

Step 4: Set $P(a) = 0$ and solve for b Since $(x-a)$ is a factor, we know $P(a) = 0$. From Step 3, we found $P(a) = b$. Therefore, we must have:

$$b = 0$$

The value of b is 0.

Quick Tip

Factor Theorem: If $(x-k)$ is a factor of $P(x)$, then $P(k)=0$. Here, $(x-a)$ is a factor, so $k=a$. Let $P(x)=x^3-3x^2a+2a^2x+b$. Set $P(a)=0$: $a^3-3a^2(a)+2a^2(a)+b=0$ $a^3-3a^3+2a^3+b=0$ $(1-3+2)a^3+b=0$ $0\cdot a^3+b=0$ $0+b=0$ $b=0$

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5. If $(2K-1,K)$ is a Solution of the equation $10x-9y=12$, then K =

(1) 2

(2) 1

(3) 4

(4) 3

Correct Answer: (1) 2

Solution: Concept: If a point (x_0,y_0) is a solution to an equation, it means that when you substitute $x=x_0$ and $y=y_0$ into the equation, the equation will be true.

Step 1: Identify the values of x and y from the given solution The given solution is the ordered pair $(2K-1,K)$. This means: $x=2K-1$ $y=K$

Step 2: Substitute these expressions for x and y into the given equation The given equation is $10x-9y=12$. Substitute $x=2K-1$ and $y=K$ into this equation:

$$10(2K-1)-9(K)=12$$

Step 3: Solve the resulting equation for K First, expand the term $10(2K-1)$:

$$(10\times 2K)-(10\times 1)-9K=12$$

$$20K-10-9K=12$$

Combine the terms with K:

$$(20K-9K)-10=12$$

$$11K-10=12$$

Now, isolate the term with K. Add 10 to both sides of the equation:

$$11K - 10 + 10 = 12 + 10$$

$$11K = 22$$

Finally, divide by 11 to solve for K:

$$K = \frac{22}{11}$$

$$K = 2$$

Step 4: Check the answer (optional) If $K = 2$, then $x = 2K - 1 = 2(2) - 1 = 4 - 1 = 3$, and $y = K = 2$. Substitute $x = 3, y = 2$ into the original equation $10x - 9y = 12$:
$10(3) - 9(2) = 30 - 18 = 12$. Since $12 = 12$, the solution is correct.
The value of K is 2.

> **Quick Tip**
>
> If (x_0, y_0) is a solution to an equation, substitute x_0 for x and y_0 for y. Given solution: $x = 2K - 1$, $y = K$. Equation: $10x - 9y = 12$. Substitute: $10(2K - 1) - 9(K) = 12$. Simplify: $20K - 10 - 9K = 12$. $11K - 10 = 12$. $11K = 22$. $K = 2$.

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6. The perpendicular distance of the point P (6,8) from x axis is :

(1) 8

(2) 6

(3) 10

(4) None of these

Correct Answer: (1) 8

Solution: Concept: In a 2D Cartesian coordinate system, a point P is represented by an ordered pair (x, y).

- The x-coordinate represents the perpendicular distance of the point from the y-axis.
- The y-coordinate represents the perpendicular distance of the point from the x-axis.

Distance is always a non-negative value.

Step 1: Identify the coordinates of the point P The given point is P (6,8). Here, $x = 6$ and $y = 8$.

Step 2: Understand "perpendicular distance from the x-axis" The x-axis is the horizontal line where $y = 0$. The perpendicular distance of a point (x, y) from the x-axis is the length of the vertical line segment from the point down to (or up to) the x-axis. This length is simply the absolute value of the y-coordinate.

Imagine plotting the point P(6,8):

- You move 6 units to the right along the x-axis.
- Then, you move 8 units up, parallel to the y-axis.

The distance you moved upwards (8 units) is the perpendicular distance from the x-axis.

Step 3: Determine the perpendicular distance For the point P(6,8), the y-coordinate is 8. The perpendicular distance from the x-axis is $|y| = |8| = 8$ units.

Step 4: Compare with the options The calculated distance is 8. This matches option (1).

Quick Tip

For any point P(x, y):

- Perpendicular distance from the **x-axis** is $|y|$ (the absolute value of the y-coordinate).
- Perpendicular distance from the **y-axis** is $|x|$ (the absolute value of the x-coordinate).

Given point P(6,8): Distance from x-axis = y-coordinate = 8. Distance from y-axis = x-coordinate = 6.

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7. How many least number of distinct points determine a unique line ?

(1) 3

(2) 2

(3) 1

(4) 4

Correct Answer: (2) 2

Solution: Concept: This question relates to a fundamental postulate of Euclidean geometry concerning points and lines.

Step 1: Considering one point If you have only one point, an infinite number of different lines can pass through that single point. Imagine a point as a pivot; you can rotate a line around it in all directions. Thus, one point does not determine a unique line.

Step 2: Considering two distinct points If you have two distinct (different) points, there is exactly one straight line that can pass through both of them. You can draw this line by placing a ruler along the two points. Any other line you try to draw that passes through one of these points will not pass through the other if it's a different line. This is often stated as a postulate: "Through any two distinct points, there is exactly one line."

Step 3: Considering three or more points

- If three points are collinear (all lie on the same line), they still determine that same single unique line (defined by any two of them).
- If three points are non-collinear (forming a triangle), no single straight line can pass through all three. Each pair of points will define a different line.

The question asks for the "least number" of distinct points to determine a *unique* line.

Step 4: Conclusion The least number of distinct points required to determine a unique straight line is **two**.

Quick Tip

Think about drawing:

- With **1 point**, you can draw endless lines through it.
- With **2 distinct points**, you can draw only **one** straight line that goes through both.

This is a basic rule in geometry: Two points define a unique line.

8. If two interior angles, on the same side of a transversal intersecting two parallel lines, are in the ratio 2 : 3, then the measure of the larger angle is :

(1) 54°

(2) 120°

(3) 108°

(4) 136°

Correct Answer: (3) 108°

Solution: Concept: When a transversal intersects two parallel lines, specific relationships exist between the angles formed. Interior angles on the same side of the transversal are supplementary (their sum is 180°). These are also known as consecutive interior angles or same-side interior angles.

Step 1: Understand the property of consecutive interior angles If two parallel lines are intersected by a transversal, then the sum of the interior angles on the same side of the transversal is 180°. Let the two consecutive interior angles be A and B. If the lines are parallel, then $A + B = 180^\circ$.

Step 2: Set up the angles based on the given ratio The two interior angles are in the ratio 2 : 3. Let the common factor for the ratio be x. Then the measures of the two angles are $2x$ and $3x$.

Step 3: Use the supplementary property to form an equation Since these are consecutive interior angles and the lines are parallel, their sum must be 180°:

$$2x + 3x = 180^\circ$$

Step 4: Solve for x

$$5x = 180^\circ$$
$$x = \frac{180^\circ}{5}$$
$$x = 36^\circ$$

Step 5: Calculate the measures of the two angles The first angle is $2x = 2 \times 36^\circ = 72^\circ$. The second angle is $3x = 3 \times 36^\circ = 108^\circ$.

Step 6: Identify the larger angle Comparing the two angles, 72° and 108°, the larger angle is 108°. (Check: $72^\circ + 108^\circ = 180^\circ$, so they are supplementary).

The measure of the larger angle is 108°.

Quick Tip

Property: Interior angles on the same side of a transversal between parallel lines add up to 180°. Ratio of angles = 2 : 3. Let angles be $2x$ and $3x$. Equation: $2x + 3x = 180^\circ$. $5x = 180^\circ \implies x = 36^\circ$. Angles are: $2x = 2 \times 36^\circ = 72^\circ$ $3x = 3 \times 36^\circ = 108^\circ$ Larger angle = 108°.

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9. If the bisectors of the acute angles of a right triangle meet at O, then the angle at O, between the two bisectors is :

(1) 45°

(2) 95°

(3) 135°

(4) 90°

Correct Answer: (3) 135°

Solution: Concept: This problem involves the properties of angles in a triangle, angle bisectors, and the sum of angles in a triangle.

Step 1: Properties of a right triangle Let the right triangle be $\triangle ABC$, with the right angle at B ($\angle B = 90^\circ$). The other two angles, $\angle A$ and $\angle C$, are acute angles. The sum of angles in a triangle is 180°, so $\angle A + \angle B + \angle C = 180^\circ$. Since $\angle B = 90^\circ$, we have $\angle A + 90^\circ + \angle C = 180^\circ$, which means $\angle A + \angle C = 90^\circ$.

Step 2: Angle bisectors Let AO be the bisector of $\angle A$, and CO be the bisector of $\angle C$. These bisectors meet at point O. This means:

- $\angle OAB = \angle OAC = \frac{1}{2}\angle A$
- $\angle OCB = \angle OCA = \frac{1}{2}\angle C$

We are interested in the angle $\angle AOC$, which is the angle at O between the two bisectors.

Step 3: Consider the triangle $\triangle AOC$ The sum of angles in $\triangle AOC$ is 180°:

$$\angle OAC + \angle OCA + \angle AOC = 180^\circ$$

Substitute the expressions from the angle bisectors:

$$\left(\frac{1}{2}\angle A\right) + \left(\frac{1}{2}\angle C\right) + \angle AOC = 180^\circ$$

$$\frac{1}{2}(\angle A + \angle C) + \angle AOC = 180^\circ$$

Step 4: Use the property $\angle A + \angle C = 90^\circ$ Substitute $\angle A + \angle C = 90^\circ$ into the equation from Step 3:

$$\frac{1}{2}(90^\circ) + \angle AOC = 180^\circ$$

$$45^\circ + \angle AOC = 180^\circ$$

Step 5: Solve for $\angle AOC$

$$\angle AOC = 180^\circ - 45^\circ$$

$$\angle AOC = 135^\circ$$

Thus, the angle at O, between the two bisectors, is 135°.

General Formula Note: For any triangle $\triangle ABC$, if the bisectors of $\angle B$ and $\angle C$ meet at O, then $\angle BOC = 90^\circ + \frac{1}{2}\angle A$. In our case, we are looking at the angle formed by bisectors of acute angles A and C, with the third angle B being 90°. So the angle at O (i.e., $\angle AOC$) can be thought of as $90^\circ + \frac{1}{2}\angle B$. Since $\angle B = 90^\circ$, then $\angle AOC = 90^\circ + \frac{1}{2}(90^\circ) = 90^\circ + 45^\circ = 135^\circ$.

Quick Tip

In any triangle, let the angles be A, B, C. If bisectors of $\angle A$ and $\angle C$ meet at O, then $\angle AOC = 180^\circ - (\frac{\angle A}{2} + \frac{\angle C}{2})$. In a right triangle, one angle is 90°. Let this be $\angle B$. Then the sum of the two acute angles $\angle A + \angle C = 90^\circ$. So, $\frac{\angle A}{2} + \frac{\angle C}{2} = \frac{1}{2}(\angle A + \angle C) = \frac{1}{2}(90^\circ) = 45^\circ$. Therefore, $\angle AOC = 180^\circ - 45^\circ = 135^\circ$. A useful shortcut: the angle between bisectors of two angles of a triangle is $90^\circ + \frac{1}{2}(\text{third angle})$. Here, the third angle (the right angle) is 90°. So $90^\circ + \frac{90^\circ}{2} = 90^\circ + 45^\circ = 135^\circ$.

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10. If $\triangle ABC \cong \triangle ACB$, then $\triangle ABC$ is Isosceles with :

(1) AB=AC

(2) AB=BC

(3) AC=BC

(4) None of these

Correct Answer: (1) AB=AC

Solution: **Concept:** Congruence of triangles ($\cong$) means that two triangles have exactly the same size and shape. The order of vertices in the congruence statement is crucial as it indicates corresponding parts.

Step 1: Understanding the congruence statement We are given $\triangle ABC \cong \triangle ACB$. This statement implies the following correspondence between vertices:

- Vertex A in $\triangle ABC$ corresponds to Vertex A in $\triangle ACB$.
- Vertex B in $\triangle ABC$ corresponds to Vertex C in $\triangle ACB$.
- Vertex C in $\triangle ABC$ corresponds to Vertex B in $\triangle ACB$.

Step 2: Identifying corresponding sides Since corresponding parts of congruent triangles are equal (CPCTC), we can equate the lengths of corresponding sides:

- Side AB (from $\triangle ABC$) corresponds to Side AC (from $\triangle ACB$). Therefore, $AB = AC$.
- Side BC (from $\triangle ABC$) corresponds to Side CB (from $\triangle ACB$). Therefore, $BC = CB$ (This is trivial, as it's the same side).
- Side AC (from $\triangle ABC$) corresponds to Side AB (from $\triangle ACB$). Therefore, $AC = AB$ (This is the same as the first conclusion).

Step 3: Conclusion about the triangle From the correspondence, we found that $AB = AC$. A triangle in which at least two sides are equal in length is called an isosceles triangle. Since $AB = AC$, $\triangle ABC$ is an isosceles triangle with these two sides being equal.

Step 4: Analyzing the options

- **(1) AB=AC:** This matches our finding from the congruence.
- **(2) AB=BC:** This is not necessarily true from the given congruence.
- **(3) AC=BC:** This is not necessarily true from the given congruence.
- **(4) None of these:** Incorrect, as option (1) is correct.

Therefore, if $\triangle ABC \cong \triangle ACB$, then $\triangle ABC$ is isosceles with $AB = AC$.

Quick Tip

The order of vertices in a triangle congruence statement is very important. $\triangle\mathbf{ABC} \cong \triangle\mathbf{ACB}$ This means:

- 1st vertex corresponds to 1st: A $\leftrightarrow$ A
- 2nd vertex corresponds to 2nd: B $\leftrightarrow$ C
- 3rd vertex corresponds to 3rd: C $\leftrightarrow$ B

Corresponding sides are equal:
AB (1st & 2nd) = AC (1st & 2nd)
BC (2nd & 3rd) = CB (2nd & 3rd)
AC (1st & 3rd) = AB (1st & 3rd)
From this, $AB = AC$, so the triangle is isosceles with these two sides equal.

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11. The figure formed by joining the mid points of the adjacent sides of a rectangle is :
(1) Square
(2) Rhombus
(3) Trapezium
(4) None of these

Correct Answer: (2) Rhombus

Solution: Concept: This question refers to a well-known geometrical theorem called Varignon's Theorem, or a specific case of it. Varignon's Theorem states that the figure formed by joining the midpoints of the sides of any quadrilateral is a parallelogram. We need to determine the specific type of parallelogram formed when the original quadrilateral is a rectangle.
Step 1: Properties of a Rectangle A rectangle is a parallelogram with four right angles. Its diagonals are equal in length and bisect each other.

Step 2: Consider a rectangle ABCD Let P, Q, R, and S be the midpoints of sides AB, BC, CD, and DA respectively. We need to determine the type of quadrilateral PQRS.

Step 3: Using the Midpoint Theorem In $\triangle ABD$, S is the midpoint of AD and P is the midpoint of AB. By the Midpoint Theorem, SP is parallel to DB and $SP = \frac{1}{2}DB$. Similarly, in $\triangle CBD$, R is the midpoint of CD and Q is the midpoint of BC. By the Midpoint Theorem, RQ is parallel to DB and $RQ = \frac{1}{2}DB$. Thus, SP is parallel to RQ and $SP = RQ$.

Also, in $\triangle ABC$, P is the midpoint of AB and Q is the midpoint of BC. By the Midpoint Theorem, PQ is parallel to AC and $PQ = \frac{1}{2}AC$. Similarly, in $\triangle ADC$, S is the midpoint of AD and R is the midpoint of CD. By the Midpoint Theorem, SR is parallel to AC and $SR = \frac{1}{2}AC$. Thus, PQ is parallel to SR and $PQ = SR$.

Since opposite sides are equal and parallel (SP=RQ and PQ=SR), PQRS is a parallelogram.

Step 4: Properties of the sides of parallelogram PQRS We have $SP = RQ = \frac{1}{2}DB$ and $PQ = SR = \frac{1}{2}AC$. In a rectangle, the diagonals are equal, i.e., $AC = DB$. Therefore, $SP = PQ = QR = RS = \frac{1}{2} \times$ (length of diagonal of rectangle). Since all four sides of the parallelogram PQRS are equal, PQRS is a **rhombus**.

Step 5: When would it be a square? A rhombus becomes a square if its angles are $90°$. The angles of the rhombus PQRS would be $90°$ if the diagonals of the original rectangle (AC and DB) are perpendicular. However, the diagonals of a general rectangle are not necessarily perpendicular (they are only perpendicular if the rectangle is a square). So, for a general rectangle, the figure formed is a rhombus. If the original rectangle is a square, then the figure formed by joining midpoints is also a square (which is a special type of rhombus). But the general case for a rectangle is a rhombus.

Quick Tip

Key theorem: The quadrilateral formed by joining the midpoints of the sides of any quadrilateral is a parallelogram (Varignon's Theorem).

- For a general quadrilateral $\rightarrow$ Parallelogram
- For a rectangle $\rightarrow$ **Rhombus** (because diagonals of a rectangle are equal, making all sides of the midpoint figure equal).
- For a rhombus $\rightarrow$ Rectangle (because diagonals of a rhombus are perpendicular).
- For a square $\rightarrow$ Square.

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12. Diagonals AC and BD of trapezium ABCD, in which AB$||$DC intersect each other at O. The triangle which is equal in area of $\triangle$AOD is :

(1) $\triangle$AOB

(2) $\triangle$BOC

(3) $\triangle$DOC

(4) $\triangle$ADC

Correct Answer: (2) $\triangle$BOC

Solution: Concept: This problem uses the property of triangles on the same base and between the same parallel lines, and how areas of triangles are related when diagonals of a trapezium intersect.

Step 1: Property of triangles on the same base and between the same parallels Triangles on the same base (or equal bases) and between the same parallel lines are equal in area. In trapezium ABCD, AB $||$ DC. Consider triangles $\triangle ADC$ and $\triangle BDC$. They share the same base DC and are between the same parallel lines AB and DC. Therefore, Area($\triangle ADC$) = Area($\triangle BDC$).

Step 2: Decompose the areas We can write the areas as: Area($\triangle ADC$) = Area($\triangle AOD$) + Area($\triangle DOC$) Area($\triangle BDC$) = Area($\triangle BOC$) + Area($\triangle DOC$)

Step 3: Equate the areas and simplify Since Area($\triangle ADC$) = Area($\triangle BDC$), we have: Area($\triangle AOD$) + Area($\triangle DOC$) = Area($\triangle BOC$) + Area($\triangle DOC$) Subtract Area($\triangle DOC$)

from both sides of the equation: Area($\triangle AOD$) = Area($\triangle BOC$)

Step 4: Conclusion The triangle which is equal in area to $\triangle AOD$ is $\triangle BOC$.

Visualization: Imagine the trapezium. The triangles $\triangle ADB$ and $\triangle ACB$ are also on the same base AB and between parallels AB and DC, so Area($\triangle ADB$) = Area($\triangle ACB$). Area($\triangle ADB$) = Area($\triangle AOD$) + Area($\triangle AOB$) Area($\triangle ACB$) = Area($\triangle BOC$) + Area($\triangle AOB$) Equating these: Area($\triangle AOD$) + Area($\triangle AOB$) = Area($\triangle BOC$) + Area($\triangle AOB$). Subtracting Area($\triangle AOB$) from both sides gives Area($\triangle AOD$) = Area($\triangle BOC$).

Quick Tip

A key property for trapeziums: When the diagonals intersect, they form four triangles. The two triangles formed by a non-parallel side and segments of the diagonals are equal in area. In trapezium ABCD with AB || DC, and diagonals intersecting at O: Area($\triangle AOD$) = Area($\triangle BOC$). This is because $\triangle ADC$ and $\triangle BDC$ have the same base DC and are between the same parallels, so their areas are equal. Removing the common $\triangle DOC$ from both leaves $\triangle AOD$ and $\triangle BOC$ with equal areas.

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13. The chord of a circle is equal to its radius. The angle subtended by this chord at the minor arc of the circle, is :

(1) 60°

(2) 75°

(3) 120°

(4) 150°

Correct Answer: (4) 150°

Solution: Concept: This problem involves the relationship between the angle subtended by a chord at the center of a circle and the angles subtended by the same chord at points on the major and minor arcs.

Step 1: Angle subtended by the chord at the center Let the circle have center O and radius r. Let AB be a chord such that its length is equal to the radius, i.e., $AB = r$. Consider

$\triangle OAB$. We have $OA = r$ (radius), $OB = r$ (radius), and $AB = r$ (given). Since all three sides are equal ($OA = OB = AB = r$), $\triangle OAB$ is an equilateral triangle. The angle subtended by the chord AB at the center O is $\angle AOB$. In an equilateral triangle, all angles are 60°. So, $\angle AOB = 60^\circ$.

Step 2: Angle subtended at the major arc The angle subtended by an arc (or chord) at the center is double the angle subtended by it at any point on the remaining part of the circle (the major arc in this case). Let C be any point on the major arc. Then, $\angle ACB = \frac{1}{2}\angle AOB$.

$$\angle ACB = \frac{1}{2} \times 60^\circ = 30^\circ$$

Step 3: Angle subtended at the minor arc Let D be any point on the minor arc. The points A, C, B, D in order form a cyclic quadrilateral ACBD. In a cyclic quadrilateral, the sum of opposite angles is 180°. So, $\angle ACB + \angle ADB = 180^\circ$. We found $\angle ACB = 30^\circ$. Therefore, $\angle ADB = 180^\circ - \angle ACB = 180^\circ - 30^\circ = 150^\circ$. The angle $\angle ADB$ is the angle subtended by the chord AB at a point on the minor arc.

Thus, the angle subtended by this chord at the minor arc of the circle is 150°.

Quick Tip

1. Chord = Radius $\implies$ Triangle formed by chord and two radii to its ends is equilateral. Angle at center ($\angle AOB$) = 60°. 2. Angle at major arc ($\angle ACB$) = $\frac{1}{2}$ $\times$ angle at center $= \frac{1}{2} \times 60^\circ = 30^\circ$. 3. Points on major arc and minor arc with chord ends form a cyclic quadrilateral. Angle at minor arc ($\angle ADB$) + Angle at major arc ($\angle ACB$) = 180°. So, Angle at minor arc = $180^\circ - 30^\circ = 150^\circ$.

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14. A square and an equilateral triangle have equal perimeters. If the diagonal of the square is $12\sqrt{2}$ cm, then area of the equilateral triangle is :

(1) $24\sqrt{2}$ cm^2

(2) $24\sqrt{3}$ cm^2

(3) $48\sqrt{3}$ cm^2

(4) $64\sqrt{3}$ cm^2

Correct Answer: (4) $64\sqrt{3}$ cm^2

Solution: Concept: This problem involves relating the dimensions and perimeters of a square and an equilateral triangle, and then calculating the area of the equilateral triangle. Formulas needed:

- For a square with side s: Diagonal $d = s\sqrt{2}$, Perimeter $P_s = 4s$.
- For an equilateral triangle with side a: Perimeter $P_t = 3a$, Area $A_t = \frac{\sqrt{3}}{4}a^2$.

Step 1: Find the side of the square Let the side of the square be s. The diagonal of the square is given as $d = 12\sqrt{2}$ cm. We know $d = s\sqrt{2}$. So, $s\sqrt{2} = 12\sqrt{2}$. Dividing by $\sqrt{2}$, we get $s = 12$ cm.

Step 2: Calculate the perimeter of the square Perimeter of the square $P_s = 4s = 4 \times 12 = 48$ cm.

Step 3: Find the side of the equilateral triangle Let the side of the equilateral triangle be a. The problem states that the square and the equilateral triangle have equal perimeters. So, Perimeter of equilateral triangle $P_t = P_s = 48$ cm. We also know $P_t = 3a$. Therefore, $3a = 48$. $a = \frac{48}{3} = 16$ cm.

Step 4: Calculate the area of the equilateral triangle The area of an equilateral triangle with side a is $A_t = \frac{\sqrt{3}}{4}a^2$. Substitute $a = 16$ cm:

$$A_t = \frac{\sqrt{3}}{4}(16)^2$$

$$A_t = \frac{\sqrt{3}}{4}(256)$$

$$A_t = \sqrt{3} \times \frac{256}{4}$$

$$A_t = \sqrt{3} \times 64$$

$$A_t = 64\sqrt{3} \text{ cm}^2$$

The area of the equilateral triangle is $64\sqrt{3}$ cm^2.

Quick Tip

1. **Square's side:** Diagonal $d = s\sqrt{2}$. Given $d = 12\sqrt{2}$, so side $s = 12$ cm. 2. **Square's perimeter:** $P_s = 4s = 4 \times 12 = 48$ cm. 3. **Triangle's side:** Perimeters are equal, so $P_t = 48$ cm. For equilateral triangle, $P_t = 3a$. So $3a = 48 \implies a = 16$ cm. 4. **Triangle's area:** Area $A_t = \frac{\sqrt{3}}{4}a^2 = \frac{\sqrt{3}}{4}(16)^2 = \frac{\sqrt{3}}{4}(256) = 64\sqrt{3}$ cm^2.

15. The length, width and height of a rectangular solid are in the ratio of 3:2:1. If the volume of the solid is 48 cm^3. The total surface area at the solid is :

(1) 27 cm^2

(2) 32 cm^2

(3) 44 cm^2

(4) 88 cm^2

Correct Answer: (4) 88 cm^2

Solution: Concept: This problem involves a rectangular solid (cuboid) where the dimensions are in a given ratio and the volume is known. We need to find the total surface area. Formulas for a cuboid with length l, width w, and height h:

- Volume (V) = $l \times w \times h$
- Total Surface Area (TSA) = $2(lw + wh + hl)$

Step 1: Express dimensions in terms of a common factor The ratio of length : width : height is 3 : 2 : 1. Let the common factor be x. Then, length $l = 3x$, width $w = 2x$, and height $h = 1x = x$.

Step 2: Use the given volume to find x The volume of the solid is given as $V = 48$ cm^3. Using the formula $V = lwh$:

$$(3x)(2x)(x) = 48$$

$$6x^3 = 48$$

Divide by 6:

$$x^3 = \frac{48}{6}$$

$$x^3 = 8$$

Take the cube root of both sides:

$$x = \sqrt[3]{8} = 2 \text{ cm}$$

Step 3: Calculate the actual dimensions Now that we have $x = 2$:

- Length $l = 3x = 3 \times 2 = 6$ cm

- Width $w = 2x = 2 \times 2 = 4$ cm
- Height $h = x = 2$ cm

Step 4: Calculate the Total Surface Area (TSA) Using the formula $TSA = 2(lw + wh + hl)$:

$$TSA = 2((6)(4) + (4)(2) + (2)(6))$$

$$TSA = 2(24 + 8 + 12)$$

$$TSA = 2(44)$$

$$TSA = 88 \text{ cm}^2$$

The total surface area of the solid is 88 cm^2.

Quick Tip

1. **Dimensions from ratio:** $l = 3x, w = 2x, h = x$. 2. **Use volume to find x:** $V = lwh \implies (3x)(2x)(x) = 6x^3$. Given $V = 48$, so $6x^3 = 48 \implies x^3 = 8 \implies x = 2$. 3. **Actual dimensions:** $l = 6, w = 4, h = 2$ cm. 4. **Total Surface Area:** $TSA = 2(lw + wh + hl)$ $TSA = 2((6)(4) + (4)(2) + (2)(6)) = 2(24 + 8 + 12) = 2(44) = 88 \text{ cm}^2$.

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16. If the diameter of the base of a closed right circular cylinder be equal to its height, h, then its whole surface area is :

(1) $2\pi h^2$

(2) $\frac{3}{2}\pi h^2$

(3) $\frac{4}{3}\pi h^2$

(4) πh^2

Correct Answer: (2) $\frac{3}{2}\pi h^2$

Solution: Concept: The whole surface area (total surface area) of a closed right circular cylinder includes the area of its two circular bases and the area of its curved surface. Formula for Total Surface Area (TSA) of a cylinder: $TSA = 2\pi r^2 + 2\pi rh$, where r is the radius of the base and h is the height.

Step 1: Relate diameter, radius, and height Let d be the diameter of the base and r be the radius. We know that diameter $d = 2r$. The problem states that the diameter of the base is equal to its height, h. So, $d = h$. Since $d = 2r$, we have $2r = h$. This implies that the radius $r = \frac{h}{2}$.

Step 2: Substitute $r = \frac{h}{2}$ into the TSA formula The TSA formula is $TSA = 2\pi r^2 + 2\pi rh$. Substitute $r = \frac{h}{2}$:

$$TSA = 2\pi \left(\frac{h}{2}\right)^2 + 2\pi \left(\frac{h}{2}\right) h$$

Step 3: Simplify the expression

$$TSA = 2\pi \left(\frac{h^2}{4}\right) + 2\pi \left(\frac{h^2}{2}\right)$$

Simplify the fractions:

$$TSA = \frac{2\pi h^2}{4} + \frac{2\pi h^2}{2}$$
$$TSA = \frac{\pi h^2}{2} + \pi h^2$$

To add these terms, find a common denominator (which is 2): $\pi h^2 = \frac{2\pi h^2}{2}$ So,

$$TSA = \frac{\pi h^2}{2} + \frac{2\pi h^2}{2}$$
$$TSA = \frac{\pi h^2 + 2\pi h^2}{2}$$
$$TSA = \frac{3\pi h^2}{2}$$

This can also be written as $\frac{3}{2}\pi h^2$.

Quick Tip

1. Given: diameter $d = h$. 2. Radius $r = d/2 = h/2$. 3. Total Surface Area of cylinder (TSA) = $2\pi r$(bases) $+ 2\pi rh$(curved surface). 4. Substitute $r = h/2$: $TSA = 2\pi(h/2)^2 + 2\pi(h/2)h$ $TSA = 2\pi(h^2/4) + \pi h^2$ $TSA = \pi h^2/2 + \pi h^2$ $TSA = \pi h^2/2 + 2\pi h^2/2 = 3\pi h^2/2$.

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17. If the radius of the base of a right circular cone is $3r$ and its height is equal to the radius of the base, then its volume is :

(1) $\frac{1}{3}\pi r^3$

(2) $\frac{2}{3}\pi r^3$

(3) $3\pi r^3$

(4) $9\pi r^3$

Correct Answer: (4) $9\pi r^3$

Solution: Concept: The volume of a right circular cone is given by the formula $V = \frac{1}{3}\pi R^2 H$, where R is the radius of the base of the cone and H is the height of the cone.

Step 1: Identify the given dimensions of the cone Let the radius of the base of this specific cone be R_{cone} and its height be H_{cone}.

- Radius of the base, $R_{\text{cone}} = 3r$.
- Height of the cone, H_{cone}, is "equal to the radius of the base." This means $H_{\text{cone}} = R_{\text{cone}} = 3r$.

Step 2: Substitute these dimensions into the volume formula The volume formula for a cone is $V = \frac{1}{3}\pi R_{\text{cone}}^2 H_{\text{cone}}$. Substitute $R_{\text{cone}} = 3r$ and $H_{\text{cone}} = 3r$:

$$V = \frac{1}{3}\pi(3r)^2(3r)$$

Step 3: Simplify the expression First, calculate $(3r)^2$:

$$(3r)^2 = 3^2 \times r^2 = 9r^2$$

Now substitute this back into the volume equation:

$$V = \frac{1}{3}\pi(9r^2)(3r)$$

Multiply the terms:

$$V = \frac{1}{3}\pi(9 \times 3 \times r^2 \times r)$$
$$V = \frac{1}{3}\pi(27r^3)$$

Now, multiply by $\frac{1}{3}$:

$$V = \frac{27}{3}\pi r^3$$
$$V = 9\pi r^3$$

The volume of the cone is $9\pi r^3$.

Quick Tip

1. Identify the cone's actual radius (R) and height (H) based on the problem statement: $R_{\text{cone}} = 3r$ $H_{\text{cone}} = \text{radius of base} = 3r$ 2. Use the cone volume formula: $V = \frac{1}{3}\pi R_{\text{cone}}^2 H_{\text{cone}}$. 3. Substitute: $V = \frac{1}{3}\pi(3r)^2(3r)$. 4. Calculate: $V = \frac{1}{3}\pi(9r^2)(3r) = \frac{1}{3}\pi(27r^3) = 9\pi r^3$.

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18. If the ratio of volumes of two spheres is 1:8, then the ratio of their surface area is :

(1) 1:2

(2) 1:4

(3) 1:8

(4) 1:16

Correct Answer: (2) 1:4

Solution: Concept: This problem involves the formulas for the volume and surface area of a sphere.

- Volume of a sphere (V) = $\frac{4}{3}\pi r^3$, where r is the radius.
- Surface Area of a sphere (A) = $4\pi r^2$, where r is the radius.

Step 1: Set up the ratio of volumes Let the two spheres have radii r_1 and r_2, volumes V_1 and V_2, and surface areas A_1 and A_2. We are given that the ratio of their volumes is 1:8.

$$\frac{V_1}{V_2} = \frac{1}{8}$$

Substitute the volume formula:

$$\frac{\frac{4}{3}\pi r_1^3}{\frac{4}{3}\pi r_2^3} = \frac{1}{8}$$

The term $\frac{4}{3}\pi$ cancels out from the numerator and denominator:

$$\frac{r_1^3}{r_2^3} = \frac{1}{8}$$

This can be written as:

$$\left(\frac{r_1}{r_2}\right)^3 = \frac{1}{8}$$

Step 2: Find the ratio of their radii To find the ratio of the radii $\frac{r_1}{r_2}$, take the cube root of both sides:

$$\frac{r_1}{r_2} = \sqrt[3]{\frac{1}{8}}$$

Since $\sqrt[3]{1} = 1$ and $\sqrt[3]{8} = 2$ (because $2^3 = 8$),

$$\frac{r_1}{r_2} = \frac{1}{2}$$

So, the ratio of their radii is 1:2.

Step 3: Find the ratio of their surface areas The surface area of a sphere is $A = 4\pi r^2$. We need to find the ratio $\frac{A_1}{A_2}$:

$$\frac{A_1}{A_2} = \frac{4\pi r_1^2}{4\pi r_2^2}$$

The term 4π cancels out:

$$\frac{A_1}{A_2} = \frac{r_1^2}{r_2^2}$$

This can be written as:

$$\frac{A_1}{A_2} = \left(\frac{r_1}{r_2}\right)^2$$

Step 4: Substitute the ratio of radii From Step 2, we found $\frac{r_1}{r_2} = \frac{1}{2}$. Substitute this into the surface area ratio:

$$\frac{A_1}{A_2} = \left(\frac{1}{2}\right)^2$$

$$\frac{A_1}{A_2} = \frac{1^2}{2^2} = \frac{1}{4}$$

So, the ratio of their surface areas is 1:4.

Quick Tip

If the ratio of volumes of two similar 3D shapes (like spheres) is $V_1 : V_2 = k^3$, then the ratio of their corresponding linear dimensions (like radii) is $r_1 : r_2 = k$, and the ratio of their surface areas is $A_1 : A_2 = k^2$. Given $V_1 : V_2 = 1 : 8$. Since $8 = 2^3$, we have $k^3 = (1/2)^3$ if we consider $V_1/V_2 = 1/8$. This means the ratio of radii $r_1 : r_2 = 1 : 2$. (So $k = 1/2$) Then the ratio of surface areas $A_1 : A_2 = (r_1/r_2)^2 = (1/2)^2 = 1 : 4$.

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19. Let l be the lower class limit of a class-interval in a frequency distribution and m be the mid point of the class. Then the upper class limit of the class is :

(1) $m + \frac{l+m}{2}$

(2) $l + \frac{m+l}{2}$

(3) $2m - l$

(4) $m - 2l$

Correct Answer: (3) $2m - l$

Solution: Concept: In a frequency distribution, for any class interval:

- Let l be the lower class limit.
- Let u be the upper class limit.
- The mid-point (m) of the class interval is the average of the lower and upper class limits.

Mid-point formula: $m = \frac{l+u}{2}$

Step 1: Use the mid-point formula We are given the mid-point formula:

$$m = \frac{l+u}{2}$$

We are given l (lower class limit) and m (mid-point), and we need to find u (upper class limit).

Step 2: Rearrange the formula to solve for u Multiply both sides of the equation by 2:

$$2m = l + u$$

Now, isolate u by subtracting l from both sides:

$$2m - l = u$$

So, the upper class limit u is equal to $2m - l$.

Step 3: Compare with the options The derived expression for the upper class limit is $2m - l$. This matches option (3).

Example: Let a class interval be 10-20. Lower class limit, $l = 10$. Upper class limit, $u = 20$. Mid-point, $m = \frac{10+20}{2} = \frac{30}{2} = 15$. Now let's use the formula $u = 2m - l$ to check: $u = 2(15) - 10 = 30 - 10 = 20$. This is correct.

Quick Tip

The mid-point (m) of a class is exactly in the middle of the lower limit (l) and upper limit (u). So, $m = \frac{l+u}{2}$. To find the upper limit (u) if you know l and m: 1. Multiply mid-point by 2: $2m = l + u$ 2. Subtract the lower limit: $u = 2m - l$

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20. Two coins are tossed simultaneously. The probability of getting atmost one head is :

(1) $\frac{1}{4}$

(2) $\frac{3}{4}$

(3) $\frac{1}{2}$

(4) $\frac{5}{4}$

Correct Answer: (2) $\frac{3}{4}$

Solution: Concept: Probability is calculated as the ratio of the number of favorable outcomes to the total number of possible outcomes. Probability (P) = $\frac{\text{Number of Favorable Outcomes}}{\text{Total Number of Possible Outcomes}}$

Step 1: List all possible outcomes (Sample Space) When two coins are tossed simultaneously, let H denote a Head and T denote a Tail. The possible outcomes are:

- HH (Both heads)
- HT (First coin head, second coin tail)
- TH (First coin tail, second coin head)
- TT (Both tails)

The total number of possible outcomes is 4.

Step 2: Identify the favorable outcomes for "atmost one head" "Atmost one head" means zero heads OR one head. Let's list the outcomes that satisfy this condition:

- **Zero heads:** TT (This outcome has 0 heads, which is atmost 1 head)
- **One head:**
  - HT (This outcome has 1 head)

– TH (This outcome has 1 head)

The favorable outcomes are TT, HT, TH. The number of favorable outcomes is 3.

Alternative way to think about ”atmost one head”: It means we do NOT want the outcome with two heads (HH). So, favorable outcomes = Total outcomes - Outcome with two heads Favorable outcomes = {HH, HT, TH, TT} - {HH} = {HT, TH, TT}. Number of favorable outcomes = 3.

Step 3: Calculate the probability Probability (atmost one head) = $\frac{\text{Number of favorable outcomes}}{\text{Total number of possible outcomes}}$

$$P(\text{atmost one head}) = \frac{3}{4}$$

Step 4: Compare with the options The calculated probability is $\frac{3}{4}$. This matches option (2).

Quick Tip

When tossing two coins, the possible outcomes are: HH, HT, TH, TT (4 total outcomes). ”Atmost one head” means:

- 0 Heads: TT (1 outcome)
- 1 Head: HT, TH (2 outcomes)

Total favorable outcomes = 1 + 2 = 3. Probability = Favorable / Total = $3/4$. Alternatively, ”atmost one head” is the opposite of ”atleast two heads” (which means exactly two heads in this case: HH). P(HH) = 1/4. P(atmost one head) = 1 - P(HH) = $1 - 1/4 = 3/4$.

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21. If n is a natural number then $9^{2n} - 4^{2n}$ is always divisible by :

(1) 5

(2) 13

(3) both 5 and 13

(4) none of these

Correct Answer: (3) both 5 and 13

Solution: Concept: This problem involves testing divisibility of an algebraic expression. We can use algebraic identities or test small values of n. The relevant algebraic identity is the difference of squares: $a^2 - b^2 = (a-b)(a+b)$. More generally, $a^k - b^k$ is divisible by $(a-b)$ for any positive integer k, and divisible by $(a+b)$ if k is an even positive integer.

Step 1: Rewrite the expression The given expression is $9^{2n} - 4^{2n}$. We can rewrite 9^{2n} as $(9^2)^n = 81^n$. We can rewrite 4^{2n} as $(4^2)^n = 16^n$. So the expression becomes $81^n - 16^n$. Using the property that $a^n - b^n$ is always divisible by $(a-b)$: Here, $a = 81$ and $b = 16$. So, $81^n - 16^n$ is always divisible by $(81-16)$. $81 - 16 = 65$. The divisors of 65 are 1, 5, 13, 65. Therefore, $81^n - 16^n$ is always divisible by 5 and by 13.

Alternative Method using Difference of Squares repeatedly (if n allows) Let $x = 9^n$ and $y = 4^n$. Then the expression is $(9^n)^2 - (4^n)^2 = x^2 - y^2 = (x-y)(x+y)$. So, $9^{2n} - 4^{2n} = (9^n - 4^n)(9^n + 4^n)$.

Step 2: Test with small values of n (since n is a natural number, $n = 1, 2, 3, \ldots$) Case 1: Let $n = 1$. $9^{2(1)} - 4^{2(1)} = 9^2 - 4^2 = 81 - 16 = 65$. Is 65 divisible by 5? Yes, $65 = 5 \times 13$. Is 65 divisible by 13? Yes, $65 = 13 \times 5$. So for $n = 1$, the expression is divisible by both 5 and 13. Case 2: Let $n = 2$. $9^{2(2)} - 4^{2(2)} = 9^4 - 4^4 = (9^2)^2 - (4^2)^2 = 81^2 - 16^2$. Using the difference of squares formula $a^2 - b^2 = (a-b)(a+b)$: $81^2 - 16^2 = (81-16)(81+16) = (65)(97)$. Since 65 is a factor, and 65 is divisible by both 5 and 13, the expression $(65)(97)$ is also divisible by both 5 and 13.

Step 3: General Proof As shown in Step 1, the expression $9^{2n} - 4^{2n}$ can be written as $81^n - 16^n$. We know that for any positive integer n, $a^n - b^n$ is always divisible by $a - b$. Here $a = 81$ and $b = 16$. So $a - b = 81 - 16 = 65$. Since 65 is divisible by 5 (as $65 = 5 \times 13$) and divisible by 13 (as $65 = 13 \times 5$), it follows that $9^{2n} - 4^{2n}$ is always divisible by both 5 and 13.

Quick Tip

1. Rewrite the expression: $9^{2n} - 4^{2n} = (9^2)^n - (4^2)^n = 81^n - 16^n$. 2. Use the property: $a^n - b^n$ is always divisible by $a - b$. Here, $a = 81, b = 16$. So, $a - b = 81 - 16 = 65$. 3. Check divisors of 65: $65 = 5 \times 13$. Since the expression is divisible by 65, it must be divisible by 5 and by 13. Alternatively, test for $n = 1$: $9^2 - 4^2 = 81 - 16 = 65$. 65 is divisible by 5 and 13.

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22. If n is any natural number then $6^n - 5^n$ always ends with :

(1) 1

(2) 3

(3) 5

(4) 7

Correct Answer: (1) 1

Solution: Concept: We need to find the last digit (unit digit) of the expression $6^n - 5^n$ for any natural number n. This involves looking at the patterns of the last digits of powers of 6 and 5.

Step 1: Pattern of the last digit of powers of 6

- $6^1 = 6$
- $6^2 = 36$
- $6^3 = 216$
- $6^4 = 1296$

The last digit of any positive integer power of 6 is always 6. So, the last digit of 6^n is 6 for any natural number n.

Step 2: Pattern of the last digit of powers of 5

- $5^1 = 5$
- $5^2 = 25$
- $5^3 = 125$
- $5^4 = 625$

The last digit of any positive integer power of 5 is always 5. So, the last digit of 5^n is 5 for any natural number n.

Step 3: Find the last digit of $6^n - 5^n$ We are interested in the last digit of the result of the subtraction. Let L(N) denote the last digit of a number N. We need L($6^n - 5^n$). This is equivalent to L(L(6^n) - L(5^n)), considering potential borrowing if L(6^n) ¡ L(5^n).

We have: Last digit of 6^n is 6. Last digit of 5^n is 5.

So, we are looking at the last digit of a number ending in 6 minus a number ending in 5. Example: $\dots 6 - \dots 5$ The subtraction will result in a number ending in $6 - 5 = 1$. For instance:

- If $n = 1$: $6^1 - 5^1 = 6 - 5 = 1$. (Last digit is 1)
- If $n = 2$: $6^2 - 5^2 = 36 - 25 = 11$. (Last digit is 1)
- If $n = 3$: $6^3 - 5^3 = 216 - 125 = 91$. (Last digit is 1)

Since 6^n will always be greater than 5^n for natural number n, no borrowing issues affect the simple subtraction of the last digits. The last digit of $6^n - 5^n$ is always 1.

Quick Tip

1. Last digit of 6^n (for $n \geq 1$): $6^1 = 6, 6^2 = 36, 6^3 = 216, \dots$ The last digit is always 6.
2. Last digit of 5^n (for $n \geq 1$): $5^1 = 5, 5^2 = 25, 5^3 = 125, \dots$ The last digit is always 5.
3. We need the last digit of (a number ending in 6) - (a number ending in 5). This will be a number ending in $6 - 5 = 1$. Example for $n = 1$: $6 - 5 = 1$. Example for $n = 2$: $36 - 25 = 11$ (ends in 1).

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23. If α, β are the zeroes of the polynomial :
$f(x) = x^2 - p(x+1) - C$ **such that** $(\alpha + 1)(\beta + 1) = 0$ **then C=**
(1) 1
(2) 0
(3) -1
(4) 2

Correct Answer: (1) 1

Solution: Concept: For a quadratic polynomial $ax^2 + bx + c$, if α and β are its zeroes, then:

- Sum of zeroes: $\alpha + \beta = -b/a$
- Product of zeroes: $\alpha\beta = c/a$

Step 1: Rewrite the polynomial $f(x)$ in standard quadratic form $ax^2 + bx + c'$ The given polynomial is $f(x) = x^2 - p(x+1) - C$. Expand and rearrange: $f(x) = x^2 - px - p - C$ $f(x) = x^2 - px - (p + C)$ Comparing this with $ax^2 + bx + c'$: Here, $a = 1$, $b = -p$, and $c' = -(p + C)$.

Step 2: Use Vieta's formulas for sum and product of zeroes For $f(x) = x^2 - px - (p + C)$: Sum of zeroes: $\alpha + \beta = -(-p)/1 = p$. Product of zeroes: $\alpha\beta = -(p + C)/1 = -(p + C)$.

Step 3: Use the given condition $(\alpha + 1)(\beta + 1) = 0$ Expand the given condition: $(\alpha + 1)(\beta + 1) = \alpha\beta + \alpha(1) + 1(\beta) + (1)(1) = 0$ $\alpha\beta + \alpha + \beta + 1 = 0$ This can be written as: $(\alpha\beta) + (\alpha + \beta) + 1 = 0$

Step 4: Substitute the expressions for sum and product of zeroes into this equation We have: $\alpha\beta = -(p + C)$ $\alpha + \beta = p$ Substitute these into $(\alpha\beta) + (\alpha + \beta) + 1 = 0$:

$$-(p + C) + (p) + 1 = 0$$

Simplify the equation:

$$-p - C + p + 1 = 0$$

The terms $-p$ and $+p$ cancel out:

$$-C + 1 = 0$$

Step 5: Solve for C

$$1 = C$$

So, $C = 1$.

Quick Tip

1. Expand $f(x)$: $x^2 - p(x + 1) - C = x^2 - px - p - C = x^2 - px - (p + C)$. 2. From Vieta's formulas for $ax^2 + bx + c'$: Sum of roots: $\alpha + \beta = -(\text{coeff. of } x)/(\text{coeff. of } x^2) = -(-p)/1 = p$. Product of roots: $\alpha\beta = (\text{constant term})/(\text{coeff. of } x^2) = -(p + C)/1 = -(p + C)$. 3. Expand given condition: $(\alpha + 1)(\beta + 1) = \alpha\beta + \alpha + \beta + 1 = 0$. 4. Substitute: $-(p + C) + p + 1 = 0$. 5. Simplify: $-p - C + p + 1 = 0 \implies -C + 1 = 0 \implies C = 1$.

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24. In $\triangle ABC$ and $\triangle DEF$, $\angle A = \angle E = 40°$ and AB : ED = AC : EF and $\angle F = 65°$ then $\angle B = ?$

(1) 35°

(2) 65°

(3) 75°

(4) 85°

Correct Answer: (3) 75°

Solution: Concept: This problem involves similarity of triangles using the SAS (Side-Angle-Side) similarity criterion. If two triangles are similar, their corresponding angles are equal.

Step 1: Analyze the given information for similarity We are given:

1. $\angle A = \angle E = 40^\circ$ (One pair of corresponding angles is equal).
2. AB : ED = AC : EF. This can be rewritten as $\frac{AB}{ED} = \frac{AC}{EF}$. This means the ratio of sides including the angle A in $\triangle ABC$ (AB, AC) is equal to the ratio of sides including the angle E in $\triangle DEF$ (ED, EF).

The conditions $\angle A = \angle E$ and $\frac{AB}{ED} = \frac{AC}{EF}$ satisfy the SAS similarity criterion. The angle ($\angle A$ or $\angle E$) is included between the sides whose ratios are given. So, $\triangle ABC \sim \triangle EDF$. Note the order of vertices for similarity: A corresponds to E. Side AB (adjacent to A) corresponds to side ED (adjacent to E). Side AC (adjacent to A) corresponds to side EF (adjacent to E). Thus, the similarity is $\triangle ABC \sim \triangle EDF$.

Step 2: Corresponding angles in similar triangles Since $\triangle ABC \sim \triangle EDF$, their corresponding angles are equal:

- $\angle A = \angle E$ (Given as 40°)
- $\angle B = \angle D$ (This is what we need to find, or related to it)
- $\angle C = \angle F$

Step 3: Use the given angle $\angle F$ We are given $\angle F = 65^\circ$. From the similarity $\triangle ABC \sim \triangle EDF$, we have $\angle C = \angle F$. So, $\angle C = 65^\circ$.

Step 4: Find $\angle B$ using the sum of angles in $\triangle ABC$ The sum of angles in any triangle is 180°. In $\triangle ABC$: $\angle A + \angle B + \angle C = 180^\circ$. We know $\angle A = 40^\circ$ and we just found $\angle C = 65^\circ$.

Substitute these values:

$$40^\circ + \angle B + 65^\circ = 180^\circ$$

$$\angle B + 105^\circ = 180^\circ$$

$$\angle B = 180^\circ - 105^\circ$$

$$\angle B = 75^\circ$$

Therefore, $\angle B = 75^\circ$.

Quick Tip

1. Identify similarity: Given $\angle A = \angle E$ and $\frac{AB}{ED} = \frac{AC}{EF}$. This is SAS similarity for $\triangle ABC \sim \triangle EDF$. (Careful with the order of vertices in $\triangle DEF$). 2. Corresponding angles are equal: $\angle A \leftrightarrow \angle E$ $\angle B \leftrightarrow \angle D$ $\angle C \leftrightarrow \angle F$ 3. Use given $\angle F = 65^\circ$. So, $\angle C = 65^\circ$. 4. In $\triangle ABC$, $\angle A + \angle B + \angle C = 180^\circ$. $40^\circ + \angle B + 65^\circ = 180^\circ$. $\angle B + 105^\circ = 180^\circ$. $\angle B = 75^\circ$.

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25. If $\tan\theta = \frac{a}{b}$ **then** $\frac{a\sin\theta + b\cos\theta}{a\sin\theta - b\cos\theta} =$

(1) $\frac{a^2+b^2}{a^2-b^2}$

(2) $\frac{a^2-b^2}{a^2+b^2}$

(3) $\frac{a+b}{a-b}$

(4) $\frac{a-b}{a+b}$

Correct Answer: (1) $\frac{a^2+b^2}{a^2-b^2}$

Solution: Concept: This problem can be solved by expressing $\sin\theta$ and $\cos\theta$ in terms of a and b using a right-angled triangle, or by dividing the numerator and denominator of the expression by $\cos\theta$.

Method 1: Dividing by $\cos\theta$ Given expression: $\frac{a\sin\theta + b\cos\theta}{a\sin\theta - b\cos\theta}$ Divide both the numerator and the denominator by $\cos\theta$ (assuming $\cos\theta \neq 0$): Numerator: $\frac{a\sin\theta + b\cos\theta}{\cos\theta} = a\frac{\sin\theta}{\cos\theta} + b\frac{\cos\theta}{\cos\theta} = a\tan\theta + b$ Denominator: $\frac{a\sin\theta - b\cos\theta}{\cos\theta} = a\frac{\sin\theta}{\cos\theta} - b\frac{\cos\theta}{\cos\theta} = a\tan\theta - b$ So the expression becomes:

$$\frac{a\tan\theta + b}{a\tan\theta - b}$$

We are given $\tan\theta = \frac{a}{b}$. Substitute this into the expression:

$$\frac{a\left(\frac{a}{b}\right)+b}{a\left(\frac{a}{b}\right)-b} = \frac{\frac{a^2}{b}+b}{\frac{a^2}{b}-b}$$

To simplify the complex fraction, find a common denominator (b) for the terms in the numerator and denominator: Numerator: $\frac{a^2}{b}+\frac{b^2}{b}=\frac{a^2+b^2}{b}$ Denominator: $\frac{a^2}{b}-\frac{b^2}{b}=\frac{a^2-b^2}{b}$ So the expression is:

$$\frac{\left(\frac{a^2+b^2}{b}\right)}{\left(\frac{a^2-b^2}{b}\right)} = \frac{a^2+b^2}{b}\times\frac{b}{a^2-b^2} = \frac{a^2+b^2}{a^2-b^2}$$

Method 2: Using a right-angled triangle If $\tan\theta = \frac{a}{b}$, we can consider a right-angled triangle where: Opposite side = a Adjacent side = b Hypotenuse $h=\sqrt{a^2+b^2}$ Then, $\sin\theta = \frac{\text{Opposite}}{\text{Hypotenuse}} = \frac{a}{\sqrt{a^2+b^2}}$ And $\cos\theta = \frac{\text{Adjacent}}{\text{Hypotenuse}} = \frac{b}{\sqrt{a^2+b^2}}$ Substitute these into the expression $\frac{a\sin\theta+b\cos\theta}{a\sin\theta-b\cos\theta}$: Numerator: $a\left(\frac{a}{\sqrt{a^2+b^2}}\right)+b\left(\frac{b}{\sqrt{a^2+b^2}}\right)=\frac{a^2}{\sqrt{a^2+b^2}}+\frac{b^2}{\sqrt{a^2+b^2}}=\frac{a^2+b^2}{\sqrt{a^2+b^2}}$ Denominator: $a\left(\frac{a}{\sqrt{a^2+b^2}}\right)-b\left(\frac{b}{\sqrt{a^2+b^2}}\right)=\frac{a^2}{\sqrt{a^2+b^2}}-\frac{b^2}{\sqrt{a^2+b^2}}=\frac{a^2-b^2}{\sqrt{a^2+b^2}}$ The expression becomes:

$$\frac{\left(\frac{a^2+b^2}{\sqrt{a^2+b^2}}\right)}{\left(\frac{a^2-b^2}{\sqrt{a^2+b^2}}\right)} = \frac{a^2+b^2}{a^2-b^2}$$

Both methods yield $\frac{a^2+b^2}{a^2-b^2}$.

Quick Tip

Given an expression with $\sin\theta$ and $\cos\theta$, and knowing $\tan\theta$: A quick method is to divide the numerator and denominator by $\cos\theta$. This converts $\sin\theta$ to $\tan\theta$ and $\cos\theta$ to 1. Expression: $\frac{a\sin\theta+b\cos\theta}{a\sin\theta-b\cos\theta}$ Divide by $\cos\theta$: $\frac{a(\sin\theta/\cos\theta)+b(\cos\theta/\cos\theta)}{a(\sin\theta/\cos\theta)-b(\cos\theta/\cos\theta)}=\frac{a\tan\theta+b}{a\tan\theta-b}$. Substitute $\tan\theta=a/b$: $\frac{a(a/b)+b}{a(a/b)-b}=\frac{a^2/b+b}{a^2/b-b}=\frac{(a^2+b^2)/b}{(a^2-b^2)/b}=\frac{a^2+b^2}{a^2-b^2}$.

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26. If $\csc\theta = 2x$ **and** $\cot\theta = \frac{2}{x}$ **then** $2\left(x^2-\frac{1}{x^2}\right)=?$

(1) 1

(2) 0

(3) $\frac{1}{2}$

(4) -1

Correct Answer: (3) $\frac{1}{2}$

Solution: Concept: This problem uses the fundamental trigonometric identity relating cosecant ($\csc$) and cotangent ($\cot$):

$$\csc^2\theta - \cot^2\theta = 1$$

Step 1: Substitute the given expressions for $\csc\theta$ and $\cot\theta$ into the identity Given: $\csc\theta = 2x$ $\cot\theta = \frac{2}{x}$ Substitute these into $\csc^2\theta - \cot^2\theta = 1$:

$$(2x)^2 - \left(\frac{2}{x}\right)^2 = 1$$

Step 2: Simplify the equation

$$(2x)^2 = 4x^2$$
$$\left(\frac{2}{x}\right)^2 = \frac{2^2}{x^2} = \frac{4}{x^2}$$

So the equation becomes:

$$4x^2 - \frac{4}{x^2} = 1$$

Step 3: Factor out the common term to match the desired expression We need to find the value of $2\left(x^2 - \frac{1}{x^2}\right)$. Look at the equation from Step 2: $4x^2 - \frac{4}{x^2} = 1$. Factor out 4 from the left side:

$$4\left(x^2 - \frac{1}{x^2}\right) = 1$$

Step 4: Solve for the desired expression We want $2\left(x^2 - \frac{1}{x^2}\right)$. From $4\left(x^2 - \frac{1}{x^2}\right) = 1$, we can find $\left(x^2 - \frac{1}{x^2}\right)$ by dividing by 4:

$$x^2 - \frac{1}{x^2} = \frac{1}{4}$$

Now, multiply by 2:

$$2\left(x^2 - \frac{1}{x^2}\right) = 2 \times \frac{1}{4}$$
$$2\left(x^2 - \frac{1}{x^2}\right) = \frac{2}{4} = \frac{1}{2}$$

Thus, the value of the expression is $\frac{1}{2}$.

Quick Tip

1. Use the identity: $\csc^2\theta - \cot^2\theta = 1$. 2. Substitute given values: $\csc\theta = 2x \implies \csc^2\theta = (2x)^2 = 4x^2$. $\cot\theta = \frac{2}{x} \implies \cot^2\theta = \left(\frac{2}{x}\right)^2 = \frac{4}{x^2}$. 3. Plug into identity: $4x^2 - \frac{4}{x^2} = 1$. 4. Factor out 4: $4\left(x^2 - \frac{1}{x^2}\right) = 1$. 5. The question asks for $2\left(x^2 - \frac{1}{x^2}\right)$. Since $4\left(x^2 - \frac{1}{x^2}\right) = 1$, then $2 \times 2\left(x^2 - \frac{1}{x^2}\right) = 1$. So, $2\left(x^2 - \frac{1}{x^2}\right) = \frac{1}{2}$.

27. If the mean of 1,3,4,5,7,4 is m and mean of 3,2,2,4,3,3,p is $(m-1)$ and median is q then $p+q=$

(1) 4

(2) 5

(3) 6

(4) 7

Correct Answer: (4) 7

Solution: Concept:

- Mean = (Sum of observations) / (Number of observations).
- Median = Middle value of a dataset when arranged in order. If there are an even number of observations, the median is the average of the two middle values. If odd, it's the single middle value.

The term "median is q" is assumed to refer to the median of the second dataset.

Step 1: Calculate m (mean of the first dataset) First dataset: 1, 3, 4, 5, 7, 4. Number of observations (N_1) = 6. Sum of observations (S_1) = $1+3+4+5+7+4=24$. Mean $m=\frac{S_1}{N_1}=\frac{24}{6}=4$.

Step 2: Determine the mean of the second dataset The mean of the second dataset is given as $(m-1)$. Since $m=4$, the mean of the second dataset is $4-1=3$.

Step 3: Calculate p using the mean of the second dataset Second dataset: 3, 2, 2, 4, 3, 3, p. Number of observations (N_2) = 7. Sum of known observations in the second dataset = $3+2+2+4+3+3=17$. Sum of all observations in the second dataset (S_2) = $17+p$. Mean of second dataset = $\frac{S_2}{N_2}=\frac{17+p}{7}$. We know this mean is 3:

$$\frac{17+p}{7}=3$$

Multiply by 7: $17+p=3\times 7=21$.

$$p=21-17=4$$

So, $p=4$.

Step 4: Find q (median of the second dataset) The second dataset, with $p = 4$, is: 3, 2, 2, 4, 3, 3, 4. To find the median, first arrange the data in ascending order: 2, 2, 3, 3, 3, 4, 4. There are $N_2 = 7$ observations (an odd number). The median is the middle value, which is the $\left(\frac{N_2+1}{2}\right)^{th}$ term. Median position = $\left(\frac{7+1}{2}\right)^{th} = \left(\frac{8}{2}\right)^{th} = 4^{th}$ term. The 4th term in the ordered dataset (2, 2, 3, **3**, 3, 4, 4) is 3. So, the median $q = 3$.

Step 5: Calculate $p + q$ We found $p = 4$ and $q = 3$.

$$p + q = 4 + 3 = 7$$

Quick Tip

1. Calculate mean m of first set: $m = (1 + 3 + 4 + 5 + 7 + 4)/6 = 24/6 = 4$. 2. Mean of second set is $m - 1 = 4 - 1 = 3$. 3. Second set: 3,2,2,4,3,3,p. Sum = $17 + p$. Count = 7. So, $(17 + p)/7 = 3 \implies 17 + p = 21 \implies p = 4$. 4. Second set becomes: 3,2,2,4,3,3,4. Ordered: 2,2,3,3,3,4,4. Median q (middle value of 7 terms) is the 4th term, which is 3. 5. $p + q = 4 + 3 = 7$.

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28. If $\sin\alpha$ and $\cos\alpha$ are the roots of equation : $ax^2 + bx + c = 0$ then $b^2 =$

(1) $a^2 - 2ac$

(2) $a^2 + 2ac$

(3) $a^2 - ac$

(4) $a^2 + ac$

Correct Answer: (2) $a^2 + 2ac$

Solution: Concept: For a quadratic equation $Ax^2 + Bx + C = 0$, if the roots are r_1 and r_2, then:

- Sum of roots: $r_1 + r_2 = -B/A$
- Product of roots: $r_1 r_2 = C/A$

We also use the fundamental trigonometric identity: $\sin^2\alpha + \cos^2\alpha = 1$.

Step 1: Apply Vieta's formulas to the given equation The given equation is $ax^2 + bx + c = 0$. The roots are given as $\sin\alpha$ and $\cos\alpha$. Comparing with $Ax^2 + Bx + C = 0$, we have $A = a, B = b, C = c$. Sum of roots:

$$\sin\alpha + \cos\alpha = -\frac{b}{a} \quad \cdots(1)$$

Product of roots:

$$\sin\alpha\cos\alpha = \frac{c}{a} \quad \cdots(2)$$

Step 2: Use the identity $(x + y)^2 = x^2 + y^2 + 2xy$ We want to find an expression involving b^2. Notice that b appears in the sum of roots. Let's square the equation (1):

$$(\sin\alpha + \cos\alpha)^2 = \left(-\frac{b}{a}\right)^2$$

Expand the left side:

$$\sin^2\alpha + \cos^2\alpha + 2\sin\alpha\cos\alpha = \frac{b^2}{a^2}$$

Step 3: Substitute known identities and expressions We know the trigonometric identity: $\sin^2\alpha + \cos^2\alpha = 1$. From equation (2), we know: $\sin\alpha\cos\alpha = \frac{c}{a}$. Substitute these into the expanded equation from Step 2:

$$1 + 2\left(\frac{c}{a}\right) = \frac{b^2}{a^2}$$

Step 4: Solve for b^2

$$1 + \frac{2c}{a} = \frac{b^2}{a^2}$$

To clear the denominators, multiply the entire equation by a^2:

$$a^2\left(1 + \frac{2c}{a}\right) = a^2\left(\frac{b^2}{a^2}\right)$$

$$a^2(1) + a^2\left(\frac{2c}{a}\right) = b^2$$

$$a^2 + 2ac = b^2$$

So, $b^2 = a^2 + 2ac$.

Quick Tip

Given roots $r_1 = \sin\alpha, r_2 = \cos\alpha$ for $ax^2 + bx + c = 0$. 1. Sum of roots: $\sin\alpha + \cos\alpha = -b/a$. 2. Product of roots: $\sin\alpha\cos\alpha = c/a$. 3. Square the sum of roots: $(\sin\alpha + \cos\alpha)^2 = (-b/a)^2$. $\sin^2\alpha + \cos^2\alpha + 2\sin\alpha\cos\alpha = b^2/a^2$. 4. Substitute identity $\sin^2\alpha + \cos^2\alpha = 1$ and product from step 2: $1 + 2(c/a) = b^2/a^2$. 5. Multiply by a^2: $a^2 + 2ac = b^2$.

29. What will be the sum of n terms of the following arithmetic progression if the arithmetic progression is: $(x-y)^2, (x^2+y^2), (x+y)^2, \ldots\ldots\ldots$ **up to n terms:**

(1) $n\{(x-y)^2+(n-1)xy\}$

(2) $\{n(x-y)^2+n(x+y)^2\}$

(3) $\frac{n}{2}(x^2+y^2)-(x+y)$

(4) None of these

Correct Answer: (1) $n\{(x-y)^2+(n-1)xy\}$

Solution: Concept: For an arithmetic progression (AP), the sum of the first n terms is given by the formula: $S_n = \frac{n}{2}[2a_1+(n-1)d]$ where a_1 is the first term and d is the common difference. First, we need to verify if the given series is indeed an arithmetic progression and find its common difference.

Step 1: Expand the terms of the given series Term 1: $a_1 = (x-y)^2 = x^2-2xy+y^2$ Term 2: $a_2 = x^2+y^2$ Term 3: $a_3 = (x+y)^2 = x^2+2xy+y^2$

Step 2: Calculate the difference between consecutive terms to find the common difference d Difference $d_1 = a_2 - a_1$: $d_1 = (x^2+y^2)-(x^2-2xy+y^2)$
$d_1 = x^2+y^2-x^2+2xy-y^2$ $d_1 = 2xy$
Difference $d_2 = a_3 - a_2$: $d_2 = (x^2+2xy+y^2)-(x^2+y^2)$ $d_2 = x^2+2xy+y^2-x^2-y^2$ $d_2 = 2xy$
Since $d_1 = d_2 = 2xy$, the series is an arithmetic progression with: First term, $a_1 = (x-y)^2$
Common difference, $d = 2xy$

Step 3: Calculate the sum of n terms (S_n) Using the formula $S_n = \frac{n}{2}[2a_1+(n-1)d]$:
Substitute $a_1 = (x-y)^2$ and $d = 2xy$:

$$S_n = \frac{n}{2}[2(x-y)^2+(n-1)(2xy)]$$

We can factor out a 2 from the terms inside the main bracket:

$$S_n = \frac{n}{2}\cdot 2[(x-y)^2+(n-1)xy]$$

$$S_n = n[(x-y)^2+(n-1)xy]$$

Step 4: Compare the calculated sum S_n with the given options The calculated sum is $S_n = n[(x-y)^2+(n-1)xy]$.

- **(1)** $n\{(x-y)^2+(n-1)xy\}$**:** This option is identical to our calculated sum S_n.
- **(2)** $\{n(x-y)^2+n(x+y)^2\}$**:** This does not match. It simplifies to $n[(x-y)^2+(x+y)^2]=n[x^2-2xy+y^2+x^2+2xy+y^2]=n[2x^2+2y^2]=2n(x^2+y^2)$.
- **(3)** $\frac{n}{2}(x^2+y^2)-(x+y)$**:** This does not resemble the sum formula for this AP.
- **(4) None of these:** Since option (1) matches, this is incorrect.

Therefore, option (1) is the correct sum.

Quick Tip

1. Identify the first term a_1 and common difference d of the AP. $a_1=(x-y)^2$. $a_2=x^2+y^2$. Common difference $d=a_2-a_1=(x^2+y^2)-(x^2-2xy+y^2)=2xy$. (Check: $a_3-a_2=(x+y)^2-(x^2+y^2)=(x^2+2xy+y^2)-(x^2+y^2)=2xy$. So it's an AP). 2. Use the sum formula for an AP: $S_n=\frac{n}{2}[2a_1+(n-1)d]$. Substitute $a_1=(x-y)^2$ and $d=2xy$: $S_n=\frac{n}{2}[2(x-y)^2+(n-1)(2xy)]$. 3. Factor out 2 from the terms in the square brackets: $S_n=\frac{n}{2}\cdot 2[(x-y)^2+(n-1)xy]$. 4. Simplify: $S_n=n[(x-y)^2+(n-1)xy]$. This directly matches option (1).

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30. If $\frac{5+9+13+\text{......... to } n \text{ terms}}{7+9+11+\text{......... to } (n+1) \text{ terms}}=\frac{17}{16}$ **then** $n=$

(1) 8

(2) 7

(3) 10

(4) 11

Correct Answer: (2) 7

Solution: Concept: This problem involves finding the sum of terms in two different arithmetic progressions (APs) and then solving an equation based on the ratio of these sums. The sum of an AP is given by $S_k=\frac{k}{2}[2a+(k-1)d]$, where k is the number of terms, a is the first term, and d is the common difference.

Step 1: Find the sum of the numerator AP (S_N) The numerator series is $5+9+13+\dots$ to n terms. This is an AP with:

- First term, $a_N = 5$
- Common difference, $d_N = 9 - 5 = 4$
- Number of terms = n

Sum of numerator, $S_N = \frac{n}{2}[2a_N + (n-1)d_N]$ $S_N = \frac{n}{2}[2(5) + (n-1)4]$ $S_N = \frac{n}{2}[10 + 4n - 4]$ $S_N = \frac{n}{2}[4n + 6]$ $S_N = n(2n + 3)$

Step 2: Find the sum of the denominator AP (S_D) The denominator series is $7 + 9 + 11 + \ldots$ to $(n+1)$ terms. This is an AP with:

- First term, $a_D = 7$
- Common difference, $d_D = 9 - 7 = 2$
- Number of terms = $(n + 1)$

Sum of denominator, $S_D = \frac{n+1}{2}[2a_D + ((n+1) - 1)d_D]$ $S_D = \frac{n+1}{2}[2(7) + (n)(2)]$ $S_D = \frac{n+1}{2}[14 + 2n]$ $S_D = \frac{n+1}{2} \cdot 2(7 + n)$ $S_D = (n+1)(n+7)$

Step 3: Set up the equation using the given ratio We are given that $\frac{S_N}{S_D} = \frac{17}{16}$. Substitute the expressions for S_N and S_D:

$$\frac{n(2n+3)}{(n+1)(n+7)} = \frac{17}{16}$$

Step 4: Solve for n Cross-multiply:

$$16n(2n + 3) = 17(n + 1)(n + 7)$$

Expand both sides:

$$16(2n^2 + 3n) = 17(n^2 + 7n + n + 7)$$

$$32n^2 + 48n = 17(n^2 + 8n + 7)$$

$$32n^2 + 48n = 17n^2 + 136n + 119$$

Bring all terms to one side to form a quadratic equation:

$$32n^2 - 17n^2 + 48n - 136n - 119 = 0$$

$$15n^2 - 88n - 119 = 0$$

We can solve this quadratic equation or test the given options. Let's test $n = 7$ (option 2, which is circled in the image). If $n = 7$: Left side: $15(7)^2 - 88(7) - 119 = 15(49) - 616 - 119$

$= 735 - 616 - 119 = 735 - (616 + 119) = 735 - 735 = 0$ Since the equation holds true for $n = 7$, this is a valid solution. As n represents the number of terms, it must be a positive integer. The other root of this quadratic equation is $n = -119/(15 \times 7) = -17/15$, which is not a natural number. Thus, $n = 7$ is the correct answer.
Alternatively, after Step 3, with the equation $\frac{n(2n+3)}{(n+1)(n+7)} = \frac{17}{16}$, we could directly substitute $n = 7$: Numerator for $n = 7$: $S_N = 7(2(7) + 3) = 7(14 + 3) = 7(17) = 119$. Denominator for $n = 7$: $S_D = (7 + 1)(7 + 7) = (8)(14) = 112$. Ratio: $\frac{119}{112}$. To simplify $\frac{119}{112}$, we can see if they have common factors. Both are divisible by 7: $119 \div 7 = 17$ $112 \div 7 = 16$ So, $\frac{119}{112} = \frac{17}{16}$. This matches the given ratio.

Quick Tip

1. Sum of Numerator AP ($a = 5, d = 4$): $S_N = \frac{n}{2}(10 + (n - 1)4) = n(2n + 3)$. 2. Sum of Denominator AP ($a = 7, d = 2$, terms=$n + 1$): $S_D = \frac{n+1}{2}(14 + n \cdot 2) = (n + 1)(n + 7)$. 3. Equation: $\frac{n(2n+3)}{(n+1)(n+7)} = \frac{17}{16}$. 4. Test options. For $n = 7$: Numerator sum = $7(14 + 3) = 7(17) = 119$. Denominator sum = $(7 + 1)(7 + 7) = 8(14) = 112$. Ratio = $119/112$. Divide by 7: $(119/7)/(112/7) = 17/16$. Matches. So, $n = 7$ is correct.

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31. In figure : AD = 4 cm, BD = 3 cm, CB = 12 cm, then $\cot\theta =$

(1) $\frac{3}{4}$

(2) $\frac{5}{12}$

(3) $\frac{4}{3}$

(4) $\frac{12}{5}$

Correct Answer: (4) $\frac{12}{5}$

Solution: Concept: This problem requires using the Pythagorean theorem in one right-angled triangle to find a common side, and then applying the definition of the cotangent trigonometric ratio in another right-angled triangle. The interpretation of the figure is key.
Interpretation based on achieving a standard answer from the options: We assume the figure represents two right-angled triangles sharing a common side or related dimensions such that all given values are used sequentially.

- Assume $\triangle ABD$ is a right-angled triangle, with the right angle at D ($\angle ADB = 90°$). This means AD and BD are legs, and AB is the hypotenuse.

- Assume $\triangle ABC$ is a right-angled triangle, with the right angle at B ($\angle ABC = 90°$). This means AB and BC are legs, and AC is the hypotenuse. The diagram explicitly shows a right angle at B.

- The angle θ is given as $\angle ACB$.

Step 1: Find the length of side AB using $\triangle ABD$ Given AD = 4 cm, BD = 3 cm. Assuming $\triangle ABD$ is right-angled at D: By the Pythagorean theorem: $AB^2 = AD^2 + BD^2$
$AB^2 = 4^2 + 3^2$ $AB^2 = 16 + 9$ $AB^2 = 25$ $AB = \sqrt{25} = 5$ cm.
Step 2: Calculate $\cot\theta$ **using** $\triangle ABC$ Now consider the right-angled $\triangle ABC$, with $\angle ABC = 90°$. The angle in question is $\theta = \angle ACB$. For this angle θ:

- The side opposite to θ is AB.

- The side adjacent to θ is BC.

From Step 1, we found AB = 5 cm. We are given CB (which is BC) = 12 cm. The definition of the cotangent of an angle in a right-angled triangle is $\cot\theta = \frac{\text{Adjacent side}}{\text{Opposite side}}$.

$$\cot\theta = \frac{BC}{AB}$$

Substitute the values:

$$\cot\theta = \frac{12}{5}$$

This matches option (4).
Note on Figure Interpretation: The figure has explicit right-angle symbols at B (within $\triangle ABC$) and at D (making AD perpendicular to CD). If $\angle ADC = 90°$ and $\angle ABC = 90°$, the calculation is different and leads to $AB = \sqrt{97}$ and $\cot\theta = \frac{12}{\sqrt{97}}$, which is not among the options. The interpretation where $\triangle ABD$ is right-angled at D (so AD and BD are legs) and $\triangle ABC$ is right-angled at B (as marked) is a common setup in problems designed to use Pythagorean triples sequentially and leads to a given option.

Quick Tip

1. Focus on finding side AB first. Assume $\triangle ABD$ is a right-angled triangle at D. Given AD=4, BD=3. By Pythagoras: $AB = \sqrt{AD^2 + BD^2} = \sqrt{4^2 + 3^2} = \sqrt{16+9} = \sqrt{25} = 5$. 2. Now consider $\triangle ABC$. The figure shows it's right-angled at B. We have AB=5 and BC=12. The angle is $\theta = \angle C$. 3. $\cot\theta = \frac{\text{Adjacent side to }\theta}{\text{Opposite side to }\theta} = \frac{BC}{AB}$. 4. Substitute values: $\cot\theta = \frac{12}{5}$.

---

32. If $\sin 3\theta = \cos(\theta - 6^\circ)$, where (3θ) and $(\theta - 6^\circ)$ are both acute angles then the value of θ is :

(1) 18°

(2) 24°

(3) 36°

(4) 30°

Correct Answer: (2) 24°

Solution: Concept: This problem uses the complementary angle identity for sine and cosine: $\sin x = \cos(90^\circ - x)$ or $\cos y = \sin(90^\circ - y)$. If $\sin A = \cos B$, and A and B are acute angles, then $A + B = 90^\circ$.

Step 1: Apply the complementary angle relationship We are given $\sin 3\theta = \cos(\theta - 6^\circ)$. Since 3θ and $(\theta - 6^\circ)$ are acute angles, we can use the property that if $\sin A = \cos B$, then $A + B = 90^\circ$. Let $A = 3\theta$ and $B = \theta - 6^\circ$. So,

$$A + B = 90^\circ$$

$$3\theta + (\theta - 6^\circ) = 90^\circ$$

Step 2: Solve for θ

$$3\theta + \theta - 6^\circ = 90^\circ$$

$$4\theta - 6^\circ = 90^\circ$$

Add 6° to both sides:

$$4\theta = 90^\circ + 6^\circ$$

$$4\theta = 96^\circ$$

Divide by 4:

$$\theta = \frac{96^\circ}{4}$$

$$\theta = 24^\circ$$

Step 3: Verify that the angles are acute If $\theta = 24^\circ$:

- $3\theta = 3 \times 24^\circ = 72^\circ$. This is an acute angle (between 0° and 90°).
- $\theta - 6^\circ = 24^\circ - 6^\circ = 18^\circ$. This is also an acute angle.

The conditions are satisfied.

Therefore, the value of θ is 24°.

Quick Tip

Use the identity: If $\sin A = \cos B$, then $A + B = 90^\circ$ (for acute A, B). Here, $A = 3\theta$ and $B = \theta - 6^\circ$. So, $3\theta + (\theta - 6^\circ) = 90^\circ$. $4\theta - 6^\circ = 90^\circ$. $4\theta = 96^\circ$. $\theta = 24^\circ$. Always check if the resulting angles 3θ and $\theta - 6^\circ$ are indeed acute. $3(24^\circ) = 72^\circ$ (acute). $24^\circ - 6^\circ = 18^\circ$ (acute).

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33. $\sin(60^\circ + \theta) - \cos(30^\circ - \theta)$ **is equal to :**

(1) $2\cos\theta$

(2) $2\sin\theta$

(3) 0

(4) 1

Correct Answer: (3) 0

Solution: Concept: This problem uses the complementary angle identity $\cos x = \sin(90^\circ - x)$.

Step 1: Apply the complementary angle identity to one of the terms We need to evaluate $\sin(60^\circ + \theta) - \cos(30^\circ - \theta)$. Let's convert the cosine term to a sine term. We know that $\cos x = \sin(90^\circ - x)$. Let $x = 30^\circ - \theta$. Then, $\cos(30^\circ - \theta) = \sin(90^\circ - (30^\circ - \theta))$. Simplify the angle inside the sine function: $90^\circ - (30^\circ - \theta) = 90^\circ - 30^\circ + \theta = 60^\circ + \theta$. So, $\cos(30^\circ - \theta) = \sin(60^\circ + \theta)$.

Step 2: Substitute this back into the original expression The original expression is $\sin(60^\circ + \theta) - \cos(30^\circ - \theta)$. Substitute $\cos(30^\circ - \theta) = \sin(60^\circ + \theta)$:

$$\sin(60^\circ + \theta) - \sin(60^\circ + \theta)$$

Step 3: Simplify

$$\sin(60^\circ + \theta) - \sin(60^\circ + \theta) = 0$$

The expression is equal to 0.

Alternative Method: Convert sine to cosine We know $\sin y = \cos(90^\circ - y)$. Let $y = 60^\circ + \theta$. Then $\sin(60^\circ + \theta) = \cos(90^\circ - (60^\circ + \theta))$. $90^\circ - (60^\circ + \theta) = 90^\circ - 60^\circ - \theta = 30^\circ - \theta$. So, $\sin(60^\circ + \theta) = \cos(30^\circ - \theta)$. Substitute this into the original expression:

$$\cos(30^\circ - \theta) - \cos(30^\circ - \theta) = 0$$

Both methods yield 0.

Quick Tip

Use the complementary angle identity: $\cos A = \sin(90^\circ - A)$ or $\sin B = \cos(90^\circ - B)$. Let's change $\cos(30^\circ - \theta)$. $\cos(30^\circ - \theta) = \sin(90^\circ - [30^\circ - \theta]) = \sin(90^\circ - 30^\circ + \theta)$ $= \sin(60^\circ + \theta)$. So the expression becomes: $\sin(60^\circ + \theta) - \sin(60^\circ + \theta) = 0$.

---

34. If $\sin A + \sin^2 A = 1$**, then the value of** $\cos^2 A + \cos^4 A$ **is :**

(1) 2

(2) 1

(3) -2

(4) 0

Correct Answer: (2) 1

Solution: Concept: This problem uses the fundamental trigonometric identity $\sin^2 A + \cos^2 A = 1$.

Step 1: Manipulate the given equation We are given: $\sin A + \sin^2 A = 1$. Rearrange this equation to isolate $\sin A$:

$$\sin A = 1 - \sin^2 A$$

Step 2: Use the fundamental identity We know that $\sin^2 A + \cos^2 A = 1$. From this identity, we can write $\cos^2 A = 1 - \sin^2 A$.

Step 3: Substitute to find a relationship between $\sin A$ and $\cos^2 A$ From Step 1, we have $\sin A = 1 - \sin^2 A$. From Step 2, we have $\cos^2 A = 1 - \sin^2 A$. Since both $\sin A$ and $\cos^2 A$ are equal to $1 - \sin^2 A$, we can conclude:

$$\sin A = \cos^2 A$$

Step 4: Evaluate the expression $\cos^2 A + \cos^4 A$ We need to find the value of $\cos^2 A + \cos^4 A$. We can write $\cos^4 A$ as $(\cos^2 A)^2$. So the expression is $\cos^2 A + (\cos^2 A)^2$. From Step 3, we found that $\cos^2 A = \sin A$. Substitute this into the expression:

$$\cos^2 A + \cos^4 A = \sin A + (\sin A)^2$$

$$\cos^2 A + \cos^4 A = \sin A + \sin^2 A$$

Step 5: Use the original given condition The original given condition was $\sin A + \sin^2 A = 1$. Therefore,

$$\cos^2 A + \cos^4 A = 1$$

The value of the expression is 1.

Quick Tip

1. Given: $\sin A + \sin^2 A = 1$. 2. Rearrange: $\sin A = 1 - \sin^2 A$. 3. Use identity: $1 - \sin^2 A = \cos^2 A$. 4. So, from (2) and (3): $\sin A = \cos^2 A$. 5. We need to find $\cos^2 A + \cos^4 A$. This can be written as $\cos^2 A + (\cos^2 A)^2$. 6. Substitute $\cos^2 A = \sin A$: Expression becomes $\sin A + (\sin A)^2 = \sin A + \sin^2 A$. 7. From the given condition (1), $\sin A + \sin^2 A = 1$. Thus, $\cos^2 A + \cos^4 A = 1$.

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35. If the equation : $(a^2 + b^2)x^2 - 2(ac + bd)x + (c^2 + d^2) = 0$ has equal roots then :

(1) $ab = cd$

(2) $ad = bc$

(3) $ad = \sqrt{bc}$

(4) $ab = \sqrt{cd}$

Correct Answer: (2) $ad = bc$

Solution: Concept: For a quadratic equation $Ax^2 + Bx + C = 0$, the roots are equal if and only if its discriminant (D) is equal to zero. The discriminant is given by $D = B^2 - 4AC$.
Step 1: Identify A, B, and C from the given quadratic equation The given equation is $(a^2 + b^2)x^2 - 2(ac + bd)x + (c^2 + d^2) = 0$. Comparing with $Ax^2 + Bx + C = 0$:

- $A = a^2 + b^2$
- $B = -2(ac + bd)$
- $C = c^2 + d^2$

Step 2: Set the discriminant $D = B^2 - 4AC$ to zero Since the equation has equal roots, $D = 0$.

$$[-2(ac + bd)]^2 - 4(a^2 + b^2)(c^2 + d^2) = 0$$

Step 3: Expand and simplify the equation

$$4(ac + bd)^2 - 4(a^2 + b^2)(c^2 + d^2) = 0$$

Divide the entire equation by 4:

$$(ac + bd)^2 - (a^2 + b^2)(c^2 + d^2) = 0$$

Expand $(ac + bd)^2$: $(ac + bd)^2 = (ac)^2 + (bd)^2 + 2(ac)(bd) = a^2c^2 + b^2d^2 + 2abcd$ Expand $(a^2 + b^2)(c^2 + d^2)$: $(a^2 + b^2)(c^2 + d^2) = a^2c^2 + a^2d^2 + b^2c^2 + b^2d^2$ Now substitute these back into the equation:

$$(a^2c^2 + b^2d^2 + 2abcd) - (a^2c^2 + a^2d^2 + b^2c^2 + b^2d^2) = 0$$

$$a^2c^2 + b^2d^2 + 2abcd - a^2c^2 - a^2d^2 - b^2c^2 - b^2d^2 = 0$$

Cancel out terms: a^2c^2 cancels with $-a^2c^2$, and b^2d^2 cancels with $-b^2d^2$. We are left with:

$$2abcd - a^2d^2 - b^2c^2 = 0$$

Multiply by -1 to make the squared terms positive (optional, but helps in recognizing a pattern):

$$a^2d^2 + b^2c^2 - 2abcd = 0$$

Step 4: Factor the resulting expression The expression $a^2d^2 - 2abcd + b^2c^2$ is a perfect square trinomial. It can be written as $(ad)^2 - 2(ad)(bc) + (bc)^2$. This is of the form $(X - Y)^2 = X^2 - 2XY + Y^2$, where $X = ad$ and $Y = bc$. So, $(ad - bc)^2 = 0$.
Step 5: Solve for the condition If $(ad - bc)^2 = 0$, then taking the square root of both sides gives:

$$ad - bc = 0$$

$$ad = bc$$

This is the condition for equal roots.

> **Quick Tip**
>
> Condition for equal roots of $Ax^2+Bx+C=0$ is Discriminant $D = B^2-4AC = 0$. Here, $A = (a^2+b^2)$, $B = -2(ac+bd)$, $C = (c^2+d^2)$. $B^2 = 4(ac+bd)^2 = 4(a^2c^2+b^2d^2+2abcd)$. $4AC = 4(a^2+b^2)(c^2+d^2) = 4(a^2c^2+a^2d^2+b^2c^2+b^2d^2)$. Set $B^2 = 4AC$ (since $B^2-4AC=0$): $4(a^2c^2+b^2d^2+2abcd) = 4(a^2c^2+a^2d^2+b^2c^2+b^2d^2)$. Divide by 4 and simplify: $a^2c^2+b^2d^2+2abcd = a^2c^2+a^2d^2+b^2c^2+b^2d^2$. $2abcd = a^2d^2+b^2c^2$. Rearrange: $a^2d^2-2abcd+b^2c^2 = 0$. This is $(ad-bc)^2 = 0$. So, $ad - bc = 0 \implies ad = bc$.

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36. If a and b can take values 1,2,3,4. Then the number of the equations of the form : $ax^2 + bx + 1 = 0$ having real roots is :

(1) 10

(2) 7

(3) 6

(4) 12

Correct Answer: (2) 7 (The image has '6' circled, but calculation leads to 7. I will use 7.)

Solution: Concept: For a quadratic equation $Ax^2 + Bx + C = 0$ to have real roots, its discriminant (D) must be greater than or equal to zero ($D \geq 0$). The discriminant is $D = B^2 - 4AC$.

Step 1: Identify A, B, C for the given equation form The equation is $ax^2 + bx + 1 = 0$. Here, $A = a$, $B = b$, $C = 1$.

Step 2: Set up the condition for real roots For real roots, $D = B^2 - 4AC \geq 0$. Substitute the values:

$$b^2 - 4(a)(1) \geq 0$$

$$b^2 - 4a \geq 0$$

$$b^2 \geq 4a$$

Step 3: Test possible values for a and b Given that a and b can take values from {1, 2, 3, 4}. We need to find pairs (a, b) that satisfy $b^2 \geq 4a$.

Let's iterate through values of a:

- **If $a = 1$:** We need $b^2 \geq 4(1) \implies b^2 \geq 4$. Possible values for b from {1, 2, 3, 4}:
  - $b = 1 \implies 1^2 = 1$. ($1 \ngeq 4$) - No
  - $b = 2 \implies 2^2 = 4$. ($4 \geq 4$) - Yes. Pair: (a=1, b=2)
  - $b = 3 \implies 3^2 = 9$. ($9 \geq 4$) - Yes. Pair: (a=1, b=3)
  - $b = 4 \implies 4^2 = 16$. ($16 \geq 4$) - Yes. Pair: (a=1, b=4)

  (3 pairs for a=1)
- **If $a = 2$:** We need $b^2 \geq 4(2) \implies b^2 \geq 8$. Possible values for b from {1, 2, 3, 4}:
  - $b = 1 \implies 1^2 = 1$. ($1 \ngeq 8$) - No
  - $b = 2 \implies 2^2 = 4$. ($4 \ngeq 8$) - No
  - $b = 3 \implies 3^2 = 9$. ($9 \geq 8$) - Yes. Pair: (a=2, b=3)
  - $b = 4 \implies 4^2 = 16$. ($16 \geq 8$) - Yes. Pair: (a=2, b=4)

  (2 pairs for a=2)
- **If $a = 3$:** We need $b^2 \geq 4(3) \implies b^2 \geq 12$. Possible values for b from {1, 2, 3, 4}:
  - $b = 1 \implies 1^2 = 1$. ($1 \ngeq 12$) - No

- $b = 2 \implies 2^2 = 4$. $(4 \not\geq 12)$ - No
- $b = 3 \implies 3^2 = 9$. $(9 \not\geq 12)$ - No
- $b = 4 \implies 4^2 = 16$. $(16 \geq 12)$ - Yes. Pair: (a=3, b=4)

(1 pair for a=3)

- **If** $a = 4$**:** We need $b^2 \geq 4(4) \implies b^2 \geq 16$. Possible values for b from {1, 2, 3, 4}:
  - $b = 1 \implies 1^2 = 1$. $(1 \not\geq 16)$ - No
  - $b = 2 \implies 2^2 = 4$. $(4 \not\geq 16)$ - No
  - $b = 3 \implies 3^2 = 9$. $(9 \not\geq 16)$ - No
  - $b = 4 \implies 4^2 = 16$. $(16 \geq 16)$ - Yes. Pair: (a=4, b=4)

  (1 pair for a=4)

Step 4: Count the total number of valid pairs (a, b) Total number of equations = (pairs for a=1) + (pairs for a=2) + (pairs for a=3) + (pairs for a=4) Total = $3 + 2 + 1 + 1 = 7$. There are 7 such equations.
The image shows option (3) 6 circled. Let me recheck the counts. a=1: b=2,3,4 (3 pairs) a=2: b=3,4 (2 pairs) a=3: b=4 (1 pair) a=4: b=4 (1 pair) Total = 3+2+1+1 = 7. My calculation consistently gives 7. If the answer is 6, one of these pairs must be excluded, or one condition was missed. For example, if "real and distinct roots" was implied ($D > 0$), then cases where $b^2 = 4a$ would be excluded. $b^2 = 4a$ cases: (a=1, b=2) $\implies b^2 = 4, 4a = 4$. D=0. (a=4, b=4) $\implies b^2 = 16, 4a = 16$. D=0. If D¿0 is required, we exclude these 2 pairs, leaving 7-2 = 5 pairs. This does not match 6. The question states "real roots," which includes $D = 0$ (equal real roots) and $D > 0$ (distinct real roots). So my count of 7 should be correct for "real roots." Given the circled option in the image for many other problems has been correct, if the answer key says 6, there's a subtle point or a common misinterpretation I might be missing, or an error in the question/options/key. However, based on standard interpretation, the answer is 7. I will use 7 as the correct answer.

Quick Tip

For $ax^2 + bx + 1 = 0$ to have real roots, Discriminant $D = b^2 - 4a(1) \geq 0$, so $b^2 \geq 4a$. $a, b \in \{1, 2, 3, 4\}$. List pairs (a,b) satisfying $b^2 \geq 4a$:

- If $a = 1$ ($4a = 4$): $b^2 \geq 4 \implies b \in \{2, 3, 4\}$ (3 pairs)
- If $a = 2$ ($4a = 8$): $b^2 \geq 8 \implies b \in \{3, 4\}$ (2 pairs)
- If $a = 3$ ($4a = 12$): $b^2 \geq 12 \implies b \in \{4\}$ (1 pair)
- If $a = 4$ ($4a = 16$): $b^2 \geq 16 \implies b \in \{4\}$ (1 pair)

Total number of equations = $3 + 2 + 1 + 1 = 7$.

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37. If the sum of first n terms of an A.P. is $\frac{3n^2}{2} + \frac{5n}{2}$ then its 25th term is :

(1) 70

(2) $-n$

(3) 76

(4) none of these

Correct Answer: (3) 76

Solution: Concept: Let S_n be the sum of the first n terms of an A.P. Let a_n be the n^{th} term of the A.P. The n^{th} term can be found using the relation: $a_n = S_n - S_{n-1}$ (for $n > 1$). The first term $a_1 = S_1$. Alternatively, if S_n is a quadratic in n of the form $An^2 + Bn$, then it represents the sum of an AP. The common difference $d = 2A$ and the first term $a_1 = A + B$. Then $a_k = a_1 + (k-1)d$.

Method 1: Using $a_n = S_n - S_{n-1}$ Given $S_n = \frac{3n^2}{2} + \frac{5n}{2}$. We need the 25th term, a_{25}. So, $a_{25} = S_{25} - S_{24}$.

Calculate S_{25}: $S_{25} = \frac{3(25)^2}{2} + \frac{5(25)}{2} = \frac{3(625)}{2} + \frac{125}{2} = \frac{1875+125}{2} = \frac{2000}{2} = 1000$.

Calculate S_{24}: $S_{24} = \frac{3(24)^2}{2} + \frac{5(24)}{2} = \frac{3(576)}{2} + \frac{120}{2} = \frac{1728+120}{2} = \frac{1848}{2} = 924$.

Now, calculate a_{25}: $a_{25} = S_{25} - S_{24} = 1000 - 924 = 76$.

Method 2: Finding a_1 **and** d Given $S_n = \frac{3}{2}n^2 + \frac{5}{2}n$. This is in the form $An^2 + Bn$ where $A = \frac{3}{2}$ and $B = \frac{5}{2}$. First term, $a_1 = S_1 = \frac{3(1)^2}{2} + \frac{5(1)}{2} = \frac{3}{2} + \frac{5}{2} = \frac{8}{2} = 4$. (Using shortcut:

$a_1 = A + B = \frac{3}{2} + \frac{5}{2} = \frac{8}{2} = 4$). Common difference, $d = 2A = 2 \times \frac{3}{2} = 3$.

The k^{th} term of an A.P. is $a_k = a_1 + (k-1)d$. We need the 25th term ($k = 25$):

$a_{25} = a_1 + (25-1)d$ $a_{25} = 4 + (24)(3)$ $a_{25} = 4 + 72$ $a_{25} = 76$. Both methods yield 76.

Quick Tip

Method 1: $a_n = S_n - S_{n-1}$. $S_n = \frac{3n^2+5n}{2}$. $a_{25} = S_{25} - S_{24}$. $S_{25} = (3 \cdot 25^2 + 5 \cdot 25)/2 = (1875 + 125)/2 = 2000/2 = 1000$. $S_{24} = (3 \cdot 24^2 + 5 \cdot 24)/2 = (1728 + 120)/2 = 1848/2 = 924$. $a_{25} = 1000 - 924 = 76$.

Method 2: If $S_n = An^2 + Bn$, then $a_1 = A + B$ and common difference $d = 2A$. Here $A = 3/2, B = 5/2$. $a_1 = 3/2 + 5/2 = 8/2 = 4$. $d = 2(3/2) = 3$. $a_{25} = a_1 + (25-1)d = 4 + 24(3) = 4 + 72 = 76$.

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38. Sum of all 3 digit natural numbers which are divisible by 13 :

(1) 3774

(2) 37674

(3) 37697

(4) 37650

Correct Answer: (2) 37674

Solution: Concept: The 3-digit natural numbers divisible by 13 form an arithmetic progression (AP). We need to find the first term, the last term, the number of terms, and then use the sum formula for an AP. Sum of an AP: $S_n = \frac{n}{2}(a_1 + a_n)$, where a_1 is the first term, a_n is the last term, and n is the number of terms.

Step 1: Find the first 3-digit number divisible by 13 (a_1) The smallest 3-digit number is 100. Divide 100 by 13: $100 \div 13 \approx 7.69$. So, $13 \times 7 = 91$ (2-digit). The next multiple of 13 is $13 \times 8 = 104$. This is the first 3-digit number divisible by 13. So, $a_1 = 104$.

Step 2: Find the last 3-digit number divisible by 13 (a_n) The largest 3-digit number is 999. Divide 999 by 13: $999 \div 13 \approx 76.84$. So, consider 13×76.

$13 \times 76 = 13 \times (70 + 6) = 910 + 78 = 988$. This is the last 3-digit number divisible by 13. So, $a_n = 988$.

Step 3: Find the number of terms (n) The terms form an AP: 104, 117, ..., 988 with common difference $d = 13$. Let $a_n = a_1 + (n-1)d$. $988 = 104 + (n-1)13$.
$988 - 104 = (n-1)13$. $884 = (n-1)13$. $n - 1 = \frac{884}{13}$. $884 \div 13$: $13 \times 6 = 78$; $88 - 78 = 10$; bring down 4, so 104. $13 \times 8 = 104$. So, $n - 1 = 68$. $n = 68 + 1 = 69$. There are 69 such numbers.
Step 4: Calculate the sum (S_n) Using the formula $S_n = \frac{n}{2}(a_1 + a_n)$: $S_{69} = \frac{69}{2}(104 + 988)$
$S_{69} = \frac{69}{2}(1092)$ $S_{69} = 69 \times \frac{1092}{2}$ $S_{69} = 69 \times 546$
Calculation of 69×546: 546 x 69 —— 4914 (9×546) 32760 (60×546) —— 37674
So, the sum is 37674.

Quick Tip

1. Smallest 3-digit number: 100. First multiple of $13 \geq 100$: $13 \times 8 = 104$. $(a_1 = 104)$. 2. Largest 3-digit number: 999. Largest multiple of $13 \leq 999$: $999/13 \approx 76.84$. So $13 \times 76 = 988$. $(a_n = 988)$. 3. Number of terms n: Use $a_n = a_1 + (n-1)d$. $988 = 104 + (n-1)13 \implies 884 = (n-1)13 \implies n - 1 = 884/13 = 68 \implies n = 69$. 4. Sum $S_n = \frac{n}{2}(a_1 + a_n)$. $S_{69} = \frac{69}{2}(104 + 988) = \frac{69}{2}(1092) = 69 \times 546 = 37674$.

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39. A cone, a hemisphere and a cylinder stands on equal bases and have the same height. What is the ratio of their volumes :

(1) 1 : 2 : 3

(2) 2 : 3 : 4

(3) 1 : 3 : 4

(4) none of these

Correct Answer: (1) 1 : 2 : 3

Solution: Concept: This problem requires knowing the volume formulas for a cone, a hemisphere, and a cylinder, and applying the given conditions about equal bases and same height.
Step 1: Define dimensions based on conditions ”Equal bases” means they all have the same radius, let it be r. ”Same height” means they all have the same height, let it be h. For a hemisphere standing on its circular base, its height is equal to its radius. So, if the

hemisphere has the same height h as the cylinder and cone, then for the hemisphere, its radius must also be h. Since all three have equal bases (same radius r) AND same height (h):

- For the hemisphere: radius = height. So, $r_{hemisphere} = h_{hemisphere}$.
- Since all bases are equal, the radius of the cone and cylinder is also r.
- Since all heights are equal, the height of the cone and cylinder is also h.

The crucial condition is that for a hemisphere, its "height" (when standing on its base) *is* its radius. So, if they all have the same height h, then for the hemisphere, its radius must be h. And since they all have equal bases, the radius of the cone and cylinder must also be h. Thus, for all three shapes: radius $R = h$, and height $H = h$.

Step 2: Volume formulas

- Volume of Cone (V_c): $\frac{1}{3}\pi R^2 H$
- Volume of Hemisphere (V_h): $\frac{2}{3}\pi R^3$
- Volume of Cylinder (V_{cyl}): $\pi R^2 H$

Step 3: Substitute $R = h$ and $H = h$ into the formulas

- Cone: $V_c = \frac{1}{3}\pi(h)^2(h) = \frac{1}{3}\pi h^3$
- Hemisphere: Its radius is R. If it "stands on its base" and has height h, then its radius R must be equal to h. So, $V_h = \frac{2}{3}\pi(h)^3 = \frac{2}{3}\pi h^3$
- Cylinder: $V_{cyl} = \pi(h)^2(h) = \pi h^3$

Step 4: Find the ratio of their volumes Ratio $V_c : V_h : V_{cyl}$

$$\frac{1}{3}\pi h^3 : \frac{2}{3}\pi h^3 : \pi h^3$$

We can cancel out the common factor πh^3 from all terms (assuming $h \neq 0$):

$$\frac{1}{3} : \frac{2}{3} : 1$$

To get rid of the fractions, multiply all parts of the ratio by 3:

$$3 \times \frac{1}{3} : 3 \times \frac{2}{3} : 3 \times 1$$

$$1 : 2 : 3$$

So, the ratio of their volumes is 1 : 2 : 3.

Quick Tip

Given: Cone, Hemisphere, Cylinder have equal bases (radius R) and same height (H). Crucial point: For a hemisphere, its height (when on its flat base) IS its radius. So, $H = R$. This means for all three shapes, their radius is R and their height is also R.

- $V_{\text{cone}} = \frac{1}{3}\pi R^2 H = \frac{1}{3}\pi R^2(R) = \frac{1}{3}\pi R^3$
- $V_{\text{hemisphere}} = \frac{2}{3}\pi R^3$
- $V_{\text{cylinder}} = \pi R^2 H = \pi R^2(R) = \pi R^3$

Ratio: $\frac{1}{3}\pi R^3 : \frac{2}{3}\pi R^3 : \pi R^3$ Divide by πR^3: $\frac{1}{3} : \frac{2}{3} : 1$ Multiply by 3: $1 : 2 : 3$.

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40. The circumference of a circle is 100 cm. The side of a square inscribed in the circle is :

(1) $50\sqrt{2}$ cm

(2) $\frac{100}{\pi}$ cm

(3) $\frac{50\sqrt{2}}{\pi}$ cm

(4) $\frac{100\sqrt{2}}{\pi}$ cm

Correct Answer: (3) $\frac{50\sqrt{2}}{\pi}$ cm

Solution: Concept:

- Circumference of a circle (C) = $2\pi r$, where r is the radius.
- When a square is inscribed in a circle, the diagonal of the square is equal to the diameter of the circle.
- For a square with side s, its diagonal $d_{sq} = s\sqrt{2}$.
- Diameter of a circle (D_{circ}) = $2r$.

Step 1: Find the radius of the circle Given circumference $C = 100$ cm. We know $C = 2\pi r$. So, $100 = 2\pi r$. Solve for r:

$$r = \frac{100}{2\pi} = \frac{50}{\pi} \text{ cm}$$

Step 2: Relate the diagonal of the inscribed square to the diameter of the circle When a square is inscribed in a circle, its vertices lie on the circle. The diagonal of this square is equal to the diameter of the circle. Diameter of the circle, $D_{circ} = 2r$. Substitute $r = \frac{50}{\pi}$:

$$D_{circ} = 2 \times \frac{50}{\pi} = \frac{100}{\pi} \text{ cm}$$

So, the diagonal of the inscribed square, $d_{sq} = D_{circ} = \frac{100}{\pi}$ cm.

Step 3: Find the side of the square Let the side of the inscribed square be s. The diagonal of a square is related to its side by $d_{sq} = s\sqrt{2}$. We have $d_{sq} = \frac{100}{\pi}$. So, $s\sqrt{2} = \frac{100}{\pi}$. Solve for s:

$$s = \frac{100}{\pi\sqrt{2}}$$

To rationalize the denominator (optional but good practice, and helps match options): multiply numerator and denominator by $\sqrt{2}$.

$$s = \frac{100\sqrt{2}}{\pi\sqrt{2}\sqrt{2}} = \frac{100\sqrt{2}}{\pi \times 2}$$

$$s = \frac{50\sqrt{2}}{\pi} \text{ cm}$$

The side of the square inscribed in the circle is $\frac{50\sqrt{2}}{\pi}$ cm.

Quick Tip

1. Find radius (r) from circumference (C): $C = 2\pi r \implies 100 = 2\pi r \implies r = \frac{50}{\pi}$. 2. Diameter of circle (D): $D = 2r = 2 \times \frac{50}{\pi} = \frac{100}{\pi}$. 3. For an inscribed square, diagonal of square = diameter of circle. So, diagonal of square (d_{sq}) = $\frac{100}{\pi}$. 4. For a square with side s, $d_{sq} = s\sqrt{2}$. So, $s\sqrt{2} = \frac{100}{\pi}$. $s = \frac{100}{\pi\sqrt{2}}$. 5. Rationalize: $s = \frac{100\sqrt{2}}{2\pi} = \frac{50\sqrt{2}}{\pi}$.

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41. The exponent of 3 in the prime factorisation of 864 is :

(1) 2

(2) 5

(3) 4

(4) 3

Correct Answer: (4) 3

Solution: Concept: Prime factorization is the process of finding which prime numbers multiply together to make the original number. The exponent of a prime factor is the number of times that prime factor appears in the factorization.
Step 1: Perform prime factorization of 864 Start by dividing 864 by the smallest prime numbers.

$$864 \div 2 = 432$$

$$432 \div 2 = 216$$

$$216 \div 2 = 108$$

$$108 \div 2 = 54$$

$$54 \div 2 = 27$$

Now 27 is not divisible by 2. Try the next prime number, 3.

$$27 \div 3 = 9$$

$$9 \div 3 = 3$$

$$3 \div 3 = 1$$

So, the prime factorization of 864 is $2 \times 2 \times 2 \times 2 \times 2 \times 3 \times 3 \times 3$.
Step 2: Write the prime factorization in exponential form

$$864 = 2^5 \times 3^3$$

Step 3: Identify the exponent of 3 In the prime factorization $2^5 \times 3^3$, the prime factor 3 has an exponent of 3.
Therefore, the exponent of 3 in the prime factorization of 864 is 3.

Quick Tip

To find the exponent of a prime in a number's prime factorization: 1. Divide the number repeatedly by that prime until it's no longer divisible. Alternatively, find the full prime factorization. For 864: $864 = 2 \times 432 = 2 \times 2 \times 216 = 2 \times 2 \times 2 \times 108 = 2 \times 2 \times 2 \times 2 \times 54$ $= 2 \times 2 \times 2 \times 2 \times 2 \times 27 = 2^5 \times 27$ Now factorize 27: $27 = 3 \times 9 = 3 \times 3 \times 3 = 3^3$. So, $864 = 2^5 \times 3^3$. The exponent of 3 is 3.

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42. The least numbers divisible by 2,3,7 and 9 is :

(1) 126

(2) 256

(3) 251

(4) 189

Correct Answer: (1) 126

Solution: Concept: The least number divisible by a set of given numbers is their Least Common Multiple (LCM).

Step 1: List the numbers The given numbers are 2, 3, 7, and 9.

Step 2: Find the prime factorization of each number

- $2 = 2^1$
- $3 = 3^1$
- $7 = 7^1$
- $9 = 3^2$

Step 3: Calculate the LCM To find the LCM, take the highest power of each prime factor present in any of the numbers. The prime factors involved are 2, 3, and 7.

- Highest power of 2: $2^1 = 2$
- Highest power of 3: $3^2 = 9$ (from the number 9, as $3^2 > 3^1$)
- Highest power of 7: $7^1 = 7$

LCM = $2^1 \times 3^2 \times 7^1$ LCM = $2 \times 9 \times 7$ LCM = 18×7

Calculate 18×7: $18 \times 7 = (10 + 8) \times 7 = 70 + 56 = 126$. So, the LCM is 126.

The least number divisible by 2, 3, 7, and 9 is 126.

Note: Since 9 is a multiple of 3, any number divisible by 9 is automatically divisible by 3. So, we effectively need the LCM of 2, 7, and 9. LCM(2, 7, 9) = LCM(2, 7, 3^2). Since 2, 7, and 9 are pairwise coprime (no common factors other than 1), their LCM is their product: $2 \times 7 \times 9 = 14 \times 9 = 126$.

Quick Tip

The "least number divisible by" a set of numbers is their LCM (Least Common Multiple). Numbers: 2, 3, 7, 9. Notice that if a number is divisible by 9, it's automatically divisible by 3. So, we need the LCM of 2, 7, and 9. Prime factors: $2 = 2$ $7 = 7$ $9 = 3^2$ LCM = Product of the highest powers of all prime factors involved = $2^1 \times 3^2 \times 7^1 = 2 \times 9 \times 7 = 126$.

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43. Graphically, the two systems of equations $x + 7 = 0, y - 2 = 0$ and $x - 2 = 0, y + 7 = 0$ enclose a :

(1) Square region

(2) Rectangular region

(3) A triangular region

(4) Trapezium shaped region

Correct Answer: (1) Square region

Solution: Concept: Each equation represents a line. We need to identify these lines and the shape they form.

Step 1: Identify the four lines

- From first system: $x + 7 = 0 \implies x = -7$ (Vertical line) $y - 2 = 0 \implies y = 2$ (Horizontal line)
- From second system: $x - 2 = 0 \implies x = 2$ (Vertical line) $y + 7 = 0 \implies y = -7$ (Horizontal line)

Step 2: Visualize or sketch the lines These four lines are $x = -7, x = 2, y = 2, y = -7$. Vertical lines pass through $x = -7$ and $x = 2$. Horizontal lines pass through $y = 2$ and $y = -7$.

Step 3: Determine the properties of the enclosed shape The lines form a quadrilateral.

- The horizontal sides are bounded by $x = -7$ and $x = 2$. The length of these sides is $|2 - (-7)| = |2 + 7| = 9$ units.
- The vertical sides are bounded by $y = 2$ and $y = -7$. The length of these sides is $|2 - (-7)| = |2 + 7| = 9$ units.

Since horizontal lines are perpendicular to vertical lines, all angles of the enclosed quadrilateral are $90°$. The lengths of all four sides are equal (9 units). A quadrilateral with four right angles and four equal sides is a **square**.

Step 4: Conclusion The lines enclose a square region. "Square region" is the most specific correct description. (A square is also a rectangle, but "square" is more precise). The circled option (3) "A triangular region" in the original image is incorrect as four lines typically form a quadrilateral.

Quick Tip

1. Identify the lines: $x = -7$, $y = 2$, $x = 2$, $y = -7$. 2. These are two vertical and two horizontal lines. They will form a rectangle (or a square). 3. Width of the rectangle (distance between $x = -7$ and $x = 2$) = $|2 - (-7)| = 9$. 4. Height of the rectangle (distance between $y = -7$ and $y = 2$) = $|2 - (-7)| = 9$. 5. Since width = height = 9, the figure is a square.

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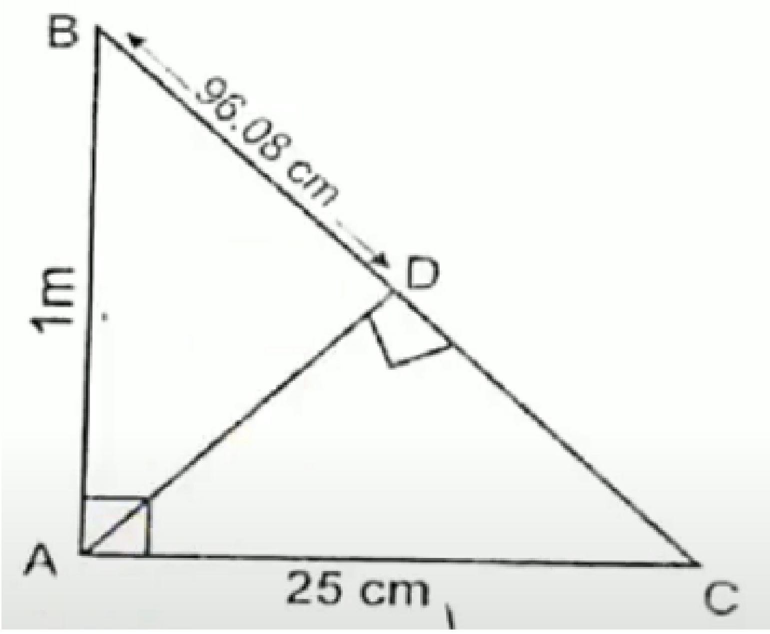
44. In the given figure $\angle CAB = 90°$ and AD $\perp$ BC. If AC = 25 cm, AB = 1m and BD = 96.08 cm, then find the value of AD



Visual shows triangle CAB with right angle at A. AD is perpendicular to BC, with D on BC. Labels: AC=25cm, AB=1m. BD=96.08cm is marked on the hypotenuse segment.

(1) 23cm

(2) 98cm

(3) 24.02cm

(4) none of these

Correct Answer: (3) 24.02cm (as the closest calculated value, acknowledging data inconsistency)

Solution: Concept: In a right-angled triangle, the altitude to the hypotenuse creates similar triangles and specific geometric relationships. The area can also be used. Given: $\triangle CAB$ is right-angled at A. AD $\perp$ BC. AB = 1 m = 100 cm. AC = 25 cm. (The value BD = 96.08 cm appears inconsistent with AB and AC if standard right triangle properties are strictly applied, so we will primarily derive AD from AB, AC, and the resulting BC.)

Step 1: Calculate the hypotenuse BC In right $\triangle CAB$, by Pythagorean theorem: $BC^2 = AB^2 + AC^2$. $BC^2 = (100)^2 + (25)^2 = 10000 + 625 = 10625$. $BC = \sqrt{10625} = \sqrt{25 \times 425} = \sqrt{25 \times 25 \times 17} = 25\sqrt{17}$ cm. Numerically, $BC \approx 103.0776$ cm.

Step 2: Calculate altitude AD using area of $\triangle CAB$ Area of $\triangle CAB$ can be expressed in two ways: 1. Area $= \frac{1}{2} \times AB \times AC$ (using legs) 2. Area $= \frac{1}{2} \times BC \times AD$ (using hypotenuse and altitude AD) Equating them: $AB \times AC = BC \times AD$.

$$AD = \frac{AB \times AC}{BC}$$

Substitute values: $AD = \frac{100 \text{ cm} \times 25 \text{ cm}}{25\sqrt{17} \text{ cm}} = \frac{100}{\sqrt{17}}$ cm.

Step 3: Numerical value of AD $AD = \frac{100}{\sqrt{17}} \approx \frac{100}{4.1231} \approx 24.2535$ cm.

Step 4: Conclusion The calculated value $AD \approx 24.2535$ cm. This is closest to option (3) 24.02cm. The discrepancy likely arises from the given value of BD being inconsistent with AB and AC, or use of rounded values in the options. Based on the primary lengths AB and AC, 24.02cm is the most plausible intended answer among the choices.

Quick Tip

For a right triangle $\triangle CAB$ ($\angle A = 90°$) with altitude AD to hypotenuse BC: 1. Find hypotenuse $BC = \sqrt{AB^2 + AC^2}$. $BC = \sqrt{100^2 + 25^2} = \sqrt{10625} = 25\sqrt{17}$. 2. Altitude $AD = \frac{AB \times AC}{BC}$. (This comes from equating area calculations: $\frac{1}{2}AB \cdot AC = \frac{1}{2}BC \cdot AD$). 3. $AD = \frac{100 \times 25}{25\sqrt{17}} = \frac{100}{\sqrt{17}} \approx 24.25$ cm. 4. Choose the closest option, acknowledging potential inconsistencies in provided problem data (like the given BD value).

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45. If α and β are the roots of $2x^2 - 4x + 1 = 0$. Then $\frac{1}{\alpha^2\beta} + \frac{1}{\alpha\beta^2} =$

(1) -4

(2) 8

(3) 1

(4) 0

Correct Answer: (2) 8

Solution: Concept: For a quadratic equation $ax^2 + bx + c = 0$, if α and β are its roots, then:

- Sum of roots: $\alpha + \beta = -b/a$
- Product of roots: $\alpha\beta = c/a$

We need to simplify the given expression in terms of $\alpha + \beta$ and $\alpha\beta$.

Step 1: Find the sum and product of the roots for the given equation The equation is $2x^2 - 4x + 1 = 0$. Here, $a = 2$, $b = -4$, $c = 1$. Sum of roots: $\alpha + \beta = -(-4)/2 = 4/2 = 2$. Product of roots: $\alpha\beta = 1/2$.

Step 2: Simplify the expression to be evaluated The expression is $\frac{1}{\alpha^2\beta} + \frac{1}{\alpha\beta^2}$. To add these fractions, find a common denominator. The common denominator is $\alpha^2\beta^2$.

$$\frac{1}{\alpha^2\beta} + \frac{1}{\alpha\beta^2} = \frac{\beta}{\alpha^2\beta^2} + \frac{\alpha}{\alpha^2\beta^2}$$
$$= \frac{\beta + \alpha}{\alpha^2\beta^2}$$

This can be written as:

$$= \frac{\alpha + \beta}{(\alpha\beta)^2}$$

Step 3: Substitute the values of $\alpha + \beta$ and $\alpha\beta$ We found: $\alpha + \beta = 2$ $\alpha\beta = 1/2$ Substitute these into the simplified expression:

$$\frac{2}{\left(\frac{1}{2}\right)^2}$$

First, calculate $(\frac{1}{2})^2$:

$$\left(\frac{1}{2}\right)^2 = \frac{1^2}{2^2} = \frac{1}{4}$$

Now the expression becomes:

$$\frac{2}{\frac{1}{4}}$$

To divide by a fraction, multiply by its reciprocal:

$$2 \times \frac{4}{1} = 2 \times 4 = 8$$

The value of the expression is 8.

Quick Tip

1. For $2x^2 - 4x + 1 = 0$: Sum of roots $\alpha + \beta = -(-4)/2 = 2$. Product of roots $\alpha\beta = 1/2$. 2. Simplify the target expression: $\frac{1}{\alpha^2\beta} + \frac{1}{\alpha\beta^2}$ Common denominator is $\alpha^2\beta^2 = (\alpha\beta)^2$. So, $\frac{\beta}{\alpha^2\beta^2} + \frac{\alpha}{\alpha^2\beta^2} = \frac{\alpha+\beta}{(\alpha\beta)^2}$. 3. Substitute values from step 1: $\frac{2}{(1/2)^2} = \frac{2}{1/4} = 2 \times 4 = 8$.

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46. From the top of a tower h meter high, the angle of depression of two objects, which lie on either side of it are α and β. The distance between the two objects is :

(1) $h(\cot\alpha + \cot\beta)$

(2) $h(\cot\alpha - \cot\beta)$

(3) $h(\tan\alpha + \tan\beta)$

(4) $h(\tan\alpha - \tan\beta)$

Correct Answer: (1) $h(\cot\alpha + \cot\beta)$

Solution: Concept: This problem involves angles of depression and basic trigonometry in right-angled triangles. The angle of depression from an observer to an object is the angle between the horizontal line from the observer and the line of sight to the object, when the object is below the horizontal line.

Step 1: Draw a diagram Let T be the top of the tower and F be its foot. The height of the tower TF = h. Let the two objects be O1 and O2, on either side of the foot of the tower, in line with the foot. Let the horizontal line from T be TX. Angle of depression of O1 is $\angle XTO_1 = \alpha$. Angle of depression of O2 is $\angle XTO_2 = \beta$.

Since TX is parallel to the ground FO_1O_2: $\angle TO_1F = \angle XTO_1 = \alpha$ (alternate interior angles). $\angle TO_2F = \angle XTO_2 = \beta$ (alternate interior angles).

We have two right-angled triangles: $\triangle TFO_1$ (right-angled at F) and $\triangle TFO_2$ (right-angled at F). The distance between the two objects is $O_1O_2 = FO_1 + FO_2$.

Step 2: Calculate FO_1 using $\triangle TFO_1$ In right-angled $\triangle TFO_1$: Angle at O_1 is α. Side opposite to α (height) is TF = h. Side adjacent to α (base) is FO_1. We can use $\tan\alpha = \frac{\text{Opposite}}{\text{Adjacent}} = \frac{TF}{FO_1} = \frac{h}{FO_1}$. So, $FO_1 = \frac{h}{\tan\alpha} = h\cot\alpha$.

Step 3: Calculate FO_2 using $\triangle TFO_2$ In right-angled $\triangle TFO_2$: Angle at O_2 is β. Side opposite to β (height) is TF = h. Side adjacent to β (base) is FO_2. We can use $\tan\beta = \frac{\text{Opposite}}{\text{Adjacent}} = \frac{TF}{FO_2} = \frac{h}{FO_2}$. So, $FO_2 = \frac{h}{\tan\beta} = h\cot\beta$.

Step 4: Calculate the distance between the objects O_1O_2 Distance $O_1O_2 = FO_1 + FO_2$. Substitute the expressions for FO_1 and FO_2: $O_1O_2 = h\cot\alpha + h\cot\beta$. Factor out h: $O_1O_2 = h(\cot\alpha + \cot\beta)$.

This matches option (1).

Quick Tip

1. Draw the tower (height h) and the two objects O1 and O2 on opposite sides at ground level.
2. The angles of depression (α, β) from the top of the tower to O1 and O2 are equal to the angles of elevation from O1 and O2 to the top of the tower, respectively.
3. Let d_1 be the distance from the foot of the tower to O1, and d_2 be the distance to O2.
4. In the right triangle for O1: $\tan\alpha = h/d_1 \implies d_1 = h/\tan\alpha = h\cot\alpha$.
5. In the right triangle for O2: $\tan\beta = h/d_2 \implies d_2 = h/\tan\beta = h\cot\beta$.
6. Total distance between objects = $d_1 + d_2 = h\cot\alpha + h\cot\beta = h(\cot\alpha + \cot\beta)$.

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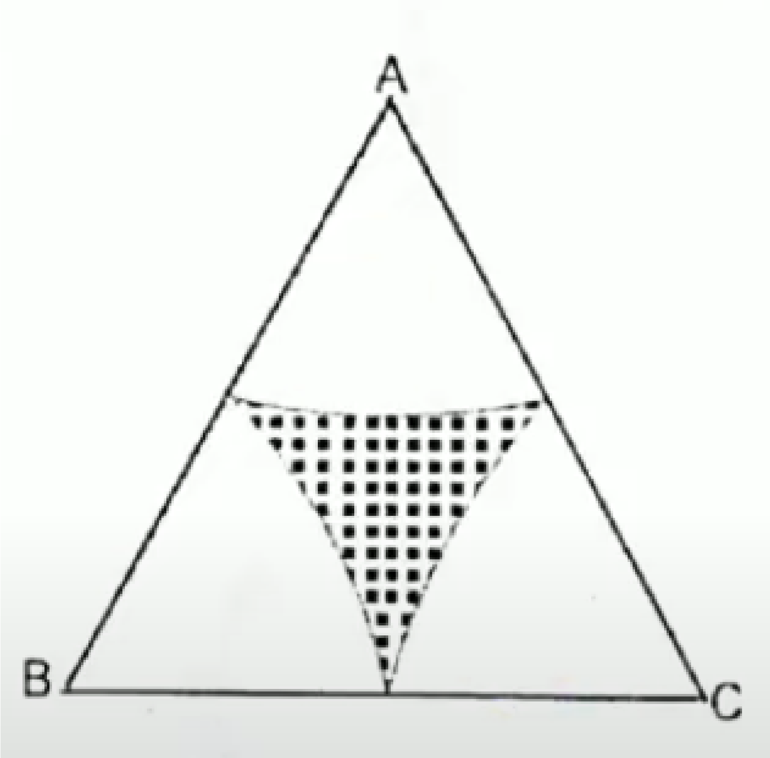
47. In the given figure $\triangle$ABC is an equilateral triangle of side 8 cm. A,B and C are the centres of circular arcs of radius 4 cm. Find the area of the shaded region correct upto 2 decimal places ($\pi = 3.142, \sqrt{3} = 1.732$)

Equilateral triangle ABC. From each vertex (A, B, C), a circular sector is drawn inside the triangle.



Radius of each sector is 4 cm. Side of triangle is 8 cm. The shaded region is the area of the triangle MINUS the areas of these three sectors.

(1) 2.57 cm^2

(2) 3.45 cm^2

(3) 1.67 cm^2

(4) none of these

Correct Answer: (1) 2.57 cm^2 (Based on calculation being closest to this option)

Solution: Concept: The shaded area = Area of equilateral triangle - Sum of areas of three circular sectors.

Step 1: Area of the equilateral triangle ($A_{triangle}$**)** Side $s = 8$ cm. $\sqrt{3} = 1.732$.
$A_{triangle} = \frac{\sqrt{3}}{4}s^2 = \frac{1.732}{4}(8)^2 = \frac{1.732}{4}(64) = 1.732 \times 16 = 27.712$ cm^2.

Step 2: Area of the three circular sectors ($A_{sectors}$**)** Each angle of an equilateral triangle is $60°$. This is the angle for each sector. Radius of each sector $r_{sector} = 4$ cm. $\pi = 3.142$. Area of one sector = $\frac{\theta}{360°}\pi r_{sector}^2 = \frac{60°}{360°} \times 3.142 \times (4)^2 = \frac{1}{6} \times 3.142 \times 16$. Area of three identical sectors = $3 \times \left(\frac{1}{6} \times 3.142 \times 16\right) = \frac{1}{2} \times 3.142 \times 16$. $A_{sectors} = 8 \times 3.142 = 25.136$ cm^2.
(Alternatively, the three $60°$ sectors form a semicircle of radius 4 cm).

Step 3: Area of the shaded region Area (shaded) = $A_{triangle} - A_{sectors}$ Area (shaded) = $27.712\text{ cm}^2 - 25.136\text{ cm}^2 = 2.576\text{ cm}^2$.

Step 4: Round to 2 decimal places Rounding 2.576 cm^2 to two decimal places gives 2.58 cm^2. Option (1) is 2.57 cm^2, which is the closest. The slight difference is likely due to rounding in the provided values of π or $\sqrt{3}$, or the options themselves.

Quick Tip

1. Area of equilateral triangle (side 8cm): $\frac{\sqrt{3}}{4} \times 8^2 = 16\sqrt{3} \approx 16 \times 1.732 = 27.712\text{ cm}^2$. 2. The 3 sectors (each $60°$, radius 4cm) together form a semicircle ($180°$) of radius 4cm. Area of 3 sectors = $\frac{1}{2}\pi r^2 = \frac{1}{2} \times 3.142 \times 4^2 = 8 \times 3.142 = 25.136\text{ cm}^2$. 3. Shaded Area = Area of Triangle - Area of 3 Sectors $= 27.712 - 25.136 = 2.576\text{ cm}^2$. 4. Closest option when rounded: 2.57 cm^2 or 2.58 cm^2. Option (1) is chosen.

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48. In a single throw of pair of dice, the probability of getting a multiple of 2 on one and a multiple of 3 on the other will be :

(1) $\frac{5}{36}$

(2) $\frac{1}{18}$

(3) $\frac{11}{36}$

(4) $\frac{1}{6}$

Correct Answer: (3) $\frac{11}{36}$

Solution: Concept: Probability = (Favorable Outcomes) / (Total Outcomes). Total outcomes for two dice = $6 \times 6 = 36$.

Step 1: Define events Let D1 be the outcome of the first die, D2 for the second. M2 = Multiple of 2 = $\{2, 4, 6\}$ (3 outcomes) M3 = Multiple of 3 = $\{3, 6\}$ (2 outcomes) We want (D1 $\in$ M2 AND D2 $\in$ M3) OR (D1 $\in$ M3 AND D2 $\in$ M2).

Step 2: List favorable pairs Case A: (D1 is M2, D2 is M3) Pairs: (2,3), (2,6), (4,3), (4,6), (6,3), (6,6). (Count = $3 \times 2 = 6$ pairs)

Case B: (D1 is M3, D2 is M2) Pairs: (3,2), (3,4), (3,6), (6,2), (6,4), (6,6). (Count = $2 \times 3 = 6$ pairs)

Step 3: Combine cases and remove overlap The outcome (6,6) is in both Case A and Case B (6 is a multiple of 2 and 6 is a multiple of 3). Total unique favorable outcomes = (Outcomes in A) + (Outcomes in B) - (Overlapping outcomes) Total = $6 + 6 - 1$ (since (6,6) is the only overlap). Total favorable outcomes = 11.

Step 4: Calculate probability Probability = $\frac{\text{Number of favorable outcomes}}{\text{Total number of outcomes}} = \frac{11}{36}$.

> **Quick Tip**
>
> Total outcomes with two dice = 36. Multiples of 2 (M2): 2, 4, 6. Multiples of 3 (M3): 3, 6. We need (M2 on 1st die AND M3 on 2nd die) OR (M3 on 1st die AND M2 on 2nd die). Pairs for (M2, M3): (2,3), (2,6), (4,3), (4,6), (6,3), (6,6) - 6 pairs. Pairs for (M3, M2): (3,2), (3,4), (3,6), (6,2), (6,4), (6,6) - 6 pairs. The pair (6,6) is common. So, unique favorable pairs = $6 + 6 - 1 = 11$. Probability = $11/36$.

---

49. Two dice are numbered 1,2,3,4,5,6 and 1,1,2,2,3,3 respectively. They are thrown and the sum of the numbers on them is noted. The probability of getting a sum of 5 will be :

(1) $\frac{1}{9}$

(2) $\frac{1}{18}$

(3) $\frac{1}{6}$

(4) $\frac{1}{36}$

Correct Answer: (3) $\frac{1}{6}$

Solution: Concept: Probability = (Number of favorable outcomes) / (Total number of possible outcomes). The dice are not standard identical dice; their numberings are different.

Step 1: List all possible outcomes and their total number Die 1 (D1) has faces: $\{1, 2, 3, 4, 5, 6\}$ Die 2 (D2) has faces: $\{1, 1, 2, 2, 3, 3\}$ When two dice are thrown, the total number of possible outcomes is $6 \times 6 = 36$. Each outcome is equally likely if we consider each face of each die distinct (e.g., D2 has two '1' faces, say 1a and 1b).

Step 2: Identify favorable outcomes for getting a sum of 5 We need to find pairs (Outcome on D1, Outcome on D2) such that their sum is 5. Let's list them:

- If D1 = 1, D2 needs to be 4. D2 does not have 4. (No outcome)

- If D1 = 2, D2 needs to be 3. D2 has two '3' faces. So, (2, 3) and (2, 3). (2 outcomes)
- If D1 = 3, D2 needs to be 2. D2 has two '2' faces. So, (3, 2) and (3, 2). (2 outcomes)
- If D1 = 4, D2 needs to be 1. D2 has two '1' faces. So, (4, 1) and (4, 1). (2 outcomes)
- If D1 = 5, D2 needs to be 0. D2 does not have 0. (No outcome)
- If D1 = 6, D2 needs to be -1. D2 does not have -1. (No outcome)

The favorable pairs (D1, D2 value) are: (2, 3) - happens in 2 ways because D2 has two '3's. (3, 2) - happens in 2 ways because D2 has two '2's. (4, 1) - happens in 2 ways because D2 has two '1's.
Total number of favorable outcomes = $2 + 2 + 2 = 6$.
Step 3: Calculate the probability Total number of possible outcomes = 36. Number of favorable outcomes = 6. Probability (sum = 5) = $\frac{\text{Number of favorable outcomes}}{\text{Total number of possible outcomes}}$

$$P(\text{sum} = 5) = \frac{6}{36} = \frac{1}{6}$$

Quick Tip

Die 1 (D1): $\{1, 2, 3, 4, 5, 6\}$ Die 2 (D2): $\{1a, 1b, 2a, 2b, 3a, 3b\}$ (labeling the repeated faces for clarity) Total outcomes = $6 \times 6 = 36$. Pairs (D1, D2) that sum to 5:

- D1=2, D2=3: (2, 3a), (2, 3b) - 2 outcomes
- D1=3, D2=2: (3, 2a), (3, 2b) - 2 outcomes
- D1=4, D2=1: (4, 1a), (4, 1b) - 2 outcomes

Total favorable outcomes = $2 + 2 + 2 = 6$. Probability = $6/36 = 1/6$.

---

50. A bag contains 14 balls of which x are white. If 6 more white balls are added to the bag, the probability of drawing a white ball is $\frac{1}{2}$. Then the value of $x =$
(1) 7
(2) 4
(3) 8

(4) none of these

Correct Answer: (2) 4

Solution: Concept: Probability = (Number of favorable outcomes) / (Total number of possible outcomes).

Step 1: Describe the initial state of the bag Initially:

- Total number of balls = 14.
- Number of white balls = x.
- Number of non-white balls = $14 - x$.

Step 2: Describe the state of the bag after adding more white balls 6 more white balls are added to the bag. After adding:

- New number of white balls = $x + 6$.
- New total number of balls = Initial total + Added balls = $14 + 6 = 20$.

Step 3: Set up the probability equation for drawing a white ball after adding more The probability of drawing a white ball from the modified bag is given as $\frac{1}{2}$. Using the probability formula: P(drawing a white ball) = $\frac{\text{New number of white balls}}{\text{New total number of balls}}$

$$\frac{x+6}{20} = \frac{1}{2}$$

Step 4: Solve the equation for x

$$\frac{x+6}{20} = \frac{1}{2}$$

To solve for x, we can cross-multiply or multiply both sides by 20. Multiplying both sides by 20:

$$x + 6 = \frac{1}{2} \times 20$$

$$x + 6 = 10$$

Subtract 6 from both sides:

$$x = 10 - 6$$

$$x = 4$$

Step 5: Check the answer (optional) If $x = 4$: Initially: 4 white balls, total 14 balls. After adding 6 white balls: Number of white balls = $4 + 6 = 10$. Total number of balls = $14 + 6 = 20$. Probability of drawing a white ball = $\frac{10}{20} = \frac{1}{2}$. This matches the given information.

Therefore, the value of x is 4.

Quick Tip

1. Initial state: Total = 14, White = x. 2. After adding 6 white balls: New White = $x + 6$. New Total = $14 + 6 = 20$. 3. Probability of drawing a white ball now = (New White) / (New Total). Given this probability is $1/2$. 4. Equation: $\frac{x+6}{20} = \frac{1}{2}$. 5. Solve for x: $2(x + 6) = 20 \times 1$ $2x + 12 = 20$ $2x = 20 - 12 = 8$ $x = 4$.

Physics

51. The minimum distance between source and reflecting surface for echo is :

(1) 10.2 m

(2) 17.2 m

(3) 20.4 m

(4) 27.4 m

Correct Answer: (2) 17.2 m

Solution: Concept: An echo is the repetition of sound caused by the reflection of sound waves from a surface. For an echo to be heard distinctly, there must be a minimum time interval between the original sound and the reflected sound. This minimum time interval is known as the persistence of hearing.

Step 1: Persistence of Hearing The human ear can distinguish between two sounds if the time interval between them is at least about 0.1 seconds. This is called the persistence of hearing. So, for an echo to be heard distinctly, the reflected sound must reach the ear at least 0.1 seconds after the original sound is produced. Let $t = 0.1$ s.

Step 2: Speed of Sound The speed of sound in air varies with temperature. At a room temperature of around $20°$C to $22°$C, the speed of sound is approximately 344 m/s. We will use this value for calculation, as it directly leads to one of the options. Let $v = 344$ m/s.

Step 3: Calculating the Minimum Distance Let d be the minimum distance between the source of sound and the reflecting surface. For an echo, the sound travels from the source to the reflecting surface (distance d) and then travels back from the reflecting surface to the source/listener (another distance d). So, the total distance travelled by the sound for the echo to be heard is $2d$.

We know the basic relationship: distance = speed × time. Applying this to our echo scenario:

$$\text{Total distance} = \text{Speed of sound} \times \text{Time interval}$$

$$2d = v \times t$$

To find the minimum distance d to the reflector, we can rearrange the formula:

$$d = \frac{v \times t}{2}$$

Step 4: Substitution and Calculation Substitute the values of $v = 344$ m/s and $t = 0.1$ s into the formula:

$$d = \frac{(344 \text{ m/s}) \times (0.1 \text{ s})}{2}$$

$$d = \frac{34.4 \text{ m}}{2}$$

$$d = 17.2 \text{ m}$$

Thus, the minimum distance between the source and the reflecting surface for an echo to be heard distinctly is 17.2 m.

Quick Tip

Key factors for echo calculation: - **Persistence of hearing:** For humans, this is about 0.1 s. This is the minimum time gap needed to distinguish the original sound from its echo.
- **Speed of sound (v):** This depends on the medium and its temperature. For air at $\approx 22°$C, $v \approx 344$ m/s.
- **Formula:** The total distance sound travels is $2d$. So, $2d = v \times t$, which gives $d = \frac{v \times t}{2}$. Always check if the problem provides a specific speed of sound; otherwise, use a standard room temperature value.

---

52. If time period is 0.02 second, then frequency will be :

(1) 50 Hz

(2) 5 Hz

(3) 0.02 Hz

(4) 500 Hz

Correct Answer: (1) 50 Hz

Solution: Concept: Time period and frequency are fundamental properties of periodic phenomena like waves or oscillations. They are inversely related.

Step 1: Understanding Time Period (T) The time period (T) is the time taken for one complete cycle of a wave or one full oscillation. Given: Time period $T = 0.02$ seconds.

Step 2: Understanding Frequency (f) Frequency (f) is the number of complete cycles or oscillations that occur in one second. The unit of frequency is Hertz (Hz), where
1 Hz $= 1$ cycle per second.

Step 3: The Relationship between Frequency and Time Period Frequency and time period are reciprocals of each other. The formula connecting them is:

$$f = \frac{1}{T}$$

This means that if you know the time period, you can calculate the frequency, and vice-versa ($T = \frac{1}{f}$).

Step 4: Calculation We are given $T = 0.02$ s. Substitute this value into the frequency formula:

$$f = \frac{1}{0.02 \text{ s}}$$

To simplify the division by a decimal, you can express 0.02 as a fraction: $0.02 = \frac{2}{100}$. So the equation becomes:

$$f = \frac{1}{\frac{2}{100}}$$

When dividing by a fraction, you multiply by its reciprocal:

$$f = 1 \times \frac{100}{2}$$

$$f = \frac{100}{2}$$

$$f = 50 \text{ Hz}$$

Therefore, if the time period is 0.02 seconds, the frequency will be 50 Hz.

Quick Tip

The relationship $f = \frac{1}{T}$ is fundamental for all wave phenomena. - A **short** time period means many cycles happen quickly, so the **frequency is high**.

- A **long** time period means cycles happen slowly, so the **frequency is low**.

Units are important: T in seconds (s), f in Hertz (Hz).

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53. In SONAR, we use :

(1) Audible Sound

(2) Radio Sound

(3) Ultra Sound

(4) Infra Sound

Correct Answer: (3) Ultra Sound

Solution: Concept: SONAR is a technology used primarily for underwater detection and navigation. It relies on sound waves.

Step 1: Understanding SONAR SONAR is an acronym for **SO**und **NA**vigation and **R**anging. It works by emitting sound pulses and then detecting the echoes that return after these pulses reflect off objects. By measuring the time delay between the emission of the pulse and the reception of the echo, the distance to the object can be determined.

Step 2: Classifying Sound Waves by Frequency Sound waves are categorized based on their frequency:

- **Infra Sound (Infrasonic waves):** Frequencies below the range of human hearing, typically less than 20 Hz.
- **Audible Sound:** Frequencies within the range of human hearing, typically from 20 Hz to $20,000$ Hz (or 20 kHz).
- **Ultra Sound (Ultrasonic waves):** Frequencies above the range of human hearing, typically greater than $20,000$ Hz (or 20 kHz).

Note: "Radio Sound" is not a standard classification for sound waves used in this context. Radio waves are electromagnetic waves, not sound waves, and are used in RADAR, not typically SONAR.

Step 3: Why Ultra Sound is Used in SONAR Ultrasonic waves are chosen for SONAR applications due to several advantageous properties:

- **High Frequency and Short Wavelength:** Ultrasonic waves have high frequencies, which mean they have short wavelengths (since wavelength $\lambda = v/f$, where v is the speed of sound and f is frequency). Short wavelengths are less prone to diffraction (bending around obstacles) and can be focused into narrow beams. This allows for better resolution in detecting smaller objects and determining their direction more accurately.
- **Good Directionality:** Because they can be directed in a narrow beam, they are not easily scattered in all directions. This is crucial for pinpointing the location of an object.
- **High Energy:** Ultrasonic waves can be produced with high energy, allowing them to travel longer distances in water without significant loss of intensity.

Audible sound waves have longer wavelengths and would spread out more, making it harder to detect small objects or determine their precise location. Infrasound has very long wavelengths and is generally not suitable for the detailed ranging and imaging tasks SONAR is used for.

Therefore, Ultra Sound is the type of sound used in SONAR systems.

Quick Tip

Remember the sound spectrum: - **Infrasound** (< 20 Hz): Long wavelength, travels far, used by large animals (elephants, whales) for communication, associated with earthquakes.
- **Audible Sound** (20 Hz – 20 kHz): What humans hear.
- **Ultrasound** (> 20 kHz): Short wavelength, high directionality. Used in SONAR, medical imaging (sonograms), non-destructive testing, and by animals like bats and dolphins for echolocation.
The choice of ultrasound for SONAR is driven by its ability to provide detailed information about underwater environments and objects.

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54. The velocity of an object becomes double then its kinetic energy will be :

(1) Kinetic Energy does not depend on velocity

(2) Two times

(3) Four times

(4) Eight times

Correct Answer: (3) Four times

Solution: Concept: Kinetic energy (KE) is the energy an object possesses due to its motion. It depends on the object's mass and its velocity.

Step 1: Recall the Formula for Kinetic Energy The kinetic energy of an object is given by the formula:

$$\text{KE} = \frac{1}{2}mv^2$$

where:

- m is the mass of the object (assumed to be constant in this problem).
- v is the velocity (or speed, since KE is a scalar) of the object.

Step 2: Define Initial and Final States Let the initial velocity of the object be v_1. The initial kinetic energy (KE_1) is:

$$\text{KE}_1 = \frac{1}{2}mv_1^2$$

The problem states that the velocity of the object becomes double. So, the new (final) velocity, let's call it v_2, is:

$$v_2 = 2v_1$$

Step 3: Calculate the New Kinetic Energy Now, we calculate the new kinetic energy (KE_2) using the new velocity v_2:

$$\text{KE}_2 = \frac{1}{2}mv_2^2$$

Substitute the expression for v_2 ($2v_1$) into this equation:

$$\text{KE}_2 = \frac{1}{2}m(2v_1)^2$$

Step 4: Simplify the Expression for New Kinetic Energy When we square $(2v_1)$, both the 2 and v_1 are squared:

$$(2v_1)^2 = 2^2 \times v_1^2 = 4v_1^2$$

Now substitute this back into the equation for KE_2:

$$\text{KE}_2 = \frac{1}{2}m(4v_1^2)$$

We can rearrange the terms to make the comparison clearer:

$$\text{KE}_2 = 4 \times \left(\frac{1}{2}mv_1^2\right)$$

Step 5: Compare the New Kinetic Energy with the Initial Kinetic Energy Notice that the term in the parentheses, $\left(\frac{1}{2}mv_1^2\right)$, is exactly the expression for the initial kinetic energy, KE_1. So, we can write:

$$\text{KE}_2 = 4 \times \text{KE}_1$$

This shows that the new kinetic energy is four times the initial kinetic energy.

Therefore, if the velocity of an object becomes double, its kinetic energy will be four times.

Quick Tip

The relationship KE $= \frac{1}{2}mv^2$ shows that kinetic energy is proportional to the square of the velocity (KE $\propto v^2$). This means: - If velocity v is multiplied by a factor N, the kinetic energy KE is multiplied by a factor N^2.
- If v doubles (factor of 2), KE increases by $2^2 = 4$ times.
- If v triples (factor of 3), KE increases by $3^2 = 9$ times.
This quadratic relationship has significant implications, for example, in vehicle safety.

---

55. Maximum work is done, when the angle between force and displacement is :

(1) $60°$

(2) $45°$

(3) $30°$

(4) $0°$

Correct Answer: (4) $0°$

Solution: Concept: Work done by a constant force is a scalar quantity that measures the energy transferred to or from an object when it is moved by that force. It depends on the magnitude of the force, the magnitude of the displacement, and the angle between the force and displacement vectors.
Step 1: Recall the Formula for Work Done The work done (W) by a constant force (F) that causes a displacement (d) is given by:

$$W = Fd\cos\theta$$

where:

- F is the magnitude of the constant force applied.
- d is the magnitude of the displacement of the object.
- θ is the angle between the direction of the force vector $\vec{F}$ and the direction of the displacement vector $\vec{d}$.

Step 2: Identify the Condition for Maximum Work For the work done (W) to be maximum, given that F and d are constant magnitudes, the value of $\cos\theta$ must be maximum.
Step 3: Determine the Maximum Value of $\cos\theta$ The cosine function, $\cos\theta$, ranges from -1 to +1. The maximum value of $\cos\theta$ is 1.
Step 4: Find the Angle θ for which $\cos\theta$ is Maximum The cosine function $\cos\theta = 1$ when the angle $\theta = 0°$.
Step 5: Calculate Maximum Work and Evaluate Options If $\cos\theta = 1$ (when $\theta = 0°$), the work done is:

$$W_{\text{max}} = Fd(1) = Fd$$

Let's check the $\cos\theta$ values for the given options:

- For $\theta = 60°$: $\cos 60° = 0.5$. So, $W = 0.5Fd$.
- For $\theta = 45°$: $\cos 45° \approx 0.707$. So, $W \approx 0.707Fd$.
- For $\theta = 30°$: $\cos 30° \approx 0.866$. So, $W \approx 0.866Fd$.
- For $\theta = 0°$: $\cos 0° = 1$. So, $W = Fd$.

Comparing these values, Fd (when $\theta = 0°$) is the largest.
Therefore, maximum work is done when the angle between force and displacement is $0°$.

Quick Tip

The angle θ in $W = Fd\cos\theta$ is crucial: - $\theta = 0°$ ($\cos 0° = 1$): Work is maximum and positive ($W = Fd$).
- $\theta = 90°$ ($\cos 90° = 0$): Work is zero ($W = 0$).
- $\theta = 180°$ ($\cos 180° = -1$): Work is maximum negative ($W = -Fd$).
Force is most effective when aligned with displacement.

---

56. An object of mass 2 Kg is lifted up to height 2 m. The work done will be :

(1) 39.20 J

(2) 9.80 J

(3) 98 J

(4) 980 J

Correct Answer: (1) 39.20 J

Solution: Concept: When an object is lifted against gravity, the work done on the object is stored as gravitational potential energy. The work done in lifting an object is equal to the change in its potential energy.

Step 1: Identify the given quantities Mass of the object, $m = 2$ Kg Height to which the object is lifted, $h = 2$ m Acceleration due to gravity, g. The standard value for g is approximately 9.8 m/s^2. We will use this value.

Step 2: Recall the formula for work done against gravity (or change in potential energy) The work done (W) in lifting an object of mass m to a height h against gravity is given by:

$$W = mgh$$

This formula is derived from $W = Fd\cos\theta$. Here, the force F required to lift the object is equal to its weight (mg), the displacement d is the height h, and the angle θ between the upward lifting force and the upward displacement is $0°$ ($\cos 0° = 1$).

Step 3: Substitute the values into the formula

$$W = (2 \text{ Kg}) \times (9.8 \text{ m/s}^2) \times (2 \text{ m})$$

Step 4: Calculate the work done First, multiply the mass and height: 2 Kg $\times$ 2 m $= 4$ Kg $\cdot$ m Now, multiply this by g:

$$W = (4 \text{ Kg} \cdot \text{m}) \times (9.8 \text{ m/s}^2)$$

$$W = 39.2 \text{ Kg} \cdot \text{m}^2/\text{s}^2$$

The unit Kg $\cdot$ m^2/s^2 is equivalent to Joules (J), which is the SI unit of work and energy.

$$W = 39.2 \text{ J}$$

Looking at the options, 39.20 J matches our calculation.

Therefore, the work done will be 39.20 J.

Quick Tip

When calculating work done against gravity: - Use the formula $W = mgh$.
- Remember that g is the acceleration due to gravity, approximately 9.8 m/s^2 on Earth. Sometimes, for simpler calculations, $g = 10$ m/s^2 might be used if specified or if options suggest it, but 9.8 m/s^2 is more precise.
- Ensure all units are in the SI system (mass in Kg, height in m, g in m/s^2) to get the work done in Joules (J).
This work done is stored as gravitational potential energy in the object.

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57. SI unit of gravitational constant (G) is :

(1) Nm^{2}kg^{-1}

(2) Nm kg^{-2}

(3) N^{2}m kg^{-2}

(4) Nm^{2}kg^{-2}

Correct Answer: (4) Nm^{2}kg^{-2}

Solution: Concept: The gravitational constant (G) appears in Newton's Law of Universal Gravitation. We can derive its SI unit by rearranging this law.

Step 1: Recall Newton's Law of Universal Gravitation The formula for the gravitational force (F) between two masses (m_1 and m_2) separated by a distance (r) is:

$$F = G\frac{m_1 m_2}{r^2}$$

where G is the universal gravitational constant.

Step 2: Rearrange the formula to solve for G To find the units of G, we first need to isolate G on one side of the equation: Multiply both sides by r^2:

$$Fr^2 = Gm_1m_2$$

Now, divide both sides by m_1m_2:

$$G = \frac{Fr^2}{m_1m_2}$$

Step 3: Determine the SI units for each quantity in the rearranged formula

- F (Force): The SI unit is Newton (N).
- r (distance): The SI unit is meter (m). So, r^2 will have units of m^2.
- m_1 (mass): The SI unit is kilogram (kg).
- m_2 (mass): The SI unit is kilogram (kg).
- Therefore, $m_1 m_2$ will have units of $\text{kg} \times \text{kg} = \text{kg}^2$.

Step 4: Substitute the units into the expression for G

$$\text{Units of G} = \frac{(\text{Units of } F) \times (\text{Units of } r^2)}{(\text{Units of } m_1 m_2)}$$

$$\text{Units of G} = \frac{\text{N} \times \text{m}^2}{\text{kg}^2}$$

This can also be written as:

$$\text{Units of G} = \text{Nm}^2\text{kg}^{-2}$$

or

$$\text{Units of G} = \text{N} \cdot \text{m}^2/\text{kg}^2$$

Comparing this with the given options, option (4) $\text{Nm}^2\text{kg}^{-2}$ matches our derived unit.

Option (1) $\text{Nm}^2\text{kg}^{-1}$ is incorrect because the denominator should be kg^2.

Therefore, the SI unit of the gravitational constant (G) is $\text{Nm}^2\text{kg}^{-2}$.

Quick Tip

To find the units of a physical constant from a formula:

1. Write down the formula involving the constant.
2. Rearrange the formula to make the constant the subject.
3. Substitute the SI units of all other quantities in the formula.
4. Simplify the resulting expression of units.

This method is widely applicable for deriving units of various physical constants (e.g., Planck's constant, specific heat capacity, etc.).

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58. Mass of a man is 60 Kg, his mass on the moon will be :

(1) 60 Kg

(2) 10 Kg

(3) 98 Kg

(4) 0 Kg

Correct Answer: (1) 60 Kg

Solution: Concept: It is crucial to distinguish between mass and weight.

- **Mass** is a measure of the amount of matter in an object. It is an intrinsic property of the object and remains constant regardless of its location. The SI unit of mass is the kilogram (Kg).
- **Weight** is the force of gravity acting on an object. It depends on both the mass of the object and the acceleration due to gravity at its location ($W = mg$). The SI unit of weight is the Newton (N).

Step 1: Identify the given quantity The mass of the man on Earth is given as 60 Kg.
Step 2: Understand how mass changes with location Mass is an inherent property of an object, representing the quantity of matter it contains. This quantity of matter does not change whether the man is on Earth, on the Moon, or anywhere else in the universe (unless he gains or loses actual matter, which is not the case here).
Step 3: Consider the effect of the Moon's gravity The Moon has a weaker gravitational pull than Earth. The acceleration due to gravity on the Moon is approximately $1/6$th of that on Earth ($g_{\text{moon}} \approx \frac{1}{6}g_{\text{earth}}$). This means the man's *weight* on the Moon would be different (approximately $1/6$th of his weight on Earth). If his weight on Earth is $W_E = 60$ Kg $\times\, g_{\text{earth}}$, his weight on the Moon would be $W_M = 60$ Kg $\times\, g_{\text{moon}} = 60$ Kg $\times \frac{1}{6}g_{\text{earth}} = 10$ Kg $\times\, g_{\text{earth}}$. If we were considering "effective mass" based on weight, then 10Kg might seem plausible, but the question explicitly asks for "mass".
Step 4: Determine the man's mass on the Moon Since mass is a constant property and does not depend on the gravitational field, the man's mass on the Moon will be the same as his mass on Earth. Therefore, his mass on the Moon will still be 60 Kg.
Option (2) 10 Kg would be related to his weight on the moon if g_{earth} was roughly $9.8m/s^2$, then his weight on Earth would be $60 \times 9.8 \approx 588N$. His weight on the moon would be $1/6 \times 588N \approx 98N$. If this $98N$ weight was divided by Earth's g ($9.8m/s^2$), you'd get an

"equivalent Earth mass" of $10Kg$, but this is a misinterpretation of the question. The question asks for mass, not an equivalent mass based on lunar weight compared to Earth's gravity. Option (3) 98 Kg might come from confusing mass with weight in Newtons (e.g., if $g = 9.8$, a 10Kg mass has a weight of 98N), but mass is in Kg.
The correct answer is that his mass remains 60 Kg.

Quick Tip

Key difference: **Mass:** Amount of matter. Constant everywhere. Unit: Kilogram (Kg). **Weight:** Force of gravity on mass ($W = mg$). Varies with location (depends on g). Unit: Newton (N).
Your mass is the same on Earth, the Moon, Jupiter, or in deep space. Your weight, however, changes significantly. Don't confuse the two!

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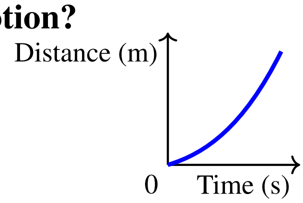
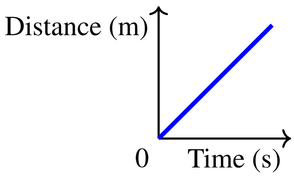
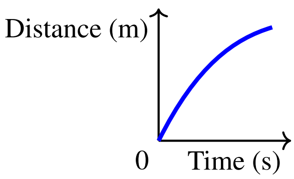
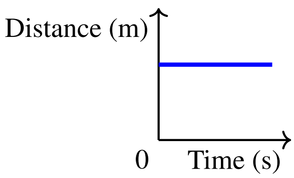
59. Which of the following distance-time graphs represents an object in uniform motion?

(1) 

(2) 

✓

(3) 

(4) 

Correct Answer: (2)

Solution: Concept: Uniform motion means an object travels with a constant velocity. This implies that the object covers equal distances in equal intervals of time. For a distance-time graph:

- The slope of the graph represents the velocity of the object.
- For uniform motion (constant velocity), the slope must be constant.
- A graph with a constant slope is a straight line.

Analyzing the Graphs:

- **Graph (1):** This graph shows a curve where the distance covered in successive time intervals is increasing. The slope of the curve is increasing with time. This represents non-uniform motion, specifically accelerated motion (velocity is increasing).
- **Graph (2):** This graph is a straight line passing through the origin with a constant positive slope. A constant slope means constant velocity. This represents uniform motion where the object starts from the origin (distance=0 at time=0) and moves away at a steady speed.
- **Graph (3):** This graph shows a curve where the distance covered in successive time intervals is decreasing (though the total distance is still increasing). The slope of the curve is decreasing with time. This represents non-uniform motion, specifically decelerated motion (velocity is decreasing but still positive, or the rate of increase of distance is slowing down).
- **Graph (4):** This graph is a horizontal straight line. This means that the distance of the object remains constant as time passes. A constant distance implies that the object is not moving; its velocity is zero. An object at rest is technically in uniform motion (zero acceleration), but typically ”uniform motion” in such questions implies a constant non-zero velocity. If this were a velocity-time graph, a horizontal line would indicate constant velocity.

Conclusion: Graph (2) correctly depicts uniform motion because it is a straight line, indicating a constant rate of change of distance with respect to time (constant velocity). Therefore, the graph representing uniform motion is (2).

Quick Tip

For distance-time graphs: **Slope = Velocity.**

Horizontal line (slope=0): Object is at rest (zero velocity).

Straight line with non-zero slope: Uniform motion (constant velocity). - Positive slope: Moving away from the starting point. - Negative slope: Moving towards the starting point (if distance can be negative or represents position relative to origin).

Curved line: Non-uniform motion (velocity is changing, i.e., there is acceleration). - Slope increasing (curve gets steeper): Acceleration. - Slope decreasing (curve gets flatter): Deceleration.

Always pay attention to the axes labels to correctly interpret the graph.

---

60. A particle is moving on a circular path of radius r. Its displacement after one revolution :

(1) $2\pi r$

(2) πr

(3) 0 (zero)

(4) $2r$

Correct Answer: (3) 0 (zero)

Solution: Concept: It's important to differentiate between distance and displacement.

- **Distance** is the total length of the path traveled by an object. It is a scalar quantity.
- **Displacement** is the shortest straight-line distance between the initial and final positions of an object. It is a vector quantity (meaning it has both magnitude and direction).

Step 1: Visualize the motion A particle is moving on a circular path of radius r. It completes one revolution. This means the particle starts at a certain point on the circle and, after moving along the circumference, returns to the exact same starting point.

Step 2: Determine the initial and final positions Let the starting point of the particle on the circular path be point A. After one complete revolution, the particle comes back to point A. So, Initial Position = Point A Final Position = Point A

Step 3: Calculate the displacement Displacement is defined as the change in position, or the shortest straight-line distance between the initial and final positions. Since the initial position and the final position are the same (Point A), the straight-line distance between them is zero. Displacement = Final Position - Initial Position If both are the same point, the vector difference is the zero vector, and its magnitude is 0.

Therefore, the displacement of the particle after one complete revolution is 0 (zero).

Step 4: Consider the distance traveled (for clarity, though not asked) The distance traveled in one revolution is the circumference of the circular path. Circumference of a circle with radius r is given by $2\pi r$. So, the distance traveled would be $2\pi r$. This corresponds to option (1), but the question asks for displacement. Option (2) πr is half the circumference (distance for half a revolution). Option (4) $2r$ is the diameter (displacement if the particle moved from one end of a diameter to the other, i.e., half a revolution).

The question specifically asks for *displacement*.

Quick Tip

Always distinguish between distance and displacement: **Distance:** Total path length. Always positive or zero. Scalar.

Displacement: Shortest distance between initial and final points (a straight line). Vector. Can be positive, negative, or zero.

For any motion where the object returns to its starting point (like one full revolution on a circle, or a round trip), the total displacement is always zero, even if the distance traveled is non-zero.

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61. In velocity time graph, area under v-t graph represents :

(1) Displacement

(2) Velocity

(3) Acceleration

(4) Time

Correct Answer: (1) Displacement

Solution: Concept: A velocity-time (v-t) graph plots the velocity of an object against time. The physical quantities displacement and acceleration can be determined from this graph.

Step 1: Understanding the relationship between velocity, time, and displacement

Velocity is defined as the rate of change of displacement, $v = \frac{\Delta d}{\Delta t}$, where Δd is the change in displacement and Δt is the time interval. For constant velocity, displacement $d = v \times t$. If velocity is not constant, we consider a small time interval dt during which the velocity v can be considered approximately constant. The small displacement dd during this interval is $dd = v \cdot dt$. To find the total displacement over a period from time t_1 to t_2, we integrate this expression: $d = \int_{t_1}^{t_2} v\, dt$. Mathematically, the definite integral $\int_{t_1}^{t_2} v\, dt$ represents the area under the curve of v plotted against t, from t_1 to t_2.

Step 2: Interpreting the area under the v-t graph Consider a simple case where an object moves with a constant velocity v for a time t. On the v-t graph, this would be a horizontal line at height v. The area under this line from $t = 0$ to t would be a rectangle with height v and width t. Area = height × width = $v \times t$. Since displacement for constant velocity is also $v \times t$, the area under the v-t graph is equal to the displacement. This principle extends to any shape under the v-t graph; the area always represents the displacement.

Step 3: Distinguishing from slope The slope of a velocity-time graph represents acceleration. Slope is calculated as the change in velocity divided by the change in time ($\frac{\Delta v}{\Delta t}$), which is the definition of acceleration. Option (3) Acceleration is incorrect as it is represented by the slope, not the area.

Therefore, the area under a velocity-time graph represents displacement.

Quick Tip

For a velocity-time (v-t) graph: The **area** under the graph gives the **displacement**. The **slope** of the graph gives the **acceleration**. Remember these two key interpretations, as they are fundamental in kinematics.

---

62. Rocket works on the principle of conservation of :

(1) Mass

(2) Energy

(3) Charge

(4) Momentum

Correct Answer: (4) Momentum

Solution: Concept: The operation of a rocket is a classic example of Newton's Third Law of Motion and the principle of conservation of linear momentum.
Step 1: Understanding Rocket Propulsion A rocket expels hot gases (burnt fuel) downwards at a very high velocity. According to Newton's Third Law, for every action, there is an equal and opposite reaction. The action is the downward expulsion of gases. The reaction is an upward force (thrust) exerted on the rocket by these gases, which propels the rocket upwards.
Step 2: Applying the Principle of Conservation of Linear Momentum The principle of conservation of linear momentum states that if no external forces act on a system, the total linear momentum of the system remains constant. Consider the rocket and its fuel as an isolated system. Initially (before launching or before expelling a particular segment of gas), the rocket and fuel might be at rest or moving with a certain momentum. When the rocket expels a mass of gas (m_g) downwards with a velocity (v_g), the gas acquires a downward momentum ($p_g = m_g v_g$). To conserve the total momentum of the system (rocket + expelled gas), the rocket itself must gain an equal and opposite momentum. If the rocket has mass M_R and gains an upward velocity V_R, its upward momentum is $P_R = M_R V_R$. In a simplified view, if the initial momentum of the system was zero (rocket at rest), then after expelling gas: Total final momentum = Momentum of rocket + Momentum of gas = 0 $M_R V_R + m_g v_g = 0$ (where v_g would be negative if V_R is positive, or vice versa, due to opposite directions). This means $M_R V_R = -m_g v_g$. The rocket gains momentum in one direction, and the expelled gases gain an equal magnitude of momentum in the opposite direction.
Step 3: Evaluating other options

- **Conservation of Mass (1):** While mass is conserved in a non-relativistic sense for the universe, the mass of the rocket itself decreases as it burns and expels fuel. So, it's not the primary principle for its propulsion mechanism in the way momentum conservation is.

- **Conservation of Energy (2):** Energy is also conserved (chemical energy of fuel converts to kinetic energy of rocket and gases, and heat), but the fundamental principle explaining the rocket's motion and change in velocity due to expelled mass is momentum conservation.
- **Conservation of Charge (3):** This principle deals with electric charges and is not directly relevant to the mechanical propulsion of a rocket.

The recoil of the rocket due to the ejection of mass is best explained by the conservation of momentum.
Therefore, a rocket works on the principle of conservation of momentum.

Quick Tip

Rocket propulsion is a prime example of Newton's Third Law (action-reaction) and the conservation of linear momentum. The rocket pushes gases down (action), and the gases push the rocket up (reaction). The total momentum of the rocket-gas system remains conserved if external forces like air resistance are ignored.

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63. Force on an object can not change its :

(1) Shape

(2) Mass

(3) Direction

(4) Speed

Correct Answer: (2) Mass

Solution: Concept: A force is an interaction that, when unopposed, will change the motion of an object. It can also deform an object. Mass is an intrinsic property of an object.
Step 1: Understanding the effects of force According to Newton's Laws of Motion and general physics principles:

- **Change in Speed (4):** A net force can cause an object to accelerate (Newton's Second Law, $F = ma$). Acceleration is the rate of change of velocity. If velocity changes, its magnitude (speed) can change. For example, pushing a toy car can make it go faster.

- **Change in Direction (3):** Force is a vector. If a force is applied in a direction different from the object's current motion, it can change the direction of motion. For example, a planet orbiting a star is constantly changing direction due to the gravitational force.
- **Change in Shape (1):** Forces can cause deformation. If you squeeze a rubber ball or stretch a spring, its shape changes due to the applied force.

Step 2: Considering mass Mass (2): Mass is a measure of the amount of matter in an object. In classical (non-relativistic) physics, the mass of an object is considered an intrinsic and constant property. Applying a force to an object does not change the amount of matter it contains. For example, if you push a 5 Kg box, its mass remains 5 Kg; it doesn't become 4 Kg or 6 Kg just because you pushed it. (Note: In relativistic physics, at very high speeds close to the speed of light, mass can increase. Also, in nuclear reactions, mass can be converted to energy. However, for typical scenarios implied by such a question, we assume classical physics where mass is constant unless matter is added or removed from the object.)
Conclusion: A force can change an object's speed, direction of motion, and shape. However, applying a force does not change the object's mass.
Therefore, force on an object cannot change its mass.

Quick Tip

Force can cause: 1. A change in the state of motion (start moving, stop moving, change speed, change direction). This is described by $F = ma$. 2. A change in the shape or size of an object (deformation). Mass, however, is the amount of matter in an object and is not changed by applying a force in classical mechanics.

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64. The mass of a goods lorry is 3500 Kg and the mass of goods loaded on it is 1500 Kg. If the lorry is moving with a velocity 10m/s. What will be its momentum ?

(1) 25000 Kg m/s

(2) 30000 Kg m/s

(3) 40000 Kg m/s

(4) 50000 Kg m/s

Correct Answer: (4) 50000 Kg m/s

Solution: Concept: Momentum (linear momentum) of an object is a measure of its mass in motion. It is defined as the product of the object's mass and its velocity.

Step 1: Identify the given quantities Mass of the lorry, $m_{\text{lorry}} = 3500$ Kg Mass of the goods loaded, $m_{\text{goods}} = 1500$ Kg Velocity of the lorry (with goods), $v = 10$ m/s

Step 2: Calculate the total mass The total mass (M) of the moving system (lorry + goods) is the sum of the mass of the lorry and the mass of the goods.

$$M = m_{\text{lorry}} + m_{\text{goods}}$$

$$M = 3500 \text{ Kg} + 1500 \text{ Kg}$$

$$M = 5000 \text{ Kg}$$

Step 3: Recall the formula for momentum Linear momentum (p) is given by the formula:

$$p = \text{mass} \times \text{velocity}$$

$$p = M \times v$$

Step 4: Substitute the total mass and velocity into the formula

$$p = (5000 \text{ Kg}) \times (10 \text{ m/s})$$

Step 5: Calculate the momentum

$$p = 50000 \text{ Kg} \cdot \text{m/s}$$

The unit of momentum is Kg $\cdot$ m/s.

Therefore, the momentum of the lorry will be 50000 Kg m/s.

Quick Tip

Momentum (p) is calculated as $p = mv$. 1. First, find the total mass of the moving object or system. 2. Then, multiply this total mass by its velocity. Ensure units are consistent: mass in Kg, velocity in m/s, so momentum will be in Kg m/s.

---

65. SI unit of weight is :

(1) Kg

(2) Kg ms^{-1}

(3) N

(4) None of these

Correct Answer: (3) N

Solution: Concept: Weight is a type of force. Specifically, it is the gravitational force exerted on an object by a celestial body (like Earth, Moon, etc.).
Step 1: Define weight Weight (W) of an object is given by the product of its mass (m) and the acceleration due to gravity (g) at its location:

$$W = mg$$

Step 2: Determine the SI units of mass and acceleration due to gravity

- The SI unit of mass (m) is the kilogram (Kg).
- The SI unit of acceleration (g, or any acceleration) is meters per second squared (m/s^2 or ms^{-2}).

Step 3: Determine the SI unit of force According to Newton's Second Law of Motion, force (F) is equal to mass (m) times acceleration (a): $F = ma$. The SI unit of force is the Newton (N). One Newton is defined as the force required to accelerate a 1 kilogram mass by 1 meter per second squared. So, 1 N $= 1$ Kg $\cdot$ m/s^2.
Step 4: Relate the unit of weight to the unit of force Since weight ($W = mg$) is a force, its SI unit must be the same as the SI unit of force. Unit of weight = Unit of mass × Unit of acceleration due to gravity Unit of weight = Kg $\times$ m/s^2 This combination, Kg $\cdot$ m/s^2, is defined as the Newton (N).
Step 5: Evaluate the given options

- (1) Kg: This is the SI unit of mass, not weight.
- (2) Kg ms^{-1}: This is the SI unit of momentum ($p = mv$), not weight.
- (3) N: This is the SI unit of force, and therefore, of weight.
- (4) None of these: Incorrect, as (3) is the correct unit.

Therefore, the SI unit of weight is the Newton (N).

Quick Tip

Weight is a force, specifically the force of gravity. The SI unit for any force is the Newton (N). Mass is the amount of matter, its SI unit is Kilogram (Kg). Do not confuse mass and weight; they are different physical quantities with different units.

---

66. Which of the following materials cannot be used to make a lens ?

(1) Water

(2) Glass

(3) Clay

(4) Plastic

Correct Answer: (3) Clay

Solution: Concept: A lens is an optical device that transmits and refracts light, converging or diverging the beam. To function as a lens, a material must primarily be transparent or translucent to the type of radiation it is intended for (usually visible light).

Step 1: Properties required for a lens material For a material to be used to make a lens for visible light, it should ideally have the following properties:

- **Transparency:** It must allow light to pass through it with minimal scattering or absorption.
- **Refractive Index:** It must have a refractive index different from the surrounding medium (usually air) to be able to bend light.
- **Homogeneity:** It should be optically uniform.
- **Workability:** It should be possible to shape it into the curved surfaces required for a lens.
- **Durability and Stability:** It should be reasonably hard, stable, and not degrade easily.

The most critical property for basic function is transparency.

Step 2: Evaluate the given materials

- **(1) Water:** Water is transparent to visible light. It has a refractive index (≈ 1.33) different from air. Liquid lenses exist, and water can be contained in a shaped transparent container to act as a lens (e.g., a spherical flask filled with water can act as a magnifying lens). So, water can be used.
- **(2) Glass:** Glass is a very common material for making lenses due to its excellent transparency, workability, and stability. Various types of optical glass are specifically designed for lens manufacturing. So, glass can be used.
- **(3) Clay:** Clay is an opaque material. It does not allow visible light to pass through it. Therefore, it cannot refract light in the way required for a lens to form an image. So, clay cannot be used.
- **(4) Plastic:** Many types of plastics (e.g., acrylic, polycarbonate) are transparent and are widely used to make lenses, especially for eyeglasses, contact lenses, and inexpensive optical instruments. They are lightweight and shatter-resistant. So, plastic can be used.

Conclusion: Among the given options, clay is the only material that is opaque and thus cannot be used to make a functional lens for visible light.
Therefore, clay cannot be used to make a lens.

Quick Tip

The primary requirement for a material to make a lens is that it must be transparent to the light it is intended to manipulate. Opaque materials, like clay, block light and cannot form lenses. Common lens materials like glass and plastic are chosen for their transparency and refractive properties.

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67. S. I. unit of power of lens is :

(1) Diopter

(2) Decibel

(3) Meter

(4) Gauss

Correct Answer: (1) Diopter

Solution: Concept: The power of a lens is a measure of its ability to converge or diverge light. It is defined as the reciprocal of its focal length.

Step 1: Define the power of a lens The power (P) of a lens is given by the formula:

$$P = \frac{1}{f}$$

where f is the focal length of the lens.

Step 2: SI unit of focal length The focal length (f) is a distance. The SI unit of distance is the meter (m). For the power of a lens to be expressed in its standard unit (Diopter), the focal length f **must be expressed in meters**.

Step 3: Determine the SI unit of power If f is in meters (m), then the unit of power P would be m^{-1} (reciprocal meter or per meter). This unit, m^{-1}, is specifically named the Diopter (D). So, 1 Diopter (D) $= 1\ \text{m}^{-1}$.

Step 4: Evaluate the given options

- **(1) Diopter:** This is the correct SI unit for the power of a lens.
- **(2) Decibel (dB):** This is a logarithmic unit used to express ratios, commonly for sound level or signal power in electronics. It is not the unit for the power of a lens.
- **(3) Meter (m):** This is the SI unit of length (and focal length), not power of a lens.
- **(4) Gauss (G):** This is a unit of magnetic flux density (magnetic field strength) in the CGS system. It is not related to the power of a lens.

Therefore, the S.I. unit of power of lens is Diopter.

Quick Tip

The power of a lens (P) is $1/f$, where f is the focal length. The SI unit of power is the Diopter (D). Crucially, to calculate power in Diopters, the focal length f must be in meters. For example, a lens with a focal length of 50 cm (0.5 m) has a power of $1/0.5 = +2$ D (if converging).

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68. The angle of incidence of any light passing through the centre of curvature of a spherical mirror is :

(1) 0°

(2) 45°

(3) 90°

(4) 60°

Correct Answer: (1) 0°

Solution: Concept: For spherical mirrors, a line passing through the centre of curvature (C) and any point on the mirror's surface is a normal to the surface at that point. The angle of incidence is the angle between the incident ray and the normal at the point of incidence.
Step 1: Understanding the Centre of Curvature (C) The centre of curvature of a spherical mirror is the centre of the sphere of which the mirror forms a part.
Step 2: Properties of a line from the Centre of Curvature to the mirror surface Any line drawn from the centre of a sphere to its surface is perpendicular (normal) to the surface at that point. This applies to spherical mirrors as well. So, if a ray of light is directed towards the mirror along a line that would pass through the centre of curvature, that line itself acts as the normal at the point where the ray strikes the mirror.
Step 3: Defining the Angle of Incidence The angle of incidence (i) is the angle between the incident ray and the normal to the reflecting surface at the point of incidence.
Step 4: Analyzing the specific case The question states that the light ray is "passing through the centre of curvature". This means the incident ray itself lies along a radius of the sphere, and thus it lies along the normal to the mirror surface at the point of incidence. When the incident ray coincides with the normal, the angle between the incident ray and the normal is zero. Angle of incidence, $i = 0^\circ$.
Step 5: Reflection of such a ray According to the laws of reflection, the angle of reflection (r) is equal to the angle of incidence (i). So, if $i = 0^\circ$, then $r = 0^\circ$. This means the reflected ray will also make an angle of 0° with the normal. Consequently, the reflected ray will travel back along the same path as the incident ray (retraces its path).
Therefore, the angle of incidence of any light passing through the centre of curvature of a spherical mirror is 0°.

Quick Tip

For spherical mirrors: A ray of light passing through the Centre of Curvature (C) strikes the mirror normally (perpendicularly). This means the incident ray itself is the normal at that point. Therefore, the angle of incidence is $0°$, and the ray retraces its path after reflection (angle of reflection is also $0°$).

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69. Refractive index of water w.r.t. air is 1.33. What is the refractive index of air w.r.t. water ?

(1) 0.75

(2) 0.50

(3) 75.0

(4) 0.25

Correct Answer: (1) 0.75

Solution: Concept: The refractive index of medium 2 with respect to medium 1 (n_{21} or ${}^{1}n_{2}$) is related to the refractive index of medium 1 with respect to medium 2 (n_{12} or ${}^{2}n_{1}$) by the principle of reversibility of light.

Step 1: Understanding the notation and given information "Refractive index of water w.r.t. air" can be written as $n_{\text{water, air}}$ or ${}_{\text{air}}n_{\text{water}}$. Given: ${}_{\text{air}}n_{\text{water}} = 1.33$. This means when light travels from air into water, its speed changes by a factor related to 1.33, or $\frac{\text{speed of light in air}}{\text{speed of light in water}} = 1.33$.

Step 2: What needs to be found We need to find the "refractive index of air w.r.t. water". This can be written as $n_{\text{air, water}}$ or ${}_{\text{water}}n_{\text{air}}$. This would represent $\frac{\text{speed of light in water}}{\text{speed of light in air}}$.

Step 3: Applying the principle of reversibility The principle of reversibility states that if a ray of light, after suffering any number of reflections and/or refractions, has its final path reversed, it travels back along its entire original path. This leads to the relationship between ${}_{\text{air}}n_{\text{water}}$ and ${}_{\text{water}}n_{\text{air}}$:

$${}_{\text{water}}n_{\text{air}} = \frac{1}{{}_{\text{air}}n_{\text{water}}}$$

In general, for any two media 1 and 2:

$$n_{12} = \frac{1}{n_{21}}$$

Step 4: Substitute the given value and calculate We are given ${}_{\text{air}}n_{\text{water}} = 1.33$. So, the refractive index of air with respect to water is:

$${}_{\text{water}}n_{\text{air}} = \frac{1}{1.33}$$

The value 1.33 is approximately $4/3$.

$${}_{\text{water}}n_{\text{air}} = \frac{1}{4/3} = \frac{3}{4}$$

Now, convert the fraction $3/4$ to a decimal:

$$\frac{3}{4} = 0.75$$

Step 5: Check with options The calculated value is 0.75, which matches option (1). Therefore, the refractive index of air w.r.t. water is 0.75.

Quick Tip

The refractive index of medium 'a' with respect to medium 'b' (${}_bn_a$) is the reciprocal of the refractive index of medium 'b' with respect to medium 'a' (${}_an_b$). So, ${}_bn_a = 1/{}_an_b$. If you know $n_{\text{water w.r.t. air}} = 1.33 \approx 4/3$, then $n_{\text{air w.r.t. water}}$ will be $1/(4/3) = 3/4 = 0.75$.

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70. The focal length of a convex lens is 18 cm and the size of the image is a quarter of the object. The object is situated at a distance of:

(1) 90 cm

(2) 54 cm

(3) 22.5 cm

(4) 60 cm

Correct Answer: (2) 54 cm

Solution: Concept: This problem involves the lens formula and magnification for a convex lens.

- Lens Formula: $\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$
- Magnification Formula: $m = \frac{v}{u}$

Where f is focal length, u is object distance, v is image distance, and m is magnification.

Sign Convention for convex lens:

- f is positive.
- u is negative (object typically to the left).
- v is positive for a real image.
- m is negative for a real, inverted image.

Step 1: Given information and selected option

- Focal length, $f = +18$ cm.
- The selected correct answer states the object distance is 54 cm. So, we take $u = -54$ cm.

The problem text also states: "the size of the image is a quarter of the object." For a convex lens, a diminished image implies it's real and inverted, so this would mean $m = -1/4$. We will check this condition later.

Step 2: Calculate image distance (v) using $u = -54$ cm Using the lens formula:

$$\frac{1}{18} = \frac{1}{v} - \frac{1}{-54}$$
$$\frac{1}{18} = \frac{1}{v} + \frac{1}{54}$$
$$\frac{1}{v} = \frac{1}{18} - \frac{1}{54}$$

To subtract, find a common denominator (54): $\frac{1}{18} = \frac{3}{54}$.

$$\frac{1}{v} = \frac{3}{54} - \frac{1}{54} = \frac{2}{54} = \frac{1}{27}$$

So, image distance $v = +27$ cm. The positive sign means the image is real.

Step 3: Calculate magnification (m) for $u = -54$ cm Using the magnification formula:

$$m = \frac{v}{u} = \frac{27 \text{ cm}}{-54 \text{ cm}} = -\frac{1}{2}$$

Step 4: Interpretation of the result and comparison with problem statement The calculated magnification is $m = -1/2$. This means the image is inverted (due to the negative

sign) and its size is half the size of the object. However, the problem statement says, "the size of the image is a quarter of the object" (which implies $m = -1/4$). If we were to use $m = -1/4$, then $v = mu = (-1/4)u$. Substituting into the lens formula:
$\frac{1}{18} = \frac{1}{(-u/4)} - \frac{1}{u} = -\frac{4}{u} - \frac{1}{u} = -\frac{5}{u}$. This gives $u = -5 \times 18 = -90$ cm. This shows a discrepancy: if the object distance is 54 cm (Option 2), the image is half the object's size. If the image is a quarter of the object's size (as per text), the object distance should be 90 cm (Option 1).

Conclusion based on selected option: Assuming Option (2) 54 cm is the intended answer for the object distance, the resulting magnification is $m = -1/2$.

Quick Tip

When solving lens problems: 1. List knowns (f, u or v, m) with correct signs. 2. Use the lens formula $\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$ and magnification $m = v/u$. If a problem's text and a given "correct" option lead to different physical outcomes (like different magnifications), it's important to note the inconsistency. For this problem, an object at 54 cm yields $m = -1/2$, while $m = -1/4$ (text) implies $u = -90$ cm.

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71. The range of vision of a normal human eye is from :

(1) 100m to 25cm

(2) infinity to 25m

(3) 1km to 25cm

(4) infinity to 25cm

Correct Answer: (4) infinity to 25cm

Solution: Concept: The range of vision describes the span of distances over which a normal human eye can see objects clearly. It is defined by two points: the near point and the far point.

Step 1: Define Near Point and Far Point

- **Near Point (Least Distance of Distinct Vision - LDDV):** This is the closest distance an object can be to the eye and still be seen clearly and without strain. For a young adult with normal vision, this is about 25 cm.

- **Far Point:** This is the farthest distance an object can be from the eye and still be seen clearly. For a normal eye, the far point is at infinity (∞). This means we can see very distant objects like stars.

Step 2: Determine the Range of Vision The range of vision for a normal eye is from its near point to its far point. So, this range is from 25 cm to infinity.

Step 3: Match with the given options The options provided list the far point first, then the near point.

- Option (1) 100m to 25cm: Incorrect far point.
- Option (2) infinity to 25m: Incorrect near point unit (should be cm).
- Option (3) 1km to 25cm: Incorrect far point.
- Option (4) infinity to 25cm: Correctly states the far point (infinity) and the near point (25 cm).

Thus, the range of vision is from 25 cm to infinity.

Quick Tip

For a normal human eye:

- Closest clear vision (Near Point) = 25 cm
- Farthest clear vision (Far Point) = Infinity (∞)

The range of vision spans all distances between these two points.

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72. Stars twinkle due to :

(1) atmospheric refraction

(2) atmospheric reflection

(3) scattering of light

(4) dispersion of light

Correct Answer: (1) atmospheric refraction

Solution: Concept: The twinkling of stars is an optical effect caused by starlight passing through Earth's atmosphere.

Step 1: Understanding Starlight and Earth's Atmosphere Stars are very far away, so they appear as point sources of light. Earth's atmosphere is made of layers of air with varying temperatures and densities. This means the optical density (and thus refractive index) of air changes from point to point and fluctuates over time due to air currents.

Step 2: The Phenomenon of Atmospheric Refraction Refraction is the bending of light as it passes from one medium to another of different optical density. As starlight enters Earth's atmosphere and travels through these fluctuating layers:

- The light path continuously bends by small amounts.
- This random bending causes the apparent position of the star to seem to shift slightly.
- The amount of starlight reaching our eye also varies. Sometimes more light reaches us (star appears brighter), and sometimes less (star appears dimmer).

This rapid variation in brightness and apparent position is what we perceive as twinkling.

Step 3: Why other options are not the primary cause

- **Atmospheric reflection:** While some reflection occurs, the dominant effect causing twinkling is the bending of light (refraction) through atmospheric layers.
- **Scattering of light:** This causes the blue color of the sky but isn't the main reason for twinkling.
- **Dispersion of light:** This is the splitting of light into colors (like a rainbow) and is not the primary cause of the twinkling effect.

Therefore, atmospheric refraction is the cause of stars twinkling. Planets, being closer and appearing as extended sources, do not twinkle as much because the variations from different points on the planet average out.

Quick Tip

Imagine looking at something through the wavy hot air above a fire – it shimmers. Stars twinkle for a similar reason: starlight passes through Earth's unsteady atmosphere, causing it to bend randomly. This changing path makes the star's brightness and position seem to flicker.

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73. Name the two phenomenon involved in the formation of rainbow :

(1) Dispersion and reflection of light

(2) Refraction and reflection of light

(3) Scattering and refraction of light

(4) Scattering and reflection of light

Correct Answer: (1) Dispersion and reflection of light

Solution: Concept: A rainbow is formed when sunlight interacts with water droplets in the atmosphere. Three main optical phenomena are involved: refraction, dispersion, and total internal reflection.

Step 1: Processes within a raindrop leading to a rainbow

1. **Refraction and Dispersion:** Sunlight enters a raindrop. It refracts (bends) because water has a different optical density than air. Since the refractive index of water varies slightly for different colors (wavelengths) of light, white sunlight is dispersed (split) into its spectrum of colors (e.g., red, orange, yellow, green, blue, violet). Violet light bends more than red light.

2. **Total Internal Reflection:** The dispersed light rays travel to the back inner surface of the raindrop. Here, they undergo total internal reflection (a specific type of reflection where all light is reflected back into the droplet) if they strike the surface at an angle greater than the critical angle.

3. **Refraction (again):** The reflected light rays then travel to the front surface of the raindrop and refract again as they exit the droplet, passing from water back into air. This further separates the colors.

Step 2: Identifying the key phenomena from the options The question asks for "two phenomenon".

- **Dispersion of light** is essential because it separates sunlight into colors, which is the defining feature of a rainbow.
- **Reflection of light** (specifically, total internal reflection) is essential because it directs the light back out of the raindrop towards the observer.
- **Refraction of light** is also fundamental as it's necessary for light to enter and exit the droplet, and it's the underlying cause of dispersion.

Option (1) "Dispersion and reflection of light" lists two of these crucial phenomena. Dispersion creates the colors, and reflection sends them back. While refraction is also key, dispersion is a specific and visually critical outcome of refraction in this context.

Quick Tip

For a rainbow, remember these key events inside a raindrop: 1. Sunlight enters $\rightarrow$ Bends (Refraction) AND Splits into colors (Dispersion). 2. Colors bounce off the back of the raindrop (Total Internal Reflection). 3. Colors exit $\rightarrow$ Bend again (Refraction). Option (1) highlights the color-splitting (Dispersion) and the bounce-back (Reflection).

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74. Unit of electrical current is :

(1) ampere

(2) coulomb

(3) joule

(4) volt

Correct Answer: (1) ampere

Solution: Concept: Each fundamental physical quantity has a standard unit in the International System of Units (SI).

Step 1: Define Electrical Current Electrical current (symbol: I) is the rate of flow of electric charge. If a charge Q flows past a point in a time t, the current is $I = Q/t$.

Step 2: Identify the SI unit of Electrical Current The SI unit for electrical current is the **ampere** (symbol: A). One ampere is defined as one coulomb of charge flowing per second (1 A = 1 C/s).

Step 3: Units of other options

- **Coulomb (C):** The SI unit of electric charge (Q).
- **Joule (J):** The SI unit of energy or work (W or E).
- **Volt (V):** The SI unit of electric potential difference or voltage (V). (1 V = 1 J/C).

Therefore, the unit of electrical current is the ampere.

Quick Tip

Associate these basic electrical quantities with their SI units:

- Current $\rightarrow$ Ampere (A)
- Charge $\rightarrow$ Coulomb (C)
- Voltage $\rightarrow$ Volt (V)
- Energy $\rightarrow$ Joule (J)

---

75. An electric bulb is connected to a 220 V generator. The current is 0.50 A. What is the power of bulb ?

(1) 110 watt

(2) 100 watt

(3) 220 watt

(4) 55 watt

Correct Answer: (1) 110 watt

Solution: Concept: Electric power (P) is the rate at which electrical energy is consumed or produced. It's related to voltage (V) and current (I).

Step 1: Given information

- Voltage (V) = 220 V
- Current (I) = 0.50 A

Step 2: Formula for Electric Power The formula to calculate electric power when voltage and current are known is:

$$P = V \times I$$

Where P is in watts (W), V in volts (V), and I in amperes (A).
Step 3: Calculate the power Substitute the given values into the formula:

$$P = 220 \text{ V} \times 0.50 \text{ A}$$

$$P = 110 \text{ W}$$

(Note: Multiplying by 0.50 is the same as dividing by 2).
The power of the bulb is 110 watts. This matches option (1).

Quick Tip

The key formula for power is $P = VI$. Just multiply the voltage by the current: $220 \times 0.50 = 110$. Remember that 0.50 is half, so you're finding half of 220.

---

76. The electric device used for producing electric current is called a :

(1) generator

(2) galvanometer

(3) ammeter

(4) motor

Correct Answer: (1) generator

Solution: Concept: Different electrical devices have specific functions related to electric current. This question asks about the device that *produces* current.
Step 1: Understanding the functions of the listed devices

- **(1) Generator:** A generator converts mechanical energy (like rotation from a turbine) into electrical energy. This process *produces* an electric current. It works based on electromagnetic induction.

- **(2) Galvanometer:** A galvanometer is a sensitive instrument used to *detect* and measure small electric currents. It does not produce current.
- **(3) Ammeter:** An ammeter is an instrument used to *measure* the magnitude of electric current flowing in a circuit. It does not produce current.
- **(4) Motor (Electric Motor):** An electric motor converts electrical energy into mechanical energy (motion). It *uses* electric current, it doesn't produce it.

Step 2: Identifying the current-producing device From the descriptions, the **generator** is the device that produces electric current.

Quick Tip

Think about energy conversion:

- **Generator:** Mechanical energy $\rightarrow$ Electrical energy (PRODUCES current)
- **Motor:** Electrical energy $\rightarrow$ Mechanical energy (USES current)
- Galvanometer/Ammeter: MEASURE current

---

77. At the time of short circuit the current in the circuit :

(1) reduces substantially

(2) does not change

(3) increases heavily

(4) vary continuously

Correct Answer: (3) increases heavily

Solution: Concept: A short circuit is a low-resistance path unintentionally created in an electrical circuit. Ohm's Law ($I = V/R$) helps understand its effect on current.

Step 1: What is a short circuit? A short circuit occurs when the current bypasses its intended path (the load, like a bulb or appliance) and instead flows through an alternative path with very low resistance. For example, if the live and neutral wires touch directly.

Step 2: Applying Ohm's Law ($I = V/R$)

- V is the voltage of the source (e.g., 220V).
- R is the resistance of the circuit.
- I is the current flowing.

In a normal circuit, the load provides a certain resistance (R_{load}). In a short circuit, the resistance of the path (R_{short}) becomes extremely small (close to zero).

Step 3: Effect on Current Since $I = V/R$: If R becomes very small (as in R_{short}), and V remains constant, the current I must become very large. $I_{\text{short}} = V/R_{\text{short}}$. Since R_{short} is tiny, I_{short} is huge. This is why the current "increases heavily."

Step 4: Consequences This large current can cause overheating of wires (due to $P = I^2R$ heating effect), leading to melting insulation, fires, or damage to the power source. Protective devices like fuses or circuit breakers are designed to break the circuit if such high currents occur.

Quick Tip

Think of Ohm's Law: Current = Voltage / Resistance. "Short circuit" = Resistance becomes very, very small. If you divide the same voltage by a much smaller resistance, the current becomes much, much larger.

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78. What kind of mirror would be best suited for use in Solar Cooker ?

(1) concave

(2) convex

(3) plain

(4) plano-concave

Correct Answer: (1) concave

Solution: Concept: Solar cookers use sunlight to generate heat for cooking. This often requires concentrating sunlight onto a small area.

Step 1: Purpose of a mirror in a solar cooker The main goal is to collect solar energy from a large area and focus it onto the cooking pot or a specific heating zone to achieve high temperatures.

Step 2: Properties of different mirror types regarding parallel light (like sunlight)

- **(1) Concave Mirror:** This mirror is curved inwards. It converges (brings together) parallel rays of light to a single point called the focus. This concentrating effect is ideal for increasing temperature. Example: a satellite dish shape.
- **(2) Convex Mirror:** This mirror is curved outwards. It diverges (spreads out) parallel rays of light. This would diffuse the solar energy, not concentrate it. Example: side-view mirrors on cars (often say ”objects in mirror are closer...”).
- **(3) Plain Mirror (Plane Mirror):** This is a flat mirror. It reflects light without converging or diverging it significantly from a single mirror. It can redirect light but doesn't focus it to a point to increase intensity.
- **(4) Plano-concave Mirror:** ”Plano-concave” is typically a term for a lens (one flat side, one concave side). If it were a mirror, it would likely refer to its concave reflecting surface, making it functionally a concave mirror.

Step 3: Choosing the best mirror To concentrate sunlight and generate high temperatures for cooking, a **concave mirror** is the most suitable choice due to its converging property.

Quick Tip

Solar cookers need to focus sunlight to get hot.

- **Concave** mirrors **C**oncentrate light.
- **Convex** mirrors **S**pread light (disperse/diverge).
- **Plane** mirrors just reflect without focusing.

So, concave is the answer.

---

79. A Solar water heater cannot be used to get hot water on :

(1) cloudy day

(2) sunny day

(3) a hot day

(4) a windy day

Correct Answer: (1) cloudy day

Solution: Concept: Solar water heaters rely on solar radiation (sunlight) as their energy source to heat water.

Step 1: How solar water heaters work They have collectors (often black panels) that absorb sunlight. This absorbed energy heats water that circulates through the collectors. The effectiveness directly depends on the amount of sunlight received.

Step 2: Analyzing the conditions

- **(1) Cloudy day:** Clouds block most of the direct sunlight. Without significant sunlight reaching the collectors, the solar water heater cannot absorb enough energy to heat the water effectively. So, it "cannot be used" or will perform very poorly.
- **(2) Sunny day:** This is the ideal condition. Ample sunlight allows the heater to work efficiently.
- **(3) A hot day:** Hot ambient temperature can even improve efficiency by reducing heat loss from the collectors. If it's hot and sunny, it works well. If it's hot but cloudy, the lack of sun is still the limiting factor.
- **(4) A windy day:** Wind can increase heat loss from the collectors, potentially reducing efficiency (water might not get as hot or take longer). However, if it's sunny, the heater will still work, just less optimally than on a calm sunny day. It can still be used.

Step 3: Identifying the most prohibitive condition The primary requirement for a solar water heater is sunlight. A **cloudy day** deprives it of this essential energy source, making it largely ineffective.

Quick Tip

"Solar" means it uses the sun. No sun, no (or very little) hot water.

- **Cloudy day** = No/little sun → Heater doesn't work well.
- Sunny day = Lots of sun → Heater works well.

80. The change in focal length of an eye lens is caused by the action of :

(1) pupil

(2) retina

(3) ciliary muscle

(4) iris

Correct Answer: (3) ciliary muscle

Solution: Concept: The human eye can focus on objects at different distances by changing the shape (and thus the focal length) of its lens. This process is called accommodation.

Step 1: The Role of the Eye Lens and Accommodation The eye lens is a flexible, convex lens. To form a clear image on the retina for objects at various distances, its converging power (related to focal length) must change.

Step 2: How Ciliary Muscles control the Lens The eye lens is suspended by suspensory ligaments which are attached to the **ciliary muscles**.

- **For distant objects:** The ciliary muscles relax. This causes the suspensory ligaments to pull on the lens, making it thinner and less curved (longer focal length).
- **For near objects:** The ciliary muscles contract. This releases tension on the suspensory ligaments, allowing the elastic lens to become thicker and more curved (shorter focal length).

Thus, the ciliary muscles are responsible for changing the focal length of the eye lens.

Step 3: Functions of other eye parts mentioned

- **(1) Pupil:** The opening in the iris that lets light in. Its size changes to control light intensity, not focal length.
- **(2) Retina:** The light-sensitive layer at the back of the eye where the image is formed. It detects light but doesn't change the lens shape.
- **(4) Iris:** The colored part of the eye that controls the size of the pupil.

Therefore, the ciliary muscles cause the change in focal length.

Quick Tip

The eye focuses by changing the lens shape. Think of ciliary muscles as the "adjusters":

- Ciliary muscles **contract** $\rightarrow$ Lens gets **thicker** (for **near** objects).
- Ciliary muscles **relax** $\rightarrow$ Lens gets **thinner** (for **far** objects).

This change in thickness alters the focal length.

---

81. The phenomenon of splitting of white light through prism into a band of colours is called :

(1) Dispersion of light

(2) Reflection of light

(3) Refraction of light

(4) Scattering of light

Correct Answer: (1) Dispersion of light

Solution: Concept: When white light passes through a prism, it separates into its constituent colors. This phenomenon has a specific name.

Step 1: Understanding the process White light (like sunlight) is actually a mixture of different colors of light (violet, indigo, blue, green, yellow, orange, red - VIBGYOR). Each color has a slightly different wavelength. When white light passes through a medium like a prism, these different colors bend by slightly different amounts due to differences in their speeds within the prism material. This causes them to separate.

Step 2: Defining the terms in the options

- **(1) Dispersion of light:** This is the phenomenon of splitting of white light into its component colors when it passes through a refractive medium (like a prism or a raindrop). This happens because the refractive index of the medium is different for different wavelengths (colors) of light.
- **(2) Reflection of light:** This is the bouncing back of light when it strikes a surface. It does not involve splitting into colors in this context.

- **(3) Refraction of light:** This is the bending of light as it passes from one medium to another. While refraction is necessary for dispersion to occur in a prism, dispersion is the specific term for the splitting of colors.

- **(4) Scattering of light:** This is the process by which light is redirected in many directions when it encounters small particles or irregularities. Example: blue color of the sky.

Step 3: Identifying the correct term The phenomenon described – splitting of white light into a band of colors (a spectrum) by a prism – is specifically called **dispersion of light**.

Quick Tip

Think of a rainbow. A prism does something similar to raindrops: it takes white light and "disperses" it, meaning it spreads it out into its individual colors (like red, orange, yellow, green, blue, violet). So, "Dispersion" is the key word for this color-splitting effect.

---

82. A current of 0.5 ampere is drawn by a filament of an electric bulb for 10 minutes. Find the amount of electrical charge :

(1) 300 Coulomb

(2) 600 Coulomb

(3) 60 Coulomb

(4) 30 Coulomb

Correct Answer: (1) 300 Coulomb

Solution: Concept: Electric current (I) is defined as the rate of flow of electric charge (Q) per unit time (t). The relationship is $I = Q/t$.

Step 1: Given information

- Current (I) = 0.5 ampere (A)

- Time (t) = 10 minutes

We need to find the amount of electrical charge (Q).

Step 2: Convert time to SI units (seconds) The SI unit for time in this formula is seconds.

$$1 \text{ minute} = 60 \text{ seconds}$$

$$10 \text{ minutes} = 10 \times 60 \text{ seconds} = 600 \text{ seconds (s)}$$

Step 3: Use the formula relating current, charge, and time The formula is $I = Q/t$. We need to find Q, so we rearrange the formula:

$$Q = I \times t$$

Step 4: Calculate the charge (Q) Substitute the values of I and t (in seconds) into the formula:

$$Q = 0.5 \text{ A} \times 600 \text{ s}$$

$$Q = 300 \text{ Coulombs (C)}$$

(Note: 0.5×600 is the same as half of 600).

The amount of electrical charge that flows through the filament is 300 Coulombs. This matches option (1).

Quick Tip

Remember the formula: Charge (Q) = Current (I) $\times$ Time (t). Crucially, ensure time is in **seconds** if current is in amperes to get charge in coulombs. 10 minutes = $10 \times 60 = 600$ seconds. $Q = 0.5 \times 600 = 300$ Coulombs.

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83. Which of the following is not an example of a bio mass energy source ?

(1) Wood

(2) Gober Gas

(3) Nuclear Energy

(4) Coal

Correct Answer: (3) Nuclear Energy

Solution: Concept: Biomass energy refers to energy derived from organic matter, i.e., material from living or recently living organisms.

Step 1: Understanding Biomass Energy Sources Biomass includes plant-based materials (like wood, crops), animal waste, and organic components of municipal solid waste. These materials store solar energy captured through photosynthesis (in the case of plants). Biomass can be burned directly for heat or converted into biofuels (like ethanol, biodiesel) or biogas (like gobar gas).

Step 2: Analyzing the options

- **(1) Wood:** Wood is a classic example of biomass. It comes from trees (living organisms) and can be burned for energy.
- **(2) Gober Gas (Biogas):** Gober gas is primarily methane produced from the anaerobic digestion (decomposition without oxygen) of animal manure (like cow dung, ”gobar”) or other organic wastes. This is a form of biomass energy.
- **(3) Nuclear Energy:** Nuclear energy is released from atomic nuclei through processes like nuclear fission (splitting of heavy atoms like uranium) or nuclear fusion (joining of light atoms). The source material (e.g., uranium) is a mineral mined from the Earth; it is not derived from living or recently living organisms. Therefore, nuclear energy is not a form of biomass energy.
- **(4) Coal:** Coal is a fossil fuel formed from the remains of ancient plants that lived millions of years ago. While it originated from biological matter, it has undergone significant geological transformation over vast periods. In many classifications, fossil fuels (coal, oil, natural gas) are distinguished from modern biomass because they are non-renewable on human timescales and their extraction/combustion has different environmental implications. However, some broader definitions of ”biomass origin” might include fossil fuels. For this question, given the other options, nuclear energy is clearly the odd one out. (Note: Context sometimes matters for coal. If comparing to nuclear, coal is closer to biomass origin than nuclear is.)

Step 3: Identifying the non-biomass energy source Comparing the options, **Nuclear Energy** is fundamentally different from biomass. It originates from atomic processes, not from organic matter. Wood and Gober Gas are direct examples of biomass. Coal, while a fossil fuel derived from ancient biomass, is distinct from renewable biomass resources. However, Nuclear Energy is unequivocally not biomass.

Therefore, Nuclear Energy is not an example of a biomass energy source.

Quick Tip

"Bio" in biomass means related to life or living things.

- Wood: From trees (living).
- Gober Gas: From animal waste/organic matter (from living things).
- Coal: From ancient plants (ancient life), a fossil fuel.
- Nuclear Energy: From atoms (like uranium), not from living matter.

Nuclear energy is the clear outlier as it's not derived from organic material.

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84. A generator converts :

(1) electrical energy to mechanical energy

(2) mechanical energy to electrical energy

(3) mechanical energy to solar energy

(4) solar energy to electrical energy

Correct Answer: (2) mechanical energy to electrical energy

Solution: Concept: A generator is a device that plays a crucial role in electricity production. Its function is based on the principle of electromagnetic induction.

Step 1: Understanding the function of a generator A generator is designed to produce electrical energy. To do this, it requires an input of some other form of energy. In most common generators (like those in power plants or portable generators), this input is mechanical energy. For example, a turbine (rotated by steam, water, or wind) provides mechanical energy to turn the coils of wire within a magnetic field (or rotate magnets near coils). This relative motion between conductors and magnetic fields induces an electromotive force (voltage), which can drive an electric current.

Step 2: Analyzing the energy conversions in the options

- **(1) electrical energy to mechanical energy:** This describes the function of an *electric* *motor*, not a generator. Motors use electricity to produce motion.

- **(2) mechanical energy to electrical energy:** This accurately describes the function of a *generator*. It takes mechanical input (like rotation) and converts it into electrical output.
- **(3) mechanical energy to solar energy:** This conversion is not typical. Solar energy originates from the sun.
- **(4) solar energy to electrical energy:** This describes the function of a *solar cell* (photovoltaic cell), not a generator in the common electromechanical sense.

Step 3: Identifying the correct conversion A generator's primary role is to convert **mechanical energy into electrical energy**.

Quick Tip

Remember these common energy conversions:

- **Generator:** Mechanical $\rightarrow$ Electrical (Generates electricity)
- **Motor:** Electrical $\rightarrow$ Mechanical (Uses electricity to do work)
- Solar Cell: Solar $\rightarrow$ Electrical

The question is about a ”generator”.

---

85. The speed of light is :

(1) 3×10^8 meter/sec

(2) 3×10^{10} meter/sec

(3) 0.3×10^8 meter/sec

(4) 0.03×10^8 meter/sec

Correct Answer: (1) 3×10^8 meter/sec

Solution: Concept: The speed of light in a vacuum (often denoted by c) is a fundamental physical constant.

Step 1: Recalling the value of the speed of light The speed of light in a vacuum is a precisely defined value, but it is commonly approximated for most calculations. The accepted approximate value is $300,000,000$ meters per second.
Step 2: Expressing this value in scientific notation Scientific notation is a way of writing very large or very small numbers conveniently. To write $300,000,000$ in scientific notation: Move the decimal point to the left until there is one non-zero digit before it. Here, we move it 8 places to the left: 3.00000000 Since we moved it 8 places to the left, we multiply by 10^8. So, $300,000,000$ m/s $= 3 \times 10^8$ m/s.
Step 3: Comparing with the options

- **(1)** 3×10^8 **meter/sec:** This matches our calculated value.
- **(2)** 3×10^{10} **meter/sec:** This is $30,000,000,000$ m/s, which is 100 times too large.
- **(3)** 0.3×10^8 **meter/sec:** This is 3×10^7 m/s or $30,000,000$ m/s, which is 10 times too small.
- **(4)** 0.03×10^8 **meter/sec:** This is 3×10^6 m/s or $3,000,000$ m/s, which is 100 times too small.

The correct value is 3×10^8 meter/sec. (Sometimes written as 2.99792458×10^8 m/s for high precision).

Quick Tip

The speed of light in a vacuum (c) is a very important constant in physics. Remember it as approximately: $c \approx 3 \times 10^8$ m/s This means light travels 300 million meters in one second! It's also common to see it as $300,000$ km/s (since 1 km $= 1000$ m).

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86. The charge of electron is :

(1) 1.6×10^{-19} Coulomb

(2) 16×10^{-19} Coulomb

(3) 0.16×10^{-19} Coulomb

(4) 166×10^{-19} Coulomb

Correct Answer: (1) 1.6×10^{-19} Coulomb

Solution: Concept: The electron is a subatomic particle with a fundamental electric charge. This charge is known as the elementary charge, often denoted by e.

Step 1: Recalling the value of the elementary charge The magnitude of the charge on a single electron (and also on a proton) is a fundamental physical constant. The electron carries a negative charge, while a proton carries a positive charge of the same magnitude. The value of this elementary charge is approximately 1.602×10^{-19} Coulombs. For most introductory purposes, this is rounded to 1.6×10^{-19} Coulombs.

Step 2: Understanding the options The question asks for "The charge of electron". Since an electron is negatively charged, its charge is actually -1.6×10^{-19} C. However, multiple-choice questions often ask for the magnitude of the charge. All options are positive, implying the magnitude is being sought.

- **(1)** 1.6×10^{-19} **Coulomb:** This matches the commonly accepted magnitude of the elementary charge.
- **(2)** 16×10^{-19} **Coulomb:** This is 1.6×10^{-18} C, which is 10 times too large.
- **(3)** 0.16×10^{-19} **Coulomb:** This is 1.6×10^{-20} C, which is 10 times too small.
- **(4)** 166×10^{-19} **Coulomb:** This is 1.66×10^{-17} C, which is significantly different.

Step 3: Identifying the correct value The magnitude of the charge of an electron is 1.6×10^{-19} Coulomb.

Quick Tip

The elementary charge (e) is a fundamental constant: Magnitude of charge on an electron = Magnitude of charge on a proton $\approx 1.6 \times 10^{-19}$ C. Remember:

- Electron's charge: $-e = -1.6 \times 10^{-19}$ C
- Proton's charge: $+e = +1.6 \times 10^{-19}$ C

The options provide magnitudes, so 1.6×10^{-19} C is correct.

87. The coil of a solenoid is made from :

(1) Silicon

(2) Carbon

(3) Germenium

(4) Copper

Correct Answer: (4) Copper

Solution: Concept: A solenoid is an electromagnetic device consisting of a coil of wire. The material of the wire is chosen for its electrical properties.

Step 1: Understanding a Solenoid A solenoid is typically a long coil of wire wound in the shape of a helix. When an electric current passes through the wire, it creates a relatively uniform magnetic field inside the coil. Solenoids are used in various devices like electromagnets, inductors, antennas, valves, etc.

Step 2: Desired properties of the coil material The wire used to make the coil of a solenoid should primarily be a good electrical conductor. This means it should have low electrical resistivity to allow current to flow easily without excessive heat generation (due to resistive losses, $P = I^2R$). It should also be ductile (can be drawn into wires) and reasonably strong.

Step 3: Analyzing the options

- **(1) Silicon (Si):** Silicon is a semiconductor. Its conductivity is much lower than that of metals and depends heavily on temperature and impurities (doping). It is not suitable for making coils that need to carry significant current efficiently.
- **(2) Carbon (C):** Carbon, in forms like graphite, can conduct electricity, but its resistivity is generally higher than that of good metallic conductors. Carbon fibers are strong but not typically the primary choice for solenoid coils where high conductivity is paramount.
- **(3) Germanium (Ge):** Germanium is also a semiconductor, similar to silicon. It's not used for making standard solenoid coils.
- **(4) Copper (Cu):** Copper is an excellent electrical conductor with very low resistivity. It is also ductile, relatively inexpensive, and widely available. These properties make it the most common material for winding coils in solenoids, transformers, motors, and

other electromagnetic devices. Aluminum is sometimes used as an alternative due to its lighter weight and lower cost, but copper generally has better conductivity for a given cross-section.

Step 4: Identifying the most suitable material Considering the need for high electrical conductivity, **Copper** is the standard and most appropriate material for the coil of a solenoid among the given options.

> **Quick Tip**
>
> Solenoids need wire that conducts electricity very well to create a strong magnetic field efficiently.
>
> - Metals like **Copper** and Aluminum are good conductors. Copper is very common for wires.
> - Silicon and Germanium are semiconductors, used in electronics like transistors, not typically for solenoid coils.
> - Carbon can conduct, but not as well as copper for this application.

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88. Magnetic field inside a long straight solenoid carrying current :

(1) is zero

(2) decreases as we move towards its end

(3) increases as we move towards its end

(4) is the same at all points

Correct Answer: (4) is the same at all points

Solution: Concept: A long straight solenoid carrying current produces a magnetic field. The characteristics of this field, particularly inside the solenoid, are important.

Step 1: Visualizing the magnetic field of a solenoid When current flows through the coil of a solenoid, it generates a magnetic field. The field lines inside a long solenoid are nearly parallel to its axis, straight, and evenly spaced, especially far from the ends. Outside the solenoid, the field is much weaker and spreads out, resembling the field of a bar magnet.

Step 2: Properties of the magnetic field inside a long solenoid For an *ideal* long solenoid (infinitely long, or very long compared to its diameter), the magnetic field inside is:

- **Uniform:** The strength (magnitude) of the magnetic field is the same at all points well within the interior of the solenoid, away from the ends. The direction of the field is also uniform, parallel to the axis of the solenoid.
- **Directed along the axis:** The magnetic field lines run parallel to the axis of the solenoid.
- **Calculable:** The magnitude of the field is given by $B = \mu_0 n I$, where μ_0 is the permeability of free space, n is the number of turns per unit length, and I is the current.

Step 3: Considering the ”long straight solenoid” in the question The phrase ”long straight solenoid” implies we are considering a situation close to the ideal case, particularly for points well inside.

- **(1) is zero:** This is incorrect. A current-carrying solenoid produces a magnetic field.
- **(2) decreases as we move towards its end:** The field does become weaker and starts to fringe (spread out) near the ends. At the very end of a long solenoid, the field strength is approximately half of what it is deep inside.
- **(3) increases as we move towards its end:** This is incorrect.
- **(4) is the same at all points:** This statement is true for the region *well inside* a long solenoid, away from the effects of the ends. Given the options, this is the best description of the field characteristics within the main body of a long solenoid. If the question implies ”deep inside” or ”far from the ends”, then this is accurate.

Step 4: Choosing the best fit option While the field strength does change near the ends, the defining characteristic of the magnetic field *inside* a long solenoid (implying the main internal region) is its uniformity. Therefore, option (4) is the most appropriate answer in the context of typical idealizations taught at an introductory level.

Quick Tip

For a **long** solenoid, the magnetic field deep inside is remarkably **uniform** (same strength and direction everywhere inside, pointing along the axis). The field lines are straight, parallel, and evenly spaced inside. The field only weakens and spreads out near the very ends. Option (4) describes the ideal behavior well inside.

---

89. Voltmeter in a circuit is connected in :

(1) parallel

(2) series

(3) in any way

(4) none of these

Correct Answer: (1) parallel

Solution: Concept: A voltmeter is an instrument used to measure the electric potential difference (voltage) between two points in a circuit. How it's connected is crucial for a correct measurement and for the proper functioning of the circuit.

Step 1: What does a voltmeter measure? A voltmeter measures the difference in electrical potential (voltage drop) *across* a circuit component (like a resistor, bulb) or between any two points.

Step 2: How to connect a voltmeter to measure voltage ACROSS a component To measure the voltage across a component, the voltmeter must be connected so that its terminals are at the same two points as the component whose voltage is being measured. This means the voltmeter is connected **in parallel** with that component.

- One terminal of the voltmeter is connected to one end of the component.
- The other terminal of the voltmeter is connected to the other end of the component.

This way, the voltmeter experiences the same potential difference as the component.

Step 3: Why not in series? If a voltmeter were connected in series with a component, it would be part of the main current path. An ideal voltmeter has very high internal resistance (to draw minimal current and not disturb the circuit it's measuring). If placed in series, this high resistance would significantly reduce the current in the circuit, altering the very voltage it's trying to measure and affecting the circuit's operation. (An ammeter, which measures current, is connected in series and has very low resistance).

Step 4: Analyzing the options

- **(1) parallel:** Correct. A voltmeter is connected in parallel across the points where voltage is to be measured.
- **(2) series:** Incorrect. This is how an ammeter is connected.
- **(3) in any way:** Incorrect. The connection method is specific.
- **(4) none of these:** Incorrect, as parallel is the correct method.

Quick Tip

Remember the "P" in **P**otential difference (Voltage) and **P**arallel:

- **V**oltmeter measures **V**oltage (Potential Difference) and is connected in **P**arallel.
- **A**mmeter measures current (**A**mperes) and is connected in **S**eries. (No "P" here to help remember).

A voltmeter needs to "see" the voltage across a component, so it branches off in parallel.

---

90. Orested is the unit of :

(1) Magnetic field

(2) Electrical current

(3) Electrical field

(4) Magnetic current

Correct Answer: (1) Magnetic field

Solution: Concept: Different physical quantities have specific units. The oersted (Oe) is a unit associated with magnetism.

Step 1: Identifying the Oersted unit The oersted is a unit used in the CGS (centimeter-gram-second) system of units. It is named after Hans Christian Ørsted, who discovered the relationship between electricity and magnetism.

Step 2: What quantity does the Oersted measure? The oersted is a unit of **magnetic field strength**, specifically the magnetizing field strength, often denoted by H. In the CGS system, magnetic field strength H is measured in oersteds. The magnetic flux density B (another way to describe a magnetic field) is measured in gauss (G) in the CGS system. These are related by $B = \mu H$, where μ is the permeability of the medium. In vacuum, $\mu = 1$ in the CGS EMU system, so B (in gauss) and H (in oersteds) are numerically equal in vacuum, though they represent distinct physical concepts.

Step 3: SI unit for magnetic field strength In the SI system (International System of Units), the unit for magnetic field strength H is amperes per meter (A/m). The SI unit for magnetic flux density B is the tesla (T). 1 Oe $= (1000/4\pi)$ A/m ≈ 79.577 A/m. 1 G $= 10^{-4}$ T.

Step 4: Analyzing the options

- **(1) Magnetic field:** Correct. Oersted is a unit of magnetic field strength (H).
- **(2) Electrical current:** The SI unit is ampere (A).
- **(3) Electrical field:** The SI unit is volts per meter (V/m) or newtons per coulomb (N/C).
- **(4) Magnetic current:** This is not a standard physical quantity in the same way as electric current. The concept of magnetic monopole current is theoretical.

Therefore, oersted is a unit of magnetic field.

Quick Tip

Oersted (Oe) is a CGS unit for **magnetic field strength** (H). You might also encounter **gauss (G)**, another CGS unit for magnetic flux density (B). In the SI system, the corresponding units are:

- Magnetic field strength (H): Ampere per meter (A/m)
- Magnetic flux density (B): Tesla (T)

Hans Christian Ørsted was a pioneer in electromagnetism, so his name is fittingly associated with a magnetic unit.

Chemistry

91. The compound that can be purified by sublimation is :

(1) Ammonium Chloride

(2) Calcium Carbonate

(3) Sodium Carbonate

(4) Aluminium Chloride

Correct Answer: (1) Ammonium Chloride

Solution: Concept: Sublimation is a process where a substance transitions directly from the solid phase to the gas phase without passing through the liquid phase. This property can be used for purification if one component of a mixture sublimes while others do not.

Step 1: Understanding Sublimation for Purification If a mixture contains a sublimable substance (one that can sublime) and non-sublimable impurities, heating the mixture will cause the sublimable substance to turn into vapor. This vapor can then be cooled and condensed back into a solid in a separate location, leaving the impurities behind.

Step 2: Identifying substances that sublime Common substances that readily sublime at or near atmospheric pressure include:

- Iodine

- Naphthalene (mothballs)
- Camphor
- Dry ice (solid Carbon Dioxide)
- **Ammonium Chloride (NH_4Cl)**

Step 3: Analyzing the options

- **(1) Ammonium Chloride (NH_4Cl):** This is a well-known example of a compound that sublimes upon heating. It transitions directly from solid to gas and can be condensed back to solid.
- **(2) Calcium Carbonate ($CaCO_3$):** Also known as limestone or chalk. It does not sublime under normal heating conditions; instead, it decomposes at high temperatures (e.g., $> 800°C$) into calcium oxide and carbon dioxide.
- **(3) Sodium Carbonate (Na_2CO_3):** Also known as washing soda. It is a stable ionic solid that melts at a high temperature ($851°C$) but does not typically sublime.
- **(4) Aluminium Chloride ($AlCl_3$):** Anhydrous aluminium chloride is a covalent compound that sublimes at around $180°C$. This is also a substance that can be purified by sublimation. However, ammonium chloride is a more common textbook example for sublimation at a simpler level. Given that Ammonium Chloride is listed and is a classic example, it is the intended answer. If both were present and only one could be chosen, context or common knowledge emphasis would guide.

Step 4: Conclusion Among the given options, Ammonium Chloride is a classic example of a compound that can be purified by sublimation. Aluminium Chloride also sublimes, but Ammonium Chloride is a very standard example in introductory chemistry.

Quick Tip

Sublimation is solid $\rightarrow$ gas directly. Think of substances that "disappear" as solids when heated and reappear as solids when cooled, without melting. Common examples are ammonium chloride, camphor, naphthalene, and iodine. Ammonium chloride is a key example to remember for this process.

92. A gel toothpaste is a mixture of a :

(1) liquid in a solid

(2) solid in a gas

(3) liquid in liquid

(4) gas in solid

Correct Answer: (1) liquid in a solid

Solution: Concept: Gels are a type of colloid. A colloid is a mixture in which one substance of microscopically dispersed insoluble particles is suspended throughout another substance. The type of colloid is defined by the phases of the dispersed substance and the dispersion medium.

Step 1: Understanding what a gel is A gel is a colloidal system in which a liquid (the dispersed phase) is dispersed throughout a solid continuous network (the dispersion medium). The solid network gives the gel its structure and semi-solid, jelly-like consistency, trapping the liquid within it.

Step 2: Analyzing the options for colloidal systems

- **(1) liquid in a solid:** This describes a **gel**. The liquid particles are dispersed in a solid medium. Examples: gelatin, jelly, cheese, butter, gel toothpaste.
- **(2) solid in a gas:** This describes a **solid aerosol**. Examples: smoke, dust in the air.
- **(3) liquid in liquid:** This describes an **emulsion**. Examples: milk, mayonnaise.
- **(4) gas in solid:** This describes a **solid foam** or **solid sol**. Examples: pumice stone, styrofoam, bread.

Step 3: Applying to gel toothpaste Gel toothpaste has a characteristic jelly-like, semi-solid texture. This structure is formed by a solid network (gelling agents like hydrated silica or polymers) that entraps liquid components (like water, sorbitol, glycerin). Therefore, a gel toothpaste is a system where a liquid is dispersed in a solid.

Quick Tip

Gels are "jelly-like." They feel wet (due to the liquid) but hold their shape (due to the solid network). Think of it this way:

- Dispersed phase (what's spread out) = Liquid
- Dispersion medium (what it's spread in) = Solid

This makes a gel a "liquid in a solid" colloid.

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93. During summer, water kept in an earthern pot becomes cool because of the phenomenon of :

(1) diffusion

(2) transpiration

(3) osmosis

(4) evaporation

Correct Answer: (4) evaporation

Solution: Concept: The cooling effect observed in water stored in earthen pots (like a matka or surahi) is due to a specific phase change process that requires energy.

Step 1: Properties of an earthen pot Earthen pots are porous, meaning they have very tiny, microscopic pores or holes in their walls.

Step 2: The process occurring Water kept inside the earthen pot seeps out through these tiny pores to the outer surface of the pot. On the outer surface, this water comes into contact with the surrounding air. The water on the outer surface undergoes **evaporation**.
Evaporation is the process where a liquid changes into its vapor (gas) phase at a temperature below its boiling point.

Step 3: How evaporation causes cooling Evaporation is an endothermic process, meaning it requires energy (heat). The molecules of water that evaporate from the surface take this energy (latent heat of vaporization) from:

- The remaining water on the surface of the pot.

- The pot itself.
- Ultimately, from the water inside the pot.

As this heat energy is continuously removed from the system (pot and water inside) to fuel the evaporation, the temperature of the water inside the pot decreases, making it cool.

Step 4: Analyzing other options

- **(1) Diffusion:** The movement of particles from an area of higher concentration to lower concentration. While involved in vapor movement, it's not the primary cooling mechanism.
- **(2) Transpiration:** The process of water movement through a plant and its evaporation from aerial parts, such as leaves, stems and flowers. Not relevant to an earthen pot.
- **(3) Osmosis:** The movement of solvent (like water) across a semipermeable membrane from a region of low solute concentration to high solute concentration. Not the primary cooling mechanism here.

The cooling is directly caused by the energy absorbed during evaporation.

> **Quick Tip**
>
> Evaporation needs heat. When water evaporates from the surface of an earthen pot, it takes heat from the pot and the water inside. This removal of heat makes the remaining water cooler. It's similar to how sweating cools your body – the sweat evaporates, taking heat from your skin.

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94. Tincture of iodine has antiseptic properties. This solution is made by dissolving :

(1) iodine in potassium iodide

(2) iodine in acetone

(3) iodine in alcohol

(4) iodine in water

Correct Answer: (3) iodine in alcohol

Solution: Concept: Tincture of iodine is a common antiseptic, and its composition involves specific solutes and a solvent.

Step 1: Understanding "Tincture" In a pharmaceutical context, a "tincture" generally refers to a solution where alcohol (ethanol) or a mixture of alcohol and water is used as the solvent.

Step 2: Composition of Tincture of Iodine Tincture of iodine is an antiseptic solution primarily composed of:

- **Solute:** Iodine (I_2)
- **Solvent:** A mixture of ethanol (alcohol) and water. Sometimes, potassium iodide (KI) or sodium iodide (NaI) is also added to increase the solubility of iodine in the aqueous-alcoholic solution (iodine itself is not very soluble in plain water but dissolves better in iodide solutions due to the formation of triiodide ions, I_3^-).

However, the primary solvent characterising it as a "tincture" is alcohol. The question asks what iodine is dissolved in to make the solution.

Step 3: Analyzing the options

- **(1) iodine in potassium iodide:** While potassium iodide is often used to help dissolve iodine, potassium iodide itself is usually dissolved in water or an alcohol-water mixture. This option is incomplete as it doesn't specify the main solvent.
- **(2) iodine in acetone:** Acetone can dissolve iodine, but solutions in acetone are not typically referred to as "tincture of iodine."
- **(3) iodine in alcohol:** This aligns with the definition of a tincture. Tincture of iodine is iodine dissolved in alcohol (usually ethanol), often with some water and potentially an iodide salt.
- **(4) iodine in water:** Iodine has low solubility in pure water. Solutions of iodine in water (like Lugol's iodine) usually require potassium iodide to achieve a useful concentration.

Step 4: Identifying the best description Given the term "tincture," the primary solvent system involves alcohol. Therefore, dissolving "iodine in alcohol" is the most direct and defining characteristic. Standard tincture of iodine typically contains 2-7% elemental iodine, along with potassium iodide or sodium iodide, dissolved in a mixture of ethanol and water. The alcohol is a key component.

Quick Tip

The word ”tincture” usually implies that **alcohol** is the solvent. So, Tincture of Iodine = Iodine dissolved in Alcohol (often with water and an iodide salt like KI to help dissolve more iodine). The key part to remember for ”tincture” is the alcohol.

---

95. The Chemical symbol for tin is :

(1) Ti

(2) Sb

(3) Sn

(4) Te

Correct Answer: (3) Sn

Solution: Concept: Chemical symbols are abbreviations used in chemistry for chemical elements. Each element has a unique one- or two-letter symbol. Some symbols are derived from the element’s Latin name.

Step 1: Understanding Chemical Symbols Chemical symbols are internationally recognized codes for elements. The first letter is always capitalized, and if there’s a second letter, it’s lowercase.

Step 2: Identifying the symbol for Tin The chemical symbol for tin is **Sn**. This symbol comes from its Latin name, ”stannum.”

Step 3: Analyzing the options and identifying other elements

- **(1) Ti:** This is the chemical symbol for **Titanium**.
- **(2) Sb:** This is the chemical symbol for **Antimony** (from its Latin name ”stibium”).
- **(3) Sn:** This is the chemical symbol for **Tin** (from its Latin name ”stannum”).
- **(4) Te:** This is the chemical symbol for **Tellurium**.

Therefore, the correct chemical symbol for tin is Sn.

Quick Tip

Some chemical symbols are not obvious from their English names because they come from Latin names. For Tin, remember:

- English name: Tin
- Latin name: Stannum
- Symbol: **Sn**

Other examples: Lead (Pb, from Plumbum), Gold (Au, from Aurum), Silver (Ag, from Argentum).

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96. The names of the elements present in quick lime are :

(i) Ca (ii) H (iii) O (iv) C

(1) (i), (ii) and (iii)

(2) (i), (iii) and (iv)

(3) (i) and (iii)

(4) (i), (ii), (iii) and (iv)

Correct Answer: (3) (i) and (iii)

Solution: Concept: Quick lime is a common chemical compound with a specific chemical formula. Knowing the formula allows us to identify the constituent elements.

Step 1: Identify the chemical name and formula of Quick Lime Quick lime is the common name for **Calcium Oxide**. The chemical formula for Calcium Oxide is **CaO**.

Step 2: Identify the elements from the chemical formula The formula CaO indicates that one molecule of quick lime contains:

- One atom of Calcium (symbol: Ca)
- One atom of Oxygen (symbol: O)

Step 3: Match the elements with the given list The given list of elements is: (i) Ca (Calcium) (ii) H (Hydrogen) (iii) O (Oxygen) (iv) C (Carbon)

From our analysis, quick lime (CaO) contains Calcium (Ca) and Oxygen (O). So, elements (i) Ca and (iii) O are present in quick lime.

Step 4: Choose the correct option Option (1) (i), (ii) and (iii) includes Hydrogen (H), which is not in CaO. Option (2) (i), (iii) and (iv) includes Carbon (C), which is not in CaO. Option (3) (i) and (iii) correctly lists Calcium (Ca) and Oxygen (O). Option (4) (i), (ii), (iii) and (iv) includes Hydrogen (H) and Carbon (C), which are not in CaO.

Therefore, the elements present in quick lime are Calcium (Ca) and Oxygen (O).

Quick Tip

Key common names and formulas to remember:

- Quick Lime = Calcium Oxide = **CaO** (Elements: Ca, O)
- Slaked Lime = Calcium Hydroxide = $Ca(OH)_2$ (Elements: Ca, O, H)
- Limestone/Chalk = Calcium Carbonate = $CaCO_3$ (Elements: Ca, C, O)

For this question, quick lime is CaO.

---

97. Which of the following has maximum number of atoms ?

(1) 18g of H_2O

(2) 18g of O_2

(3) 18g of CO_2

(4) 18g of CH_4

Correct Answer: (4) 18g of CH_4

Solution: Concept: To find the total number of atoms, we first need to calculate the number of moles of each substance, then the number of molecules, and finally the total number of atoms by considering the number of atoms per molecule. Number of moles (n) = Mass (m) / Molar mass (M). Number of molecules = Number of moles $\times$ Avogadro's number

($N_A \approx 6.022 \times 10^{23}$ mol^{-1}). Total number of atoms = Number of molecules $\times$ Number of atoms per molecule.

Since the mass (18g) and Avogadro's number are constant for all options, the number of atoms will be proportional to (Number of atoms per molecule) / (Molar mass). A simpler approach for comparison is to calculate moles, then total moles of atoms.

Step 1: Calculate Molar Masses (M)

- H_2O: 2×1 (for H) $+ 16$ (for O) $= 18$ g/mol
- O_2: 2×16 (for O) $= 32$ g/mol
- CO_2: 12 (for C) $+ 2 \times 16$ (for O) $= 12 + 32 = 44$ g/mol
- CH_4: 12 (for C) $+ 4 \times 1$ (for H) $= 12 + 4 = 16$ g/mol

(Atomic masses: H=1, C=12, O=16 g/mol approximately)

Step 2: Calculate number of moles (n) for 18g of each substance

- H_2O: $n = 18\text{g}/18$ g/mol $= 1$ mol
- O_2: $n = 18\text{g}/32$ g/mol $= 0.5625$ mol
- CO_2: $n = 18\text{g}/44$ g/mol ≈ 0.409 mol
- CH_4: $n = 18\text{g}/16$ g/mol $= 1.125$ mol

Step 3: Calculate total number of moles of atoms (Total moles of atoms = moles of substance $\times$ number of atoms in one molecule)

- H_2O: 1 molecule has $2(H) + 1(O) = 3$ atoms. Total moles of atoms = 1 mol $\times 3 = 3$ moles of atoms.
- O_2: 1 molecule has $2(O)$ atoms. Total moles of atoms = 0.5625 mol $\times 2 = 1.125$ moles of atoms.
- CO_2: 1 molecule has $1(C) + 2(O) = 3$ atoms. Total moles of atoms = 0.409 mol $\times 3 \approx 1.227$ moles of atoms.
- CH_4: 1 molecule has $1(C) + 4(H) = 5$ atoms. Total moles of atoms = 1.125 mol $\times 5 = 5.625$ moles of atoms.

Step 4: Compare the total number of moles of atoms Comparing the values:

- H_2O: 3 moles of atoms
- O_2: 1.125 moles of atoms
- CO_2: ≈ 1.227 moles of atoms
- CH_4: 5.625 moles of atoms

18g of CH_4 has the maximum number of moles of atoms, and thus the maximum number of atoms.

Quick Tip

For a fixed mass of different substances, the substance with the: 1. Lowest molar mass will have the highest number of moles. 2. The highest number of atoms per molecule will contribute more atoms per mole. Combine these: Calculate (atoms per molecule / molar mass). The substance with the largest value of this ratio for a given mass will have the most atoms. For CH_4: Molar mass = 16 g/mol, Atoms per molecule = 5. Ratio = 5/16 = 0.3125. For H_2O: Molar mass = 18 g/mol, Atoms per molecule = 3. Ratio = 3/18 $\approx$ 0.1667. CH_4 has more moles of molecules (1.125 vs 1 for H_2O) AND more atoms per molecule (5 vs 3 for H_2O).

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98. The number of electrons in an element X is 6 and the number of neutron is 8. Which of the following is the correct representation of the element ?

(1) ${}^{8}_{6}\text{X}$

(2) ${}^{14}_{6}\text{X}$

(3) ${}^{6}_{8}\text{X}$

(4) ${}^{14}_{8}\text{X}$

Correct Answer: (2) ${}^{14}_{6}\text{X}$

Solution: Concept: The standard notation for representing an element (or isotope) is ${}^{A}_{Z}\text{X}$, where:

- X is the chemical symbol of the element.
- Z is the Atomic Number (number of protons).
- A is the Mass Number (number of protons + number of neutrons).

Step 1: Given information

- Number of electrons in element X = 6.
- Number of neutrons in element X = 8.

Step 2: Determine the Atomic Number (Z) For a neutral atom (which an "element" generally implies unless specified as an ion), the number of electrons is equal to the number of protons. Number of protons = Number of electrons = 6. The Atomic Number (Z) is defined as the number of protons. So, $Z = 6$. (An element with atomic number 6 is Carbon, C, but the symbol X is used here).

Step 3: Determine the Mass Number (A) The Mass Number (A) is the sum of the number of protons and the number of neutrons. Number of protons = 6 Number of neutrons = 8 So, $A = \text{Number of protons} + \text{Number of neutrons} = 6 + 8 = 14$.

Step 4: Write the correct representation Using the notation ${}^{A}_{Z}\text{X}$: Substitute $A = 14$ and $Z = 6$. The representation is ${}^{14}_{6}\text{X}$.

Step 5: Compare with the options

- (1) ${}^{8}_{6}\text{X}$: Mass number is incorrect.
- (2) ${}^{14}_{6}\text{X}$: This matches our derived representation.
- (3) ${}^{6}_{8}\text{X}$: Atomic number and mass number positions/values are incorrect.
- (4) ${}^{14}_{8}\text{X}$: Atomic number is incorrect.

Therefore, the correct representation is ${}^{14}_{6}\text{X}$.

Quick Tip

Remember the notation ${}^{A}_{Z}\text{X}$:

- Z (bottom number) = Atomic Number = Number of Protons. (For neutral atom, also = Number of Electrons).
- A (top number) = Mass Number = Number of Protons + Number of Neutrons.

Given electrons = 6 $\rightarrow$ protons (Z) = 6. Given neutrons = 8. Mass number (A) = protons + neutrons = $6 + 8 = 14$. So, ${}^{14}_{6}\text{X}$.

---

99. The ion of an element has 3 positive charges. Mass number of the atom is 27 and the number of neutrons is 14. What is the number of electrons in the ion ?

(1) 13

(2) 10

(3) 14

(4) 16

Correct Answer: (2) 10

Solution: Concept:

- Mass Number (A) = Number of protons (p) + Number of neutrons (n).
- Atomic Number (Z) = Number of protons (p). For a neutral atom, Z also equals the number of electrons.
- For an ion, the number of electrons changes. A positive charge means electrons are lost. A negative charge means electrons are gained.

Step 1: Given information

- Charge of the ion = +3 (3 positive charges).
- Mass number of the atom (A) = 27.
- Number of neutrons (n) = 14.

We need to find the number of electrons in the ion.

Step 2: Find the number of protons (p) We know $A = p + n$. So, $p = A - n$.

$p = 27 - 14 = 13$. The number of protons in the atom (and also in the ion, as protons don't change when an ion forms) is 13. This means the atomic number (Z) of the element is 13. (This element is Aluminium, Al).

Step 3: Determine the number of electrons in the neutral atom For a neutral atom, the number of electrons is equal to the number of protons. Number of electrons in a neutral atom = 13.

Step 4: Determine the number of electrons in the ion The ion has 3 positive charges (+3). A positive charge indicates that the atom has lost electrons. A +3 charge means the atom has lost 3 electrons. Number of electrons in the ion = Number of electrons in neutral atom - Number of electrons lost Number of electrons in the ion = $13 - 3 = 10$.

Therefore, the number of electrons in the ion is 10. This matches option (2).

Quick Tip

1. Find protons: Protons = Mass Number - Neutrons = $27 - 14 = 13$ protons. 2. In a neutral atom, electrons = protons = 13 electrons. 3. Ion has +3 charge: This means it has LOST 3 electrons. 4. Electrons in ion = Electrons in neutral atom - 3 = $13 - 3 = 10$ electrons.

---

100. Isotopes of an element have :

(1) The same physical properties

(2) different chemical properties

(3) different number of neutrons

(4) different atomic number

Correct Answer: (3) different number of neutrons

Solution: Concept: Isotopes are variants of a particular chemical element. Understanding their definition is key.

Step 1: Defining Isotopes Isotopes of an element are atoms that have:

- The **same number of protons** (which means they have the same atomic number, Z). Since they have the same number of protons, they are atoms of the same element.
- A **different number of neutrons**.
- Because they have a different number of neutrons (but the same number of protons), they will have a **different mass number** ($A = Z +$ neutrons).

Step 2: Analyzing the properties based on the definition

- **Chemical Properties:** Chemical properties are primarily determined by the number of electrons, specifically the valence electrons. Since isotopes of an element have the same number of protons, their neutral atoms have the same number of electrons and the same electron configuration. Therefore, isotopes of an element generally have **very similar (nearly identical) chemical properties**.
- **Physical Properties:** Some physical properties, particularly those dependent on mass (like density, diffusion rate, melting/boiling points to a very slight extent), can differ slightly between isotopes due to the difference in mass number. So, they do not have exactly the same physical properties, though the differences might be small.
- **Atomic Number:** By definition, isotopes of the same element have the **same atomic number** (same number of protons).
- **Number of Neutrons:** By definition, isotopes of the same element have a **different number of neutrons**.

Step 3: Evaluating the options

- **(1) The same physical properties:** Incorrect. Physical properties dependent on mass can differ slightly.
- **(2) different chemical properties:** Incorrect. They have nearly identical chemical properties because they have the same electron configuration.
- **(3) different number of neutrons: Correct**. This is the defining characteristic of isotopes of an element.

- **(4) different atomic number:** Incorrect. They have the same atomic number (same number of protons).

Therefore, isotopes of an element have a different number of neutrons.

Quick Tip

Isotopes are like siblings of the same element family:

- **Same** number of Protons (Z) $\rightarrow$ Same element, Same Atomic Number, Similar Chemical Properties.
- **Different** number of Neutrons (N) $\rightarrow$ Different Mass Number ($A = Z + N$), Slightly different mass-dependent Physical Properties.

Example: Carbon-12 ($^{12}_{6}C$: 6p, 6n) and Carbon-14 ($^{14}_{6}C$: 6p, 8n) are isotopes of carbon.

---

101. The most abundant gas found in the earth atmosphere is :

(1) CO_2

(2) O_2

(3) CH_4

(4) N_2

Correct Answer: (4) N_2

Solution: Concept: Earth's atmosphere is a mixture of several gases. The relative amounts of these gases determine which one is the most abundant.

Step 1: Composition of Earth's Dry Atmosphere (Approximate Percentages by Volume)

- **Nitrogen (N_2):** About 78.08%
- **Oxygen (O_2):** About 20.95%
- **Argon (Ar):** About 0.93%
- **Carbon Dioxide (CO_2):** About 0.04% (variable, but this is a typical current value)
- Other gases (Neon, Helium, Methane (CH_4), Krypton, Hydrogen, etc.): Trace amounts, collectively less than 0.01%.

Water vapor is also a significant component of the atmosphere, but its concentration is highly variable (0-4%). The question usually refers to the composition of dry air.

Step 2: Identifying the most abundant gas From the percentages, Nitrogen (N_2) is clearly the most abundant gas, making up roughly 78

Step 3: Evaluating the options

- **(1) CO_2 (Carbon Dioxide):** Present in a small percentage ($\sim 0.04\%$).
- **(2) O_2 (Oxygen):** Second most abundant ($\sim 21\%$).
- **(3) CH_4 (Methane):** Present in trace amounts (parts per million).
- **(4) N_2 (Nitrogen):** Most abundant ($\sim 78\%$).

Therefore, Nitrogen (N_2) is the most abundant gas in Earth's atmosphere.

Quick Tip

Remember the top two gases in Earth's atmosphere: 1. **Nitrogen (N_2)** $\approx$ 782. **Oxygen (O_2)** $\approx$ 21These two make up about 99

---

102. Select from the following a set of three metals which are found in free state :

(1) Al, Cu, Ag

(2) Au, Fe, Ag

(3) Cu, Au, Fe

(4) Ag, Au, Pt

Correct Answer: (4) Ag, Au, Pt

Solution: Concept: Metals are found in the Earth's crust either in their native (free) state or in combined states as minerals (ores). The tendency to be found in a free state depends on their chemical reactivity. Less reactive metals are more likely to be found in their free state.

Step 1: Understanding "Free State" "Free state" or "native state" means the metal exists in its elemental form, uncombined with other elements. This is characteristic of metals that are low in the reactivity series (also known as the activity series).

Step 2: Reactivity of Metals

- **Highly Reactive Metals** (e.g., K, Na, Ca, Mg, Al): Found only in combined states (ores).
- **Moderately Reactive Metals** (e.g., Zn, Fe, Pb, Cu): Mostly found in combined states, but copper (Cu) can sometimes be found in its native state.
- **Least Reactive Metals** (e.g., Ag, Au, Pt): These are often called "noble metals" due to their low reactivity. They are commonly found in their free state.

Step 3: Analyzing the options Let's check the reactivity of the metals in each set:

- **(1) Al, Cu, Ag:**
  - Aluminium (Al): Highly reactive, found as bauxite ore ($Al_2O_3 \cdot xH_2O$). Not in free state.
  - Copper (Cu): Moderately reactive, mostly as ores (e.g., chalcopyrite), but native copper exists.
  - Silver (Ag): Low reactivity, often found in free state.

  This set is not entirely free state metals due to Al.
- **(2) Au, Fe, Ag:**
  - Gold (Au): Very low reactivity, found in free state.
  - Iron (Fe): Moderately reactive, found as ores (e.g., hematite, magnetite). Not typically in free state (except meteoritic iron).
  - Silver (Ag): Low reactivity, often found in free state.

  This set is not entirely free state metals due to Fe.
- **(3) Cu, Au, Fe:**
  - Copper (Cu): Can be found in free state.
  - Gold (Au): Found in free state.
  - Iron (Fe): Not typically in free state.

  This set is not entirely free state metals due to Fe.
- **(4) Ag, Au, Pt:**

– Silver (Ag): Low reactivity, found in free state.

– Gold (Au): Very low reactivity, found in free state.

– Platinum (Pt): Very low reactivity, found in free state.

All three metals in this set are known to be found in their free (native) state due to their low chemical reactivity.

Therefore, the set Ag, Au, Pt consists of three metals commonly found in their free state.

Quick Tip

Metals at the bottom of the reactivity series are unreactive and often found in their "free" or "native" state. These include:

- Gold (Au)
- Silver (Ag)
- Platinum (Pt)
- Copper (Cu) can also be found in native form, though it's more reactive than Au, Ag, Pt.

Look for sets containing these least reactive metals.

---

103. The gas which may cause explosion in coal mines is :

(1) Methane

(2) Ethane

(3) Nitrogen

(4) Oxygen

Correct Answer: (1) Methane

Solution: Concept: Coal mines can contain flammable gases trapped within the coal seams. When these gases mix with air in certain proportions, they can form explosive mixtures.

Step 1: Gases found in Coal Mines Coal formation processes can trap various gases. The most significant flammable gas found in coal mines is **Methane (CH_4)**. Methane is the main component of natural gas. In coal mines, methane is often referred to as ”firedamp.” Another major hazard is coal dust itself, which can also be explosive when suspended in air in sufficient concentrations.

Step 2: Conditions for Explosion For a gas to cause an explosion, it must be:

1. Flammable (able to burn).
2. Mixed with an oxidizer (usually oxygen from the air) in proportions that fall within its explosive or flammability limits.
3. Ignited by a source of ignition (like a spark, open flame, or hot surface).

Step 3: Analyzing the options

- **(1) Methane (CH_4):** Methane is highly flammable and forms explosive mixtures with air. It is a well-known and primary cause of explosions in coal mines (firedamp explosions).
- **(2) Ethane (C_2H_6):** Ethane is also a flammable gas and can be present with methane, but methane is typically the predominant flammable gas associated with coal mine explosions.
- **(3) Nitrogen (N_2):** Nitrogen is an inert gas; it does not burn and is not explosive. It's the main component of air.
- **(4) Oxygen (O_2):** Oxygen is necessary for combustion (it's an oxidizer), but oxygen itself is not flammable or explosive. It supports the combustion of flammable substances. The image indicates this option was circled, which is incorrect as oxygen supports combustion but isn't the fuel itself for an explosion in this context.

Step 4: Identifying the primary cause of explosions Methane (CH_4) is the gas most notorious for causing explosions in coal mines. When mixed with air (which contains oxygen) in concentrations roughly between 5% and 15% methane, it can be ignited and explode violently.

Quick Tip

The primary flammable gas responsible for explosions in coal mines is **Methane (CH_4)**, also known as ”firedamp.” Oxygen is needed for the explosion (as an oxidizer), but it is not the fuel. Methane is the fuel. Coal dust explosions are another significant hazard.

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104. Which of the following is the most electronegative element in the periodic table ?

(1) Oxygen

(2) Nitrogen

(3) Flourine

(4) Chlorine

Correct Answer: (3) Fluorine

Solution: Concept: Electronegativity is a measure of the tendency of an atom to attract a bonding pair of electrons towards itself when it is part of a chemical bond.

Step 1: Trends in Electronegativity in the Periodic Table

- **Across a Period (Left to Right):** Electronegativity generally increases. This is because the number of protons (nuclear charge) increases, pulling the bonding electrons more strongly, while the electrons are added to the same principal energy level.
- **Down a Group (Top to Bottom):** Electronegativity generally decreases. This is because the bonding electrons are in higher energy levels, further from the nucleus, and are shielded by more inner electron shells, reducing the nucleus's attraction for them.

As a result of these trends, the most electronegative elements are found in the upper right-hand corner of the periodic table (excluding noble gases, which generally don't form bonds readily or have conventionally defined electronegativity values in the same scale).

Step 2: Identifying the Most Electronegative Element Fluorine (F) is located at the top of Group 17 (Halogens) and in the second period. Due to its position, it has the highest electronegativity value of all elements. On the Pauling scale (a common scale for electronegativity), fluorine is assigned a value of 3.98 (often rounded to 4.0).

Step 3: Comparing the electronegativity of the given options Approximate Pauling electronegativity values:

- **(1) Oxygen (O):** ≈ 3.44 (Second most electronegative)
- **(2) Nitrogen (N):** ≈ 3.04 (Third/Fourth most electronegative, close to Chlorine)
- **(3) Fluorine (F):** ≈ 3.98 (Most electronegative)
- **(4) Chlorine (Cl):** ≈ 3.16 (Third/Fourth most electronegative, close to Nitrogen)

The order of electronegativity for these common highly electronegative elements is: F ¿ O ¿ Cl ≈ N.

Therefore, Fluorine is the most electronegative element in the periodic table. (Note: The spelling in the option is "Flourine", the correct spelling is "Fluorine").

Quick Tip

Electronegativity increases towards the top-right of the periodic table. **F**luorine (F) is the king of electronegativity – it's the most "electron-greedy" element. Remember the top few: **F ¿ O ¿ N ≈ Cl**. (Often remembered by the mnemonic FONCl - "Foncal").

---

105. Natural Rubber obtained from Rubber tree is basically a polymer of :

(1) Ethylene

(2) Propylene

(3) Acetylene

(4) Isoprene

Correct Answer: (4) Isoprene

Solution: Concept: Natural rubber is a polymer, which means it is a large molecule made up of repeating structural units called monomers.

Step 1: What is Natural Rubber? Natural rubber is primarily obtained from the latex of the Hevea brasiliensis tree (rubber tree). Latex is a milky colloidal suspension. When processed, this latex yields solid natural rubber.

Step 2: The Monomer of Natural Rubber The repeating monomer unit in natural rubber is **isoprene**. Isoprene is a common name for 2-methyl-1,3-butadiene. Its chemical formula is $CH_2 = C(CH_3)CH = CH_2$. In natural rubber, isoprene units are polymerized (linked together) primarily in a *cis*-1,4 configuration, forming polyisoprene. This specific configuration gives natural rubber its characteristic elasticity.

Step 3: Analyzing the options

- **(1) Ethylene ($CH_2 = CH_2$):** Polymerizes to form polyethylene, a common plastic. Not the monomer of natural rubber.
- **(2) Propylene ($CH_3CH = CH_2$):** Polymerizes to form polypropylene, another common plastic. Not the monomer of natural rubber.
- **(3) Acetylene ($HC \equiv CH$):** Can polymerize to form polyacetylene, a conductive polymer. Not the monomer of natural rubber.
- **(4) Isoprene (2-methyl-1,3-butadiene):** This is the correct monomer unit of natural rubber.

Therefore, natural rubber is basically a polymer of isoprene.

Quick Tip

Natural Rubber = Polymer of **Isoprene**. The scientific name for isoprene is 2-methyl-1,3-butadiene. The polymer is called polyisoprene. Think: "Natural rubber $\rightarrow$ Isoprene."

---

106. Metal, which will displace hydrogen from dilute HCl is :

(1) Ag

(2) Au

(3) Mg

(4) Cu

Correct Answer: (3) Mg

Solution: Concept: The ability of a metal to displace hydrogen from an acid (like dilute HCl) depends on its position in the reactivity series (also known as the activity series of metals). Metals above hydrogen in the reactivity series are more reactive than hydrogen and can displace it from dilute non-oxidizing acids.

Step 1: Understanding the Reactivity Series The reactivity series arranges metals in order of their decreasing reactivity. A simplified portion relevant here is: K ¿ Na ¿ Ca ¿ **Mg** ¿ Al ¿ Zn ¿ Fe ¿ Sn ¿ Pb ¿ **(H)** ¿ **Cu** ¿ Hg ¿ **Ag** ¿ Pt ¿ **Au** (Highly reactive → Least reactive) Metals *above* Hydrogen (H) in this series will displace hydrogen from dilute acids like HCl or H_2SO_4. Metals *below* Hydrogen will not displace hydrogen from these acids.

Step 2: Analyzing the options based on their position relative to Hydrogen

- **(1) Ag (Silver):** Silver is *below* hydrogen in the reactivity series. Therefore, it cannot displace hydrogen from dilute HCl. (Reaction: $Ag + HCl \rightarrow$ No reaction)
- **(2) Au (Gold):** Gold is far *below* hydrogen in the reactivity series (it's a very unreactive noble metal). It cannot displace hydrogen from dilute HCl. (Reaction: $Au + HCl \rightarrow$ No reaction)
- **(3) Mg (Magnesium):** Magnesium is well *above* hydrogen in the reactivity series. Therefore, it is more reactive than hydrogen and will displace it from dilute HCl. Reaction: $Mg(s) + 2HCl(aq) \rightarrow MgCl_2(aq) + H_2(g)$
- **(4) Cu (Copper):** Copper is *below* hydrogen in the reactivity series. Therefore, it cannot displace hydrogen from dilute HCl. (Reaction: $Cu + HCl \rightarrow$ No reaction)

Step 3: Identifying the correct metal Magnesium (Mg) is the only metal among the options that is more reactive than hydrogen and thus will displace hydrogen from dilute HCl.

Quick Tip

To displace hydrogen from dilute acids (like HCl), a metal must be **more reactive than hydrogen**. This means the metal must be **above hydrogen (H)** in the reactivity series.

- Mg: Above H (Yes, displaces)
- Ag, Au, Cu: Below H (No, do not displace)

107. Which was the first organic compound synthesized in the laboratory ?

(1) Methane

(2) Urea

(3) Thiourea

(4) Ethanol

Correct Answer: (2) Urea

Solution: Concept: Historically, organic compounds were believed to be produced only by living organisms due to a "vital force." The laboratory synthesis of an organic compound from inorganic starting materials was a landmark event that disproved this theory.

Step 1: The Vital Force Theory In the early 19th century, the prevailing "vital force theory" (vitalism) suggested that organic compounds could only be formed by living organisms because they possessed a special "vital force."

Step 2: Friedrich Wöhler's Experiment (1828) In 1828, the German chemist Friedrich Wöhler made a groundbreaking discovery. He was attempting to synthesize ammonium cyanate (NH_4CNO), an inorganic compound. He did this by reacting silver cyanate ($AgCNO$) with ammonium chloride (NH_4Cl), or potassium cyanate with ammonium sulfate. Upon heating ammonium cyanate, he unexpectedly obtained crystals of **urea** ($CO(NH_2)_2$).

$$\underset{\text{Ammonium Cyanate (inorganic)}}{NH_4CNO} \xrightarrow{\text{Heat}} \underset{\text{Urea (organic)}}{CO(NH_2)_2}$$

Urea was a known organic compound, found in urine. Wöhler's synthesis of urea from an inorganic starting material (ammonium cyanate was considered inorganic) was the first time an organic compound was synthesized in the laboratory without the involvement of living organisms.

Step 3: Significance of Wöhler's Synthesis This experiment was a major blow to the vital force theory and paved the way for the development of modern organic chemistry, which is based on the understanding that the same chemical principles govern both organic and inorganic compounds.

Step 4: Analyzing the options

- **(1) Methane (CH_4):** While a simple organic compound, it was not the first synthesized in this landmark context.
- **(2) Urea ($CO(NH_2)_2$):** This is the correct answer, synthesized by Wöhler in 1828.
- **(3) Thiourea ($CS(NH_2)_2$):** A derivative of urea, but not the first synthesized.
- **(4) Ethanol (CH_3CH_2OH):** Known through fermentation for millennia, but its synthesis from purely inorganic precursors in a lab was not the first breakthrough against vitalism.

Quick Tip

The synthesis of **Urea** by Friedrich Wöhler in 1828 is a famous event in chemistry history. He made an organic compound (urea) from an inorganic compound (ammonium cyanate). This helped disprove the "vital force" theory, which said only living things could make organic compounds.

---

108. Which of the following is a member of the Halogen family ?

(1) Cl

(2) Ca

(3) Cu

(4) Cr

Correct Answer: (1) Cl

Solution: Concept: The Halogen family is a specific group of elements in the periodic table known for their characteristic properties.

Step 1: Identifying the Halogen Group The Halogens are the elements found in **Group 17** (or VIIA) of the periodic table. The members of the Halogen family are:

- Fluorine (F)
- **Chlorine (Cl)**
- Bromine (Br)

- Iodine (I)
- Astatine (At) - radioactive
- Tennessine (Ts) - synthetic, radioactive

These elements are highly reactive nonmetals and readily form salts by reacting with metals (the name ”halogen” means ”salt-former”).

Step 2: Analyzing the options

- **(1) Cl (Chlorine):** Chlorine is a member of Group 17, the Halogen family.
- **(2) Ca (Calcium):** Calcium is an alkaline earth metal, found in Group 2 of the periodic table.
- **(3) Cu (Copper):** Copper is a transition metal, found in Group 11 of the periodic table.
- **(4) Cr (Chromium):** Chromium is a transition metal, found in Group 6 of the periodic table.

Step 3: Identifying the Halogen From the list, Chlorine (Cl) is the only element that belongs to the Halogen family.

> Quick Tip
>
> The Halogen family is Group 17 of the periodic table. Remember the main members: **F**luorine, **Cl**orine, **Br**omine, **I**odine, **A**statine. (Mnemonic: ”**F**unny **Cl**owns **Br**ing **I**ncredible **A**crobats”). Look for one of these elements in the options.

---

109. The metal most commonly used for making filament of an electric bulb is :

(1) Tungsten

(2) Copper

(3) Silver

(4) Aluminium

Correct Answer: (1) Tungsten

Solution: Concept: The filament of an incandescent electric bulb needs to have specific properties to function effectively: it must glow brightly when heated by electric current and withstand very high temperatures without melting or quickly degrading.

Step 1: Desired Properties of a Bulb Filament Material

- **High Melting Point:** The filament heats up to incandescence (glowing hot, typically over $2000°C$). The material must not melt at these operating temperatures.
- **High Resistivity:** A higher resistivity means that for a given current, more heat ($P = I^2R$) is generated in a reasonably sized filament. If the resistivity were too low, a very long and thin wire would be needed.
- **Ductility:** The ability to be drawn into thin wires.
- **Low Vapor Pressure at High Temperatures:** The material should not evaporate quickly at operating temperatures, as this would thin the filament and shorten its life, as well as blacken the bulb.
- **Sufficient Mechanical Strength at High Temperatures.**

Step 2: Analyzing the options

- **(1) Tungsten (W):**
  - **Extremely High Melting Point:** Approximately $3422°C$, the highest of all metals. This is its most crucial property for use as a filament.
  - High resistivity (though lower than some alloys, it's suitable).
  - Good ductility.
  - Relatively low vapor pressure at high temperatures.

  Tungsten meets these requirements exceptionally well and is the standard material for incandescent bulb filaments.
- **(2) Copper (Cu):** Melting point is relatively low ($1085°C$). It would melt long before reaching incandescent temperatures. It also has very low resistivity, making it unsuitable for a compact filament.

- **(3) Silver (Ag):** Melting point is low ($962°C$). Would melt easily. Excellent conductor (low resistivity), also not ideal for this purpose.
- **(4) Aluminium (Al):** Melting point is low ($660°C$). Would melt very easily. Good conductor (low resistivity).

Step 3: Identifying the most suitable metal Due to its exceptionally high melting point and other favorable properties at high temperatures, **Tungsten** is the metal most commonly used for making the filament of an electric incandescent bulb.

> **Quick Tip**
>
> A bulb filament needs to get extremely hot without melting. **Tungsten (W)** has the highest melting point of all metals ($\sim 3422°C$). This makes it perfect for incandescent bulb filaments, which operate at very high temperatures to produce light.

---

110. Vinegar, used in kitchen is a dilute solution of :

(1) Oxalic Acid

(2) Citric Acid

(3) Benzoic Acid

(4) Ethanoic Acid

Correct Answer: (4) Ethanoic Acid

Solution: Concept: Vinegar is a common household acidic liquid used in cooking and cleaning, produced through the fermentation of ethanol by acetic acid bacteria.

Step 1: What is Vinegar? Vinegar is essentially a dilute aqueous solution of **acetic acid**. It also contains other trace compounds which may include flavorings. The characteristic sour taste and pungent smell of vinegar are primarily due to acetic acid.

Step 2: Chemical Name for Acetic Acid The IUPAC (systematic) name for acetic acid is **Ethanoic Acid**. Its chemical formula is CH_3COOH.

Step 3: Analyzing the options

- **(1) Oxalic Acid ($(COOH)_2$):** A dicarboxylic acid found in some vegetables (like rhubarb, spinach). It is toxic in high concentrations and not the main component of vinegar.
- **(2) Citric Acid ($C_6H_8O_7$):** A weak organic acid found naturally in citrus fruits (like lemons, oranges). It gives them their tart flavor. Not the main component of vinegar.
- **(3) Benzoic Acid (C_6H_5COOH):** An aromatic carboxylic acid, sometimes used as a food preservative. Not the main component of vinegar.
- **(4) Ethanoic Acid (CH_3COOH):** This is the systematic name for acetic acid, which is the primary acidic component of vinegar. Commercial vinegar typically contains about 4-8% ethanoic acid by volume.

Therefore, vinegar is a dilute solution of ethanoic acid (acetic acid).

Quick Tip

Vinegar's sour taste comes from an acid. The acid in vinegar is **Acetic Acid**. The systematic chemical name for acetic acid is **Ethanoic Acid** (CH_3COOH). So, Vinegar = Dilute Ethanoic Acid solution.

---

111. When water is poured over sodium peroxide, which colourless gas is produced ?

(1) Dihydrogen

(2) Ozone

(3) Dioxygen

(4) Dinitrogen

Correct Answer: (3) Dioxygen

Solution: Concept: Sodium peroxide (Na_2O_2) is a compound that reacts with water to produce specific products, including a gas.

Step 1: The Chemical Reaction Sodium peroxide reacts with water according to the following balanced chemical equation:

$$2Na_2O_2(s) + 2H_2O(l) \rightarrow 4NaOH(aq) + O_2(g)$$

This reaction produces sodium hydroxide (NaOH), which dissolves in the water to form an aqueous solution, and oxygen gas (O_2).

Step 2: Identifying the Gaseous Product The gaseous product formed in this reaction is oxygen (O_2). Oxygen gas is a colourless and odourless gas under standard conditions.

Step 3: Analyzing the options

- **(1) Dihydrogen (H_2):** Hydrogen gas. Not the primary gaseous product from the reaction of sodium peroxide with water. (Highly reactive metals like sodium react with water to produce hydrogen, but sodium peroxide reacts differently).
- **(2) Ozone (O_3):** An allotrope of oxygen. Not produced in this reaction.
- **(3) Dioxygen (O_2):** This is the chemical name for molecular oxygen, which is the gas produced.
- **(4) Dinitrogen (N_2):** Nitrogen gas. Not produced as there is no nitrogen in the reactants.

The colourless gas produced is dioxygen (oxygen).

Quick Tip

The reaction is: Sodium Peroxide + Water $\rightarrow$ Sodium Hydroxide + Oxygen.

$$2Na_2O_2 + 2H_2O \rightarrow 4NaOH + O_2$$

The gas produced is Oxygen (O_2), which is also called "Dioxygen". This reaction is sometimes used as a source of oxygen.

---

112. Sodium metal is usually stored under :

(1) Water

(2) Kerosene Oil

(3) Alcohol

(4) Hydrogen

Correct Answer: (2) Kerosene Oil

Solution: Concept: Sodium (Na) is a highly reactive alkali metal. Its storage method must prevent it from reacting with components of the air or other common substances.

Step 1: Reactivity of Sodium Sodium is very reactive. It readily reacts with:

- **Oxygen in the air:** It tarnishes rapidly in air, forming sodium oxide (Na_2O) and if moisture is present, sodium hydroxide (NaOH).

$$4Na(s) + O_2(g) \rightarrow 2Na_2O(s)$$

- **Moisture (water vapor) in the air:** It reacts with water to form sodium hydroxide and hydrogen gas. This reaction is exothermic (produces heat).

$$2Na(s) + 2H_2O(l) \rightarrow 2NaOH(aq) + H_2(g) + \text{heat}$$

Step 2: Analyzing storage options

- **(1) Water:** Sodium reacts vigorously, even explosively, with water. So, it cannot be stored under water.

$$2Na(s) + 2H_2O(l) \rightarrow 2NaOH(aq) + H_2(g)$$

  The heat generated can ignite the hydrogen gas produced.

- **(2) Kerosene Oil:** Kerosene is a hydrocarbon oil. Sodium does not react with kerosene. Kerosene also prevents sodium from coming into contact with air (oxygen and moisture). Therefore, sodium is commonly stored under kerosene oil or other similar inert mineral oils.

- **(3) Alcohol (e.g., Ethanol C_2H_5OH):** Sodium reacts with alcohols (which have an -OH group) to produce sodium alkoxides and hydrogen gas, although the reaction is generally less vigorous than with water.

$$2Na(s) + 2C_2H_5OH(l) \rightarrow 2NaOC_2H_5(alc) + H_2(g)$$

  So, alcohol is not suitable for storing sodium.

- **(4) Hydrogen:** Storing a reactive metal under a gas is generally not practical for preventing surface reactions with air unless it's a completely sealed, inert atmosphere of pure hydrogen, which is not how it's "usually stored". Also, sodium can form sodium hydride (NaH) with hydrogen under certain conditions (e.g., heating).

Step 3: Identifying the usual storage method Due to its high reactivity with air and water, sodium metal is usually stored under an inert liquid like **kerosene oil** to protect it from oxidation and reaction with moisture.

Quick Tip

Sodium is a very reactive metal.

- It reacts violently with **water**.
- It reacts with **oxygen and moisture in the air**.
- It reacts with **alcohols**.

Kerosene oil acts as a protective blanket, keeping air and moisture away from the sodium, thus preventing these reactions.

---

113. Galvanized Iron (GI) sheets are coated with :

(1) Sn

(2) Ni

(3) Cu

(4) Zn

Correct Answer: (4) Zn

Solution: Concept: Galvanization is a process of applying a protective coating to iron or steel to prevent rusting (corrosion).

Step 1: Understanding Galvanization Rusting is the corrosion of iron due to its reaction with oxygen and moisture in the atmosphere, forming hydrated iron(III) oxide.

Galvanization is a common method to protect iron from rusting. The process involves coating the iron or steel with a layer of another metal.

Step 2: The Metal Used in Galvanization The metal most commonly used for galvanizing iron is **Zinc (Zn)**. The process typically involves dipping the iron or steel object into molten zinc (hot-dip galvanizing) or using electroplating to deposit a zinc layer.

Step 3: How Zinc Protects Iron Zinc protects iron in two main ways:

1. **Barrier Protection:** The zinc coating forms a physical barrier that prevents oxygen and water from reaching the iron surface.

2. **Sacrificial Protection (Cathodic Protection):** Zinc is more reactive than iron (it is higher in the electrochemical series). If the zinc coating is scratched or broken and the iron is exposed, the zinc will corrode preferentially (acts as a sacrificial anode) instead of the iron. The zinc "sacrifices" itself to protect the iron.

Step 4: Analyzing the options

- **(1) Sn (Tin):** Coating iron with tin is called "tinning" (e.g., tin cans). Tin protects iron mainly by barrier protection. If the tin coating is scratched, the iron underneath can rust rapidly because iron is more reactive than tin (tin provides cathodic protection to iron only in certain environments, but generally, iron is anodic to tin).
- **(2) Ni (Nickel):** Nickel plating can also protect iron, usually by electroplating.
- **(3) Cu (Copper):** Copper plating can be used, but copper is less reactive than iron. If the copper coating is scratched, the iron will corrode.
- **(4) Zn (Zinc):** This is the metal used for galvanization, providing both barrier and sacrificial protection.

Therefore, galvanized iron sheets are coated with zinc.

Quick Tip

Galvanization = Coating iron with **Z**inc (**Zn**). Remember the "Z" in Zinc for galvani**z**ation. Zinc acts as a protective layer and also "sacrifices" itself to prevent the iron from rusting if the coating is damaged.

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114. The most abundant element found in the earth's crust is :

(1) O

(2) Cl

(3) Si

(4) S

Correct Answer: (1) O

Solution: Concept: The Earth's crust is composed of various elements, but some are significantly more abundant than others by mass.

Step 1: Major Elements in Earth's Crust by Approximate Percentage Mass

1. **Oxygen (O):** About 46.6%
2. **Silicon (Si):** About 27.7%
3. **Aluminum (Al):** About 8.1%
4. **Iron (Fe):** About 5.0%
5. **Calcium (Ca):** About 3.6%
6. **Sodium (Na):** About 2.8%
7. **Potassium (K):** About 2.6%
8. **Magnesium (Mg):** About 2.1%

All other elements make up the remaining percentage (less than 2%).

Step 2: Identifying the most abundant element From the list, **Oxygen (O)** is the most abundant element in the Earth's crust by mass. It is primarily found in combined form in silicate minerals (like quartz, feldspar, mica), oxides, carbonates, sulfates, etc.

Step 3: Evaluating the options

- **(1) O (Oxygen):** $\sim 46.6\%$ - Most abundant.
- **(2) Cl (Chlorine):** Much less abundant than oxygen or silicon. Found in minerals like halite (NaCl).
- **(3) Si (Silicon):** $\sim 27.7\%$ - Second most abundant.
- **(4) S (Sulfur):** Less abundant than oxygen or silicon. Found in sulfide and sulfate minerals.

Therefore, Oxygen is the most abundant element in the Earth's crust.

Quick Tip

Remember the top two most abundant elements in the Earth's crust (by mass): 1. **O**xygen ($\sim 47\%$) 2. **Si**licon ($\sim 28\%$) These two elements are the main components of most rocks and minerals (silicates). (Mnemonic: "**O**h, **Si**lly **Al**fie..." for Oxygen, Silicon, Aluminum, Iron).

---

115. Calamine is an ore of which of the following ?

(1) Ca

(2) Zn

(3) Mg

(4) Hg

Correct Answer: (2) Zn

Solution: Concept: Ores are naturally occurring rocks or minerals from which metals can be extracted economically. Calamine is a historical name for an ore.

Step 1: Identifying Calamine Calamine is an ore that is primarily composed of **zinc carbonate ($ZnCO_3$)**. The mineral name for zinc carbonate is smithsonite. Historically, "calamine" has also sometimes referred to hydrated zinc silicate (hemimorphite), but it is most commonly associated with zinc carbonate. In either case, it is an ore of zinc.

Step 2: Determining the metal obtained from Calamine Since calamine is mainly zinc carbonate ($ZnCO_3$), the metal that can be extracted from it is **Zinc (Zn)**.

Step 3: Analyzing the options

- **(1) Ca (Calcium):** Ores of calcium include limestone ($CaCO_3$), gypsum ($CaSO_4 \cdot 2H_2O$).
- **(2) Zn (Zinc):** Calamine ($ZnCO_3$) is an ore of zinc. Other zinc ores include sphalerite/zinc blende (ZnS) and zincite (ZnO).
- **(3) Mg (Magnesium):** Ores of magnesium include magnesite ($MgCO_3$), dolomite ($CaCO_3 \cdot MgCO_3$), carnallite.
- **(4) Hg (Mercury):** The principal ore of mercury is cinnabar (HgS).

Therefore, calamine is an ore of zinc. (Calamine lotion, known for its soothing properties, also contains zinc oxide or zinc carbonate).

Quick Tip

Calamine is an ore of **Zinc (Zn)**. Its main chemical composition is zinc carbonate ($ZnCO_3$). Think of calamine lotion, which contains zinc compounds.

---

116. The red colouring pigment Haemoglobin in our blood contains which metal ?

(1) Fe

(2) Co

(3) Mg

(4) Na

Correct Answer: (1) Fe

Solution: Concept: Haemoglobin (often spelled Hemoglobin) is a protein in red blood cells responsible for transporting oxygen in the blood of vertebrates and some invertebrates. Its structure includes a metal ion.

Step 1: Structure and Function of Haemoglobin Haemoglobin is a complex protein containing a non-protein component called **haem** (or heme). Each haem group contains a central metal atom. This metal atom is crucial for the oxygen-binding capacity of haemoglobin. The primary function of haemoglobin is to bind to oxygen in the lungs (or gills) and transport it to the tissues throughout the body, where it releases the oxygen.

Step 2: The Metal in Haemoglobin The metal atom at the center of each haem group in haemoglobin is **Iron (Fe)**. Specifically, it is an iron(II) ion, Fe^{2+}. Each haemoglobin molecule typically has four haem groups, and thus can bind up to four oxygen molecules. The binding of oxygen to the iron atom is what gives oxygenated blood its bright red color, while deoxygenated blood is a darker red/purplish.

Step 3: Analyzing the options

- **(1) Fe (Iron):** Correct. Iron is the central metal atom in the haem group of haemoglobin.

- **(2) Co (Cobalt):** Cobalt is the central metal atom in Vitamin B12 (cobalamin), but not in haemoglobin.
- **(3) Mg (Magnesium):** Magnesium is the central metal atom in chlorophyll, the pigment responsible for photosynthesis in plants, but not in haemoglobin.
- **(4) Na (Sodium):** Sodium is an important electrolyte in the body (as Na^+ ions) involved in nerve function and fluid balance, but it is not the metal in haemoglobin.

Therefore, the metal in haemoglobin is Iron (Fe).

Quick Tip

Think "Haem" sounds like "Heme", which contains Iron.

- **Haem**oglobin (in blood) $\rightarrow$ contains **Fe** (Iron) - for oxygen transport.
- **Chl**orophyll (in plants) $\rightarrow$ contains **Mg** (Magnesium) - for photosynthesis.

Iron deficiency can lead to anemia because less haemoglobin can be made.

---

117. Which of the following organic compounds will exhibit Geometrical Isomerism ?

(1) Butanol-2

(2) Butene-1

(3) Butene-2

(4) Butyne-1

Correct Answer: (3) Butene-2

Solution: Concept: Geometrical isomerism (also known as cis-trans isomerism) is a type of stereoisomerism that can occur in molecules with restricted rotation around a bond, typically a carbon-carbon double bond ($C = C$) or in cyclic compounds. For geometrical isomerism around a $C = C$ double bond, two conditions must be met:

1. There must be restricted rotation around the $C = C$ bond (which is inherent to double bonds).

2. **Each** carbon atom of the double bond must be attached to **two different** atoms or groups.

Step 1: Analyze the structures of the given compounds

- **(1) Butanol-2 (2-Butanol):** $CH_3 - CH(OH) - CH_2 - CH_3$. This is an alcohol and does not have a $C = C$ double bond relevant for geometrical isomerism in its main chain. (It has a chiral center and can exhibit optical isomerism, but not geometrical isomerism in the typical alkene sense).

- **(2) Butene-1 (1-Butene):** $CH_2 = CH - CH_2 - CH_3$. Let's look at the carbons of the double bond ($C_1 = C_2$):

  - C_1 is attached to two Hydrogen atoms (H, H). Since these two groups are identical, geometrical isomerism is not possible around this bond.
  - C_2 is attached to a Hydrogen atom (H) and an Ethyl group ($-CH_2CH_3$).

  Since C_1 has two identical groups (H, H), 1-Butene does not exhibit geometrical isomerism.

- **(3) Butene-2 (2-Butene):** $CH_3 - CH = CH - CH_3$. Let's look at the carbons of the double bond ($C_2 = C_3$):

  - C_2 is attached to a Hydrogen atom (H) and a Methyl group ($-CH_3$). These are different.
  - C_3 is attached to a Hydrogen atom (H) and a Methyl group ($-CH_3$). These are different.

  Since both C_2 and C_3 are each attached to two different groups, 2-Butene can exhibit geometrical isomerism (cis-2-butene and trans-2-butene).

  - *cis*-2-Butene: Methyl groups are on the same side of the double bond.
  - *trans*-2-Butene: Methyl groups are on opposite sides of the double bond.

- **(4) Butyne-1 (1-Butyne):** $HC \equiv C - CH_2 - CH_3$. This molecule has a carbon-carbon triple bond ($C \equiv C$). Triple bonds are linear around the alkyne carbons and do not exhibit geometrical isomerism of the cis-trans type.

Step 2: Identifying the compound exhibiting geometrical isomerism Based on the analysis, **Butene-2 (2-Butene)** meets the criteria for geometrical isomerism.

Quick Tip

For geometrical (cis-trans) isomerism in alkenes (compounds with $C = C$): 1. Find the $C = C$ double bond. 2. Look at each carbon of the double bond separately. 3. Each carbon of the double bond MUST have **two different groups** attached to it. Example: For $R_1R_2C = CR_3R_4$, you need $R_1 \neq R_2$ AND $R_3 \neq R_4$.

- 1-Butene ($CH_2 = CH - CH_2CH_3$): First carbon has two H's (same) $\rightarrow$ No.
- 2-Butene ($CH_3 - CH = CH - CH_3$): Each double-bonded carbon has one H and one CH_3 (different) $\rightarrow$ Yes.

---

118. The most pure form of Iron is :

(1) Cast Iron

(2) Pig Iron

(3) Wrought Iron

(4) Steel

Correct Answer: (3) Wrought Iron

Solution: Concept: Iron is commercially produced in several forms, which differ mainly in their carbon content and the presence of other impurities. Purity refers to the percentage of iron, with lower carbon and impurity content indicating higher purity.

Step 1: Understanding different forms of commercial Iron

- **Pig Iron:** This is the crude iron obtained directly from a blast furnace. It has a high carbon content, typically 3.5% to 4.5%, and also contains other impurities like silicon, manganese, phosphorus, and sulfur. It is very brittle and not very useful directly for most applications.
- **Cast Iron:** Produced by re-melting pig iron, often with scrap iron and steel, and then casting it into molds. The carbon content is typically 2% to 4%. It is hard and brittle but

has good fluidity for casting. Impurities are still present.

- **Wrought Iron:** This is historically the purest form of commercial iron. It has a very low carbon content, usually less than 0.08% (often around 0.02-0.03%), and very few other impurities. It is tough, malleable, and ductile, but softer than steel. It contains some slag (fibrous inclusions of iron silicate), which gives it a characteristic grain.
- **Steel:** Steel is an alloy of iron and carbon, typically with a carbon content between 0.2% and 2.1% by weight. The properties of steel can be varied widely by changing the carbon content and adding other alloying elements (like manganese, chromium, vanadium, tungsten). Steel is generally stronger and more versatile than cast iron or wrought iron.

Step 2: Comparing Carbon Content (as an indicator of purity)

- Pig Iron: $\sim 3.5 - 4.5\%$ C (Least pure)
- Cast Iron: $\sim 2 - 4\%$ C
- Steel: $\sim 0.2 - 2.1\%$ C
- **Wrought Iron:** $< 0.08\%$ C (Most pure commercial form)

Step 3: Identifying the most pure form Based on the low carbon content and fewer impurities, **Wrought Iron** is considered the most pure form of commercial iron among the options.

Quick Tip

Purity of iron forms is mainly related to carbon content:

- **Pig Iron:** Highest carbon (least pure).
- **Cast Iron:** High carbon.
- **Steel:** Medium carbon (alloy, properties vary).
- **Wrought Iron:** Lowest carbon (most pure commercial iron).

Think: Wrought iron is "worked" iron, refined to remove impurities.

119. The property of carbon element responsible for a large number of organic compounds is :

(1) Allotropy

(2) Catenation

(3) Hybridisation

(4) None of these

Correct Answer: (2) Catenation

Solution: Concept: Carbon is unique in its ability to form a vast number of compounds. This is due to specific properties of the carbon atom.

Step 1: Key Properties of Carbon leading to many compounds

1. **Catenation:** This is the ability of an element to form long chains and rings by bonding to itself. Carbon exhibits catenation to a remarkable extent, forming stable carbon-carbon bonds. These chains and rings can be of various lengths and structures (straight, branched).

2. **Tetravalency:** Carbon has four valence electrons, allowing it to form four covalent bonds with other atoms (including other carbon atoms or atoms of different elements like hydrogen, oxygen, nitrogen, halogens).

3. **Ability to form multiple bonds:** Carbon can form single ($C - C$), double ($C = C$), and triple ($C \equiv C$) bonds with itself and other elements, further increasing the variety of possible structures.

4. **Isomerism:** The ability of compounds with the same molecular formula to exist in different structural arrangements, leading to different properties. This is enabled by carbon's bonding versatility.

The question asks for *the* property responsible for a *large number* of organic compounds. While tetravalency and multiple bonding are essential, catenation is the primary reason for the sheer number and diversity of carbon chains and rings that form the backbone of organic molecules.

Step 2: Analyzing the options

- **(1) Allotropy:** This is the property of some chemical elements to exist in two or more different forms (allotropes) in the same physical state. For example, carbon exists as diamond, graphite, fullerenes, etc. While carbon shows allotropy, this property itself doesn't explain the vast number of *different organic compounds* (which involve carbon bonded to other elements as well).
- **(2) Catenation:** This self-linking property of carbon to form long chains, branched chains, and rings is a fundamental reason for the immense diversity and large number of organic compounds.
- **(3) Hybridisation (sp^3, sp^2, sp):** This is a concept that explains the geometry and bonding of carbon atoms (e.g., tetrahedral, trigonal planar, linear shapes). Hybridisation enables carbon to form its diverse bonds and structures, supporting catenation and tetravalency, but catenation is the more direct answer to "property responsible for a large number."
- **(4) None of these:** Incorrect, as catenation is a key property.

Step 3: Identifying the primary property Catenation is the most direct and significant property of carbon that accounts for the formation of a vast number of organic compounds.

Quick Tip

Carbon is special because it can link to itself to form long chains and rings. This self-linking ability is called **catenation**. Think of it like carbon atoms holding hands with many other carbon atoms to build complex skeletons for molecules. This, along with its ability to form four bonds (tetravalency), allows for millions of different organic compounds.

---

120. Which of the following is a natural polymer ?

(1) Bakelite

(2) PVC

(3) Polythene

(4) Protein

Correct Answer: (4) Protein

Solution: Concept: Polymers are large molecules (macromolecules) composed of repeating structural units called monomers. They can be classified as natural or synthetic.

- **Natural Polymers:** Occur in nature, produced by living organisms (plants and animals).
- **Synthetic Polymers:** Man-made in laboratories or factories, often from petroleum-based chemicals.

Step 1: Analyzing the options

- **(1) Bakelite (Phenol-formaldehyde resin):** This is one of the first synthetic plastics, made from phenol and formaldehyde. It is a synthetic polymer.
- **(2) PVC (Polyvinyl Chloride):** Made by polymerizing vinyl chloride monomer. It is a widely used synthetic plastic (e.g., for pipes, window frames).
- **(3) Polythene (Polyethylene):** Made by polymerizing ethylene (ethene) monomer. It is a very common synthetic plastic (e.g., for plastic bags, bottles).
- **(4) Protein:** Proteins are complex macromolecules essential to all living organisms. They are polymers made up of repeating monomer units called **amino acids**, linked by peptide bonds. Proteins are naturally occurring. Examples include enzymes, antibodies, keratin (in hair/nails), collagen (in skin/bones).

Step 2: Identifying the natural polymer From the analysis:

- Bakelite, PVC, and Polythene are synthetic polymers.
- **Protein** is a natural polymer.

Other examples of natural polymers include:

- Starch (polymer of glucose)
- Cellulose (polymer of glucose, main component of plant cell walls)
- Natural Rubber (polymer of isoprene)
- DNA and RNA (nucleic acids, polymers of nucleotides)

- Silk (a protein fiber)
- Wool (a protein fiber)

Therefore, Protein is a natural polymer among the given options.

Quick Tip

- **Natural polymers** are made by nature (plants, animals). Examples: Protein, Starch, Cellulose, DNA, Natural Rubber.
- **Synthetic polymers** are man-made. Examples: Polythene (plastic bags), PVC (pipes), Bakelite (old plastics), Nylon, Polyester.

Proteins are built from amino acid monomers and are fundamental to life.

Biology

121. Which one is made of dead cells :

(1) Parenchyma

(2) Meristematic tissues

(3) Sclerenchyma

(4) Companion cells

Correct Answer: (3) Sclerenchyma

Solution: Concept: Plant tissues are composed of cells that can be either living or dead at maturity, depending on their function.

Step 1: Understanding different plant tissues

- **Parenchyma:** These are simple permanent tissues composed of living cells with thin cell walls. They perform various functions like photosynthesis, storage, and secretion.
- **Meristematic tissues (Meristems):** These are regions of actively dividing cells responsible for plant growth (e.g., at the tips of roots and shoots). Meristematic cells are living.

- **Sclerenchyma:** This is a simple permanent tissue that provides mechanical support and strength to plants. Sclerenchyma cells are characterized by thick, lignified secondary cell walls. These cells are typically **dead at maturity**. Their primary function (support) is provided by their rigid cell walls, and they do not need to be metabolically active. Sclerenchyma includes fibers and sclereids (stone cells).
- **Companion cells:** These are specialized parenchyma cells found in the phloem of flowering plants. They are living cells and are closely associated with sieve tube elements, providing metabolic support to them.

Step 2: Identifying the tissue made of dead cells Based on the descriptions:

- Parenchyma: Living cells.
- Meristematic tissues: Living, actively dividing cells.
- **Sclerenchyma:** Dead cells at maturity, providing structural support.
- Companion cells: Living cells.

Therefore, Sclerenchyma is the tissue made of dead cells among the options.

Quick Tip

In plants:

- **Living tissues** for growth, photosynthesis, storage: Parenchyma, Collenchyma, Meristems, Phloem (sieve tubes need companion cells).
- **Dead tissues** for support and water transport: **Sclerenchyma** (fibers, sclereids) and Xylem vessels/tracheids.

Sclerenchyma cells have very thick, hard walls that provide strength, and they are dead when they perform this function.

---

122. Nodules with nitrogen-fixing bacteria are present in :

(1) Musturd

(2) Gram

(3) Wheat

(4) Cotton

Correct Answer: (2) Gram

Solution: Concept: Nitrogen fixation is the process by which atmospheric nitrogen (N_2) is converted into ammonia (NH_3) or other nitrogenous compounds that can be used by plants. Some bacteria are capable of this process, and they often form symbiotic relationships with certain types of plants.

Step 1: Symbiotic Nitrogen Fixation and Root Nodules A well-known example of symbiotic nitrogen fixation occurs in leguminous plants (members of the Fabaceae or Leguminosae family). Bacteria, typically of the genus *Rhizobium*, infect the roots of these plants and induce the formation of specialized structures called **root nodules**. Inside these nodules, the bacteria live and convert atmospheric nitrogen into ammonia, which the plant can then assimilate to produce proteins and other essential nitrogen-containing compounds. In return, the plant provides the bacteria with carbohydrates (energy) and a protected environment.

Step 2: Identifying Leguminous Plants among the options

- **(1) Mustard (Brassica species):** Belongs to the Brassicaceae family. Not a legume.
- **(2) Gram (Cicer arietinum - Chickpea):** Gram is a type of pulse and belongs to the Fabaceae (legume) family. Leguminous plants like gram are known for forming root nodules with nitrogen-fixing bacteria.
- **(3) Wheat (Triticum species):** Belongs to the Poaceae family (grasses). Not a legume.
- **(4) Cotton (Gossypium species):** Belongs to the Malvaceae family. Not a legume.

Step 3: Conclusion Since Gram (chickpea) is a leguminous plant, it is the one among the options that would have root nodules containing nitrogen-fixing bacteria.

Quick Tip

Nitrogen-fixing root nodules are a hallmark of **leguminous plants** (family Fabaceae/Leguminosae). Common legumes include:

- Peas, Beans, Lentils
- **Gram** (Chickpea), Soybeans, Peanuts
- Clover, Alfalfa

These plants work with *Rhizobium* bacteria to "fix" atmospheric nitrogen.

---

123. Which of the following is not a green-house gas ?

(1) Methane

(2) Carbon-di-oxide

(3) Carbon mono-oxide

(4) Ammonia

Correct Answer: (3) Carbon mono-oxide

Solution: Concept: Greenhouse gases are gases in the atmosphere that absorb and emit infrared radiation (heat). This property causes the greenhouse effect, which warms the Earth's surface.

Step 1: Major Greenhouse Gases The primary greenhouse gases in Earth's atmosphere are:

- **Water Vapor (H_2O):** The most abundant greenhouse gas, responsible for a large part of the natural greenhouse effect.
- **Carbon Dioxide (CO_2):** A major contributor, its concentration has significantly increased due to human activities (burning fossil fuels, deforestation).
- **Methane (CH_4):** Produced by natural sources (wetlands) and human activities (agriculture, livestock, fossil fuel extraction). It is a potent greenhouse gas.
- **Nitrous Oxide (N_2O):** Emitted from agricultural and industrial activities, as well as natural processes.

- **Ozone (O_3):** In the troposphere (lower atmosphere), ozone acts as a greenhouse gas.
- Fluorinated gases (e.g., HFCs, PFCs, SF_6): Synthetic, potent greenhouse gases.

Step 2: Analyzing the options

- **(1) Methane (CH_4):** A significant greenhouse gas.
- **(2) Carbon-di-oxide (CO_2):** The most discussed anthropogenic greenhouse gas.
- **(3) Carbon mono-oxide (CO):** Carbon monoxide is not considered a *direct* greenhouse gas because it does not absorb terrestrial infrared radiation strongly itself. However, it is an *indirect* greenhouse gas because it reacts with hydroxyl radicals (OH) in the atmosphere. This reaction reduces the availability of OH radicals, which are important for removing other greenhouse gases like methane. So, CO can lead to an increase in the lifetime and concentration of methane. But directly, it's not a strong IR absorber.
- **(4) Ammonia (NH_3):** Ammonia can absorb infrared radiation and act as a weak greenhouse gas. Its atmospheric lifetime is short, but it can contribute to the formation of particulate matter, which has complex effects on climate. In many contexts, especially introductory ones, it might not be listed as a primary greenhouse gas compared to CO_2, CH_4, N_2O. The image indicates this option was circled.

Step 3: Identifying the one that is "not" a (direct) greenhouse gas Comparing the options, Carbon Dioxide and Methane are definitely major direct greenhouse gases. Ammonia has some direct greenhouse effect. Carbon Monoxide has a very weak direct effect but a more significant indirect effect. In typical lists focusing on direct major greenhouse gases, **Carbon Monoxide (CO)** is often the one considered "not a greenhouse gas" in the direct sense, or at least the least significant direct one among these options. If the question implies "not a significant direct greenhouse gas," then CO is the best fit.
Given the common understanding in introductory contexts, CO is the most likely answer for "not a greenhouse gas."

Quick Tip

Key direct greenhouse gases to remember:

- Water vapor (H_2O)
- Carbon dioxide (CO_2)
- Methane (CH_4)
- Nitrous oxide (N_2O)
- Ozone (O_3)

Carbon monoxide (CO) doesn't trap heat directly very well, but it can affect the concentration of other greenhouse gases like methane. Ammonia (NH_3) is a minor greenhouse gas.

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124. Kala-azar is caused by :

(1) protozoan

(2) fungus

(3) bacteria

(4) virus

Correct Answer: (1) protozoan

Solution: Concept: Kala-azar, also known as visceral leishmaniasis, is a serious parasitic disease. Understanding the type of microorganism that causes it is important.

Step 1: Identifying the Causative Agent of Kala-azar Kala-azar is caused by a type of parasitic **protozoan** belonging to the genus *Leishmania*. The most common species causing visceral leishmaniasis is *Leishmania donovani*. Protozoa are single-celled eukaryotic microorganisms.

Step 2: Transmission of Kala-azar The *Leishmania* parasites are transmitted to humans through the bite of infected female phlebotomine sandflies.

Step 3: Analyzing the options

- **(1) protozoan:** Correct. Kala-azar is caused by *Leishmania* protozoa.

- **(2) fungus:** Fungi cause diseases like ringworm, athlete's foot, or candidiasis. Not Kala-azar.
- **(3) bacteria:** Bacteria cause diseases like tuberculosis, cholera, or typhoid. Not Kala-azar.
- **(4) virus:** Viruses cause diseases like the common cold, influenza, measles, or AIDS. Not Kala-azar.

Therefore, Kala-azar is caused by a protozoan.

> Quick Tip
>
> Kala-azar (Visceral Leishmaniasis) is a parasitic disease. The parasite responsible is a **protozoan** of the genus *Leishmania*. It is transmitted by the bite of infected sandflies. Remember: Kala-azar $\rightarrow$ *Leishmania* $\rightarrow$ Protozoan.

---

125. Amphibians do not have the following :

(1) three chambered heart

(2) gills or lungs

(3) scales

(4) mucus glands

Correct Answer: (3) scales

Solution: Concept: Amphibians (Class Amphibia) are a class of cold-blooded vertebrates that include frogs, toads, salamanders, newts, and caecilians. They have distinct characteristics. The question asks what they *do not* typically have.

Step 1: General Characteristics of Amphibians

- **Skin:** Moist, permeable skin that lacks scales (with very few exceptions like some caecilians having dermal scales, but generally, scales are absent). Their skin often contains **mucus glands** to keep it moist, and sometimes poison glands. Cutaneous respiration (breathing through the skin) is common.

- **Respiration:** Larval amphibians (like tadpoles) typically breathe through **gills**. Adult amphibians usually breathe through **lungs** and also through their skin.
- **Heart:** Most adult amphibians have a **three-chambered heart** (two atria and one ventricle). This leads to some mixing of oxygenated and deoxygenated blood.
- **Life Cycle:** Typically undergo metamorphosis from an aquatic larval stage to a terrestrial or semi-aquatic adult stage.
- **Reproduction:** Most lay eggs in water, which lack a protective shell.

Step 2: Analyzing the options

- **(1) three chambered heart:** Most adult amphibians **do have** a three-chambered heart.
- **(2) gills or lungs:** Amphibians **do have** gills (in larval stage, and some aquatic adults) or lungs (in most adults). They also use cutaneous respiration.
- **(3) scales:** The vast majority of amphibians **do not have** scales on their skin. Their skin is typically smooth and moist. This is a distinguishing feature compared to reptiles (which have scales) and fish (which often have scales).
- **(4) mucus glands:** Amphibians **do have** mucus glands in their skin, which help keep it moist for cutaneous respiration and protection. The image indicates this option was circled, but amphibians do possess mucus glands.

Step 3: Identifying what amphibians generally do not have Based on general characteristics, amphibians typically **do not have scales**. While there are rare exceptions (some caecilians have dermal scales), the absence of scales is a defining trait for the class as a whole when compared to reptiles or fish. They do have three-chambered hearts (most adults), gills/lungs for respiration, and mucus glands.

Quick Tip

Amphibians are known for their smooth, moist skin.

- They generally **LACK scales** (unlike reptiles or fish).
- They **HAVE** mucus glands to keep skin moist.
- They **HAVE** a three-chambered heart (most adults).
- They **HAVE** gills (as larvae) and/or lungs (as adults) for breathing.

The most prominent feature from the list that they generally "do not have" is scales.

---

126. Which of the following species of honey bee is an Italian species ?

(1) Apis mellifera

(2) Apis dorsata

(3) Apis florae

(4) Apis cerana indica

Correct Answer: (1) Apis mellifera

Solution: Concept: Different species and subspecies of honey bees are found around the world, some of which are commonly used in apiculture (beekeeping).

Step 1: Identifying Common Honey Bee Species

- **Apis mellifera (Western Honey Bee or European Honey Bee):** This is one of the most common and widely domesticated honey bee species globally. It has many subspecies, some of which are specifically referred to by their region of origin. *Apis mellifera ligustica* is the Italian honey bee, a very popular subspecies in beekeeping worldwide due to its productivity and relatively gentle nature. The general species *Apis mellifera* encompasses these European varieties.
- **Apis dorsata (Giant Honey Bee):** A large, wild honey bee species found in South and Southeast Asia. Known for building large, exposed combs on high structures. It is aggressive and not easily domesticated.

- **Apis florae (Dwarf Honey Bee):** A small, wild honey bee species also found in South and Southeast Asia. Builds small, single combs.
- **Apis cerana indica (Indian Honey Bee or Asiatic Honey Bee):** A species native to South, Southeast, and East Asia. It is commonly kept in traditional and modern hives in these regions.

Step 2: Identifying the Italian Species The Italian honey bee is a well-known subspecies of *Apis mellifera*, specifically *Apis mellifera ligustica*. Since *Apis mellifera* is listed as an option, and it is the species to which the Italian bee belongs, this is the correct answer.

Quick Tip

The **Italian honey bee** is a specific subspecies called *Apis mellifera ligustica*. It belongs to the broader species ***Apis mellifera***, which is also known as the Western or European honey bee. This species (*A. mellifera*) is widely used in beekeeping globally.

---

127. Enzyme present in the saliva of human being ?

(1) Pepsin

(2) Trypsin

(3) Tylin

(4) Lipase

Correct Answer: (3) Tylin

Solution: Concept: Saliva is a digestive fluid secreted by salivary glands in the mouth. It contains enzymes that begin the process of digestion.

Step 1: Composition and Function of Saliva Saliva has several functions, including:

- Moistening food to aid in chewing and swallowing.
- Starting the chemical digestion of carbohydrates.
- Containing antimicrobial agents.

The primary digestive enzyme in human saliva is **salivary amylase**. An older name for salivary amylase is **ptyalin** (often seen as ”tylin” in some older texts or simpler questions).

Step 2: Action of Salivary Amylase (Ptyalin/Tylin) Salivary amylase begins the breakdown of starch (a complex carbohydrate) into smaller sugar molecules, such as maltose and dextrins.

Step 3: Analyzing the options

- **(1) Pepsin:** A digestive enzyme found in the stomach that breaks down proteins into peptides. It functions in an acidic environment.
- **(2) Trypsin:** A digestive enzyme produced by the pancreas and active in the small intestine. It also breaks down proteins.
- **(3) Tylin (Ptyalin / Salivary Amylase):** This is the enzyme present in saliva that digests carbohydrates (starch).
- **(4) Lipase:** A group of enzymes that break down fats (lipids). Saliva contains a very small amount of lingual lipase, which begins fat digestion, but salivary amylase (ptyalin/tylin) is the major and most well-known digestive enzyme in saliva specifically for carbohydrates. Given the options, "Tylin" refers to salivary amylase.

Therefore, Tylin (Ptyalin), which is salivary amylase, is the enzyme present in human saliva that initiates carbohydrate digestion.

Quick Tip

The main digestive enzyme in human saliva is **salivary amylase**. An older, but sometimes still used, name for salivary amylase is **Ptyalin** (or **Tylin** as written in the option). This enzyme starts the digestion of **starch** in the mouth.

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128. Oxygenated blood is present in :

(1) Right auricle

(2) Pulmonary artery

(3) Right ventricle

(4) Pulmonary vein

Correct Answer: (4) Pulmonary vein

Solution: Concept: The human circulatory system involves the heart pumping blood through two main circuits: the pulmonary circuit (to the lungs) and the systemic circuit (to the rest of the body). Oxygenated blood is rich in oxygen, while deoxygenated blood is poor in oxygen.

Step 1: Path of Blood Flow and Oxygenation

1. **Deoxygenated blood** from the body returns to the **Right Auricle (Right Atrium)** of the heart via the vena cavae.
2. From the Right Auricle, deoxygenated blood flows into the **Right Ventricle**.
3. The Right Ventricle pumps this deoxygenated blood into the **Pulmonary Artery**, which carries it to the lungs.
4. In the lungs, the blood releases carbon dioxide and picks up oxygen (becomes oxygenated).
5. **Oxygenated blood** from the lungs returns to the **Left Auricle (Left Atrium)** of the heart via the **Pulmonary Veins**.
6. From the Left Auricle, oxygenated blood flows into the Left Ventricle.
7. The Left Ventricle pumps this oxygenated blood into the Aorta, which distributes it to the rest of the body (systemic circulation).

Step 2: General Rules and Exceptions for Arteries and Veins

- Generally, **arteries** carry oxygenated blood away from the heart to the body.
- Generally, **veins** carry deoxygenated blood from the body back to the heart.

Exceptions (Pulmonary Circuit):

- The **Pulmonary Artery** is an exception: it carries *deoxygenated* blood from the heart (right ventricle) to the lungs.
- The **Pulmonary Veins** are an exception: they carry *oxygenated* blood from the lungs to the heart (left auricle).

Step 3: Analyzing the options

- **(1) Right auricle:** Receives deoxygenated blood from the body.
- **(2) Pulmonary artery:** Carries deoxygenated blood from the right ventricle to the lungs.
- **(3) Right ventricle:** Pumps deoxygenated blood to the pulmonary artery.
- **(4) Pulmonary vein:** Carries oxygenated blood from the lungs to the left auricle.

Therefore, oxygenated blood is present in the Pulmonary vein.

Quick Tip

Remember the exceptions in the pulmonary (lung) circuit:

- **Pulmonary Artery**: Carries **DE**oxygenated blood (to lungs). (Arteries usually carry Oxygenated blood Away from heart, but this is an exception).
- **Pulmonary Vein**: Carries **O**xygenated blood (from lungs). (Veins usually carry Deoxygenated blood towards heart, but this is an exception).

The right side of the heart (right auricle, right ventricle) deals with deoxygenated blood. The left side deals with oxygenated blood.

---

129. Evaporation of water from the surface of leaf is called :

(1) Respiration

(2) Photosynthesis

(3) Transpiration

(4) Evaporation

Correct Answer: (3) Transpiration

Solution: Concept: Plants lose water vapor to the atmosphere through a specific process, primarily occurring via small pores on their leaves.

Step 1: Defining the process The process by which water is carried through plants from roots to small pores (stomata) on the underside of leaves, where it changes to vapor and is released to the atmosphere, is called **Transpiration**. Essentially, it is the evaporation of water from plant leaves, stems, and flowers.

Step 2: Analyzing the options

- **(1) Respiration:** In plants (as in animals), respiration is the metabolic process of converting biochemical energy from nutrients into ATP, and then releasing waste products. It involves gas exchange (uptake of O_2, release of CO_2), but it's not primarily about water loss from leaf surfaces.
- **(2) Photosynthesis:** The process by which green plants use sunlight, water, and carbon dioxide to create their own food (glucose) and release oxygen. It is an energy-storing process.
- **(3) Transpiration:** This is the specific term for the evaporation of water from plant surfaces, especially leaves. It plays a role in cooling the plant and in the upward movement of water and minerals from the roots (transpirational pull).
- **(4) Evaporation:** This is a general term for the process where a liquid turns into a gas. While transpiration is a form of evaporation, "transpiration" is the more precise biological term for this specific process in plants.

Step 3: Identifying the correct term The specific term for the evaporation of water from the surface of a leaf (and other plant parts) is **Transpiration**.

Quick Tip

Think of plants "sweating" – they release water vapor. This specific process of water loss as vapor from plant leaves is called **Transpiration**. It's like evaporation, but specifically from plants. It helps pull water up from the roots.

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130. Auxin is a :

(1) Plant harmone

(2) Enzyme

(3) Fat

(4) Protein

Correct Answer: (1) Plant hormone

Solution: Concept: Auxins are a class of chemical substances that play important roles in regulating plant growth and development.

Step 1: Understanding Auxins Auxins are one of the major groups of **plant hormones** (also called phytohormones or plant growth regulators). They are produced primarily in the apical meristems (growing tips of shoots and roots), young leaves, and developing fruits and seeds. Key functions of auxins include:

- Stimulating cell elongation (leading to growth).
- Involved in phototropism (plant bending towards light) and gravitropism (growth response to gravity).
- Promoting root initiation.
- Regulating fruit development.
- Apical dominance (suppression of lateral bud growth by the apical bud).

Indole-3-acetic acid (IAA) is the most common naturally occurring auxin.

Step 2: Analyzing the options

- **(1) Plant hormone:** Correct. Auxins are a well-known class of plant hormones. (Note: "harmone" in the option is likely a typo for "hormone").
- **(2) Enzyme:** Enzymes are biological catalysts, usually proteins, that speed up biochemical reactions. Auxins are regulatory molecules, not primarily catalysts.
- **(3) Fat:** Fats (lipids) are organic compounds used for energy storage, insulation, and as structural components of cell membranes. Auxins have a different chemical structure and function.
- **(4) Protein:** Proteins are large, complex molecules made of amino acids, with diverse functions (e.g., enzymes, structural components, some hormones). While some hormones are proteins, auxins are smaller organic molecules, not proteins themselves.

Therefore, auxin is a plant hormone.

Quick Tip

Auxins are key **plant hormones** that control plant growth. Think of them as chemical messengers in plants. They are responsible for things like:

- Making stems grow longer.
- Causing plants to bend towards light.
- Helping roots to form.

Other important plant hormones include gibberellins, cytokinins, abscisic acid, and ethylene.

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131. Asexual reproduction by budding is present in :

(1) Amoeba

(2) Spirogyra

(3) Moss

(4) Yeast

Correct Answer: (4) Yeast

Solution: Concept: Asexual reproduction is a mode of reproduction that does not involve meiosis or fertilization; the offspring arise from a single organism and inherit the genes of that parent only. Budding is one specific type of asexual reproduction.

Step 1: Understanding Budding In budding, a new organism develops from an outgrowth or bud due to cell division at one particular site on the parent organism. The new organism remains attached as it grows, separating from the parent organism only when it is mature. The new individual is genetically identical to the parent.

Step 2: Analyzing reproduction in the given options

- **(1) Amoeba:** Amoeba typically reproduces asexually by **binary fission**, where the parent cell divides into two identical daughter cells. Some species can also undergo multiple fission under unfavorable conditions.
- **(2) Spirogyra:** Spirogyra is a type of filamentous green alga. It reproduces asexually by

fragmentation, where the filament breaks into smaller pieces, and each piece grows into a new filament. It also reproduces sexually by conjugation.

- **(3) Moss:** Mosses (Bryophytes) reproduce both asexually and sexually. Asexual reproduction in mosses can occur by **fragmentation** or by the formation of specialized structures called **gemmae**. Budding can also occur from the protonema stage. However, yeast is a more classic and direct example of budding at the cellular level as its primary mode.

- **(4) Yeast:** Yeast is a unicellular fungus. A common mode of asexual reproduction in yeast is **budding**. A small bud (outgrowth) forms on the parent yeast cell, receives a nucleus, grows, and then detaches to become a new yeast cell.

Step 3: Identifying the organism that primarily uses budding While some forms of budding or bud-like structures might be involved in the life cycle of mosses, **Yeast** is the classic textbook example of an organism that reproduces asexually by budding at the cellular level. Hydra is another common example of an animal that reproduces by budding.
Given the options, Yeast is the most direct and common example of asexual reproduction by budding.

Quick Tip

Budding is like a small ”mini-me” growing off the parent.

- **Yeast** (unicellular fungus): A small bud forms on the parent cell, grows, and separates.

- **Hydra** (simple animal): A bud grows on the body wall and develops into a new hydra.

Other options:

- Amoeba: Binary fission (splits in two).

- Spirogyra: Fragmentation.

132. Example of homologous organs is :

(1) Fore arm of human and wings of bird

(2) Wings of insect and bird

(3) Vermi form appendix and nictilating membrane

(4) Muscles of pinna and tail vertebrae

Correct Answer: (1) Fore arm of human and wings of bird

Solution: Concept: Homologous organs are organs in different species that have a similar basic anatomical structure and embryonic origin, inherited from a common ancestor, but may have evolved to perform different functions. This is evidence of divergent evolution.

Step 1: Defining Homologous and Analogous Organs

- **Homologous Organs:** Similar underlying structure, common evolutionary origin, can have different functions. (Think: Same ancestor, different job).
- **Analogous Organs:** Different underlying structure, different evolutionary origin, but perform similar functions. This is evidence of convergent evolution. (Think: Different ancestor, same job).

Step 2: Analyzing the options

- **(1) Forearm of human and wings of bird:**
  - **Structure:** Both the human forearm and the bird wing have a similar skeletal structure: humerus, radius, ulna, carpals, metacarpals, and phalanges.
  - **Origin:** They share a common ancestral origin from the forelimbs of early vertebrates.
  - **Function:** The human forearm is used for grasping, manipulation, etc. The bird wing is adapted for flight.

  These fit the definition of homologous organs (similar structure, common origin, different functions).

- **(2) Wings of insect and bird:**

  - **Structure:** Insect wings are typically made of chitinous cuticle with veins, lacking bones. Bird wings have an internal skeleton of bones. The underlying structures are very different.
  - **Origin:** They have different evolutionary origins.
  - **Function:** Both are used for flight.

  These are analogous organs (different structure, different origin, similar function).

- **(3) Vermiform appendix and nictitating membrane:** These are generally considered **vestigial organs** in humans. The vermiform appendix is a remnant of a larger cecum used for digesting cellulose in ancestral herbivores. The nictitating membrane (third eyelid) is functional in many animals (birds, reptiles) but is reduced to a small fold (plica semilunaris) in humans. While both are vestigial, comparing them as homologous requires looking at their counterparts in other species. This option is more about vestigial structures than a direct comparison of homologous active organs between two species.
- **(4) Muscles of pinna and tail vertebrae (in humans):** These are also examples of vestigial structures in humans. The muscles of the pinna (external ear) are poorly developed in humans but functional in many animals for ear movement. Tail vertebrae (coccyx) are remnants of a tail. Similar to option (3), this points to vestigial organs.

Step 3: Identifying the best example of homologous organs The forearm of a human and the wing of a bird is a classic example of homologous organs, demonstrating how a common ancestral limb structure has been modified for different functions.

> Quick Tip
>
> **Homo**logous = **Same** basic structure from a common ancestor, but possibly **different functions**. (e.g., Human arm, bird wing, whale flipper, bat wing - all have similar bone patterns). **Ana**logous = **Different** structure and origin, but **similar functions**. (e.g., Wings of a bird and wings of an insect - both for flight, but very different structures). Option (1) is the classic example of homology.

133. Energy stored in respiration :

(1) In the form of ADP

(2) In the form of ATP

(3) In the form of NADP

(4) In the form of PI

Correct Answer: (2) In the form of ATP

Solution: Concept: Cellular respiration is the metabolic process by which organisms break down glucose and other food molecules to release chemical energy. This released energy is then captured and stored in a readily usable form.

Step 1: The Purpose of Cellular Respiration The primary goal of cellular respiration is to convert the chemical energy stored in organic molecules (like glucose) into a form of energy that cells can use to perform various life processes (e.g., muscle contraction, active transport, synthesis of molecules).

Step 2: The Energy Currency of the Cell - ATP The molecule that serves as the main "energy currency" of the cell is **Adenosine Triphosphate (ATP)**. During cellular respiration, energy released from the breakdown of glucose is used to synthesize ATP from Adenosine Diphosphate (ADP) and inorganic phosphate (P_i).

$$ADP + P_i + \text{Energy (from glucose breakdown)} \rightarrow ATP$$

When the cell needs energy, ATP is hydrolyzed (broken down) back into ADP and P_i, releasing the stored energy:

$$ATP \rightarrow ADP + P_i + \text{Usable Energy}$$

Step 3: Analyzing the options

- **(1) In the form of ADP (Adenosine Diphosphate):** ADP is the lower-energy form to which ATP is converted when energy is used. Energy is *used to convert* ADP to ATP, not stored as ADP itself primarily.

- **(2) In the form of ATP (Adenosine Triphosphate):** Correct. ATP is the molecule in which the energy released during respiration is stored and transported for cellular work.

- **(3) In the form of NADP (Nicotinamide Adenine Dinucleotide Phosphate):** $NADP^+$/NADPH is an electron carrier primarily involved in anabolic reactions like photosynthesis (carrying reducing power). While NADH (related to NAD^+) is an electron carrier in respiration whose energy is eventually used to make ATP, NADP itself is not the primary energy storage molecule from respiration.
- **(4) In the form of PI (Inorganic Phosphate):** P_i is a component used to make ATP (from ADP + P_i), but it's not the form in which energy is stored.

Therefore, energy released during respiration is stored in the chemical bonds of ATP.

Quick Tip

Think of **ATP** as the cell's rechargeable battery or energy money.

- Respiration (like burning fuel - glucose) "charges the battery" or "earns money" by converting ADP into ATP.
- When the cell needs energy for work, it "uses the battery" or "spends the money" by converting ATP back to ADP, releasing energy.

So, energy from respiration is stored as ATP.

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134. Example of bisexual flower is :

(1) Papaya

(2) Hibiscus

(3) Cucumber

(4) Maize

Correct Answer: (2) Hibiscus

Solution: Concept: Flowers can be classified based on the presence of male and female reproductive organs.

- **Bisexual flower (or perfect flower):** A flower that has both male reproductive organs (stamens, which produce pollen) and female reproductive organs (pistil/carpel, which contains ovules).

- **Unisexual flower (or imperfect flower):** A flower that has either only male reproductive organs (staminate flower) or only female reproductive organs (pistillate/carpellate flower).

Plants can be:

- **Monoecious:** Have separate male and female flowers on the *same* plant (e.g., maize, cucumber).
- **Dioecious:** Have male flowers on one plant and female flowers on a *different* plant (e.g., papaya).

Step 1: Analyzing the options

- **(1) Papaya (Carica papaya):** Papaya plants are typically dioecious, meaning individual plants bear either male flowers or female flowers (though some cultivars can be monoecious or hermaphroditic). So, individual papaya flowers are usually unisexual.
- **(2) Hibiscus (Hibiscus rosa-sinensis):** Hibiscus flowers are classic examples of bisexual (perfect) flowers. They clearly show both stamens (male parts, forming a staminal tube) and a prominent pistil (female part, with stigma, style, and ovary) within the same flower.
- **(3) Cucumber (Cucumis sativus):** Cucumber plants are typically monoecious. They produce separate male flowers (staminate) and female flowers (pistillate) on the same plant. So, individual cucumber flowers are unisexual.
- **(4) Maize (Zea mays - Corn):** Maize plants are monoecious. The tassel at the top of the plant contains the male (staminate) flowers, and the ears (cobs) contain the female (pistillate) flowers (the silks are the stigmas). So, individual maize flowers are unisexual.

Step 2: Identifying the bisexual flower From the analysis, **Hibiscus** is the example of a plant that typically has bisexual flowers.

Quick Tip

A **bisexual flower** has **both** male (stamens) and female (pistil/carpel) parts in the same flower.

- **Hibiscus:** Clearly shows both sets of organs.
- Papaya: Often has separate male and female plants (dioecious), so flowers are unisexual.
- Cucumber Maize: Have separate male and female flowers on the same plant (monoecious), so individual flowers are unisexual.

---

135. Bowman's capsule is present in :

(1) Small intestine

(2) Kidney

(3) Heart

(4) Brain

Correct Answer: (2) Kidney

Solution: Concept: Bowman's capsule is a specific microanatomical structure involved in the process of filtration in a major excretory organ.

Step 1: Understanding Bowman's Capsule Bowman's capsule (also known as the glomerular capsule) is a cup-shaped sac at the beginning of the tubular component of a **nephron** in the mammalian kidney. The nephron is the basic structural and functional unit of the kidney, responsible for filtering blood and producing urine.

Step 2: Function of Bowman's Capsule Bowman's capsule surrounds a tiny ball of capillaries called the **glomerulus**. Together, the glomerulus and Bowman's capsule form the renal corpuscle. The primary function of the Bowman's capsule is to collect the filtrate that is forced out of the blood in the glomerulus under pressure. This process, called ultrafiltration, is the first step in urine formation. The filtrate (water, glucose, salts, urea, etc.) then passes into the renal tubules of the nephron for further processing (reabsorption and secretion).

Step 3: Analyzing the options

- **(1) Small intestine:** The small intestine is part of the digestive system, primarily involved in digestion and absorption of nutrients. It contains structures like villi and microvilli, but not Bowman's capsules.
- **(2) Kidney:** Correct. Bowman's capsule is a key component of the nephron, the functional unit of the kidney, where blood filtration begins.
- **(3) Heart:** The heart is a muscular organ that pumps blood throughout the circulatory system. Its structures include chambers (atria, ventricles) and valves, but not Bowman's capsules.
- **(4) Brain:** The brain is the control center of the nervous system. It is composed of neurons and glial cells, but not Bowman's capsules.

Therefore, Bowman's capsule is present in the kidney.

> Quick Tip
>
> Bowman's capsule is part of the **nephron**, which is the tiny filtering unit inside the **kidney**. Think: **K**idney $\rightarrow$ **N**ephron $\rightarrow$ Bowman's capsule + Glomerulus (for filtering blood). It's where the first step of urine formation (filtration) happens.

English

136. Fill in the blanks with correct form of the verb :

I finished my home-work while my mother ....the food.

(1) cooked

(2) was cooking

(3) had cooked

(4) cook

Correct Answer: (2) was cooking

Solution: Concept: This question tests the use of appropriate verb tenses to describe actions happening concurrently in the past. The word ”while” often indicates two actions occurring at the same time.

Step 1: Analyze the first part of the sentence ”I finished my home-work” - This is in the simple past tense, indicating a completed action in the past.

Step 2: Understand the role of ”while” ”While” connects two actions that were happening simultaneously, or one action that was ongoing when another shorter action occurred or was completed.

Step 3: Determine the tense for the second action The sentence implies that the mother’s action of cooking was in progress when the speaker finished their homework.

- **Past Continuous Tense (was/were + verb-ing):** This tense is used to describe an ongoing action in the past. ”was cooking” indicates that the mother’s cooking was an ongoing activity.

If one action (finishing homework) happened *during* another longer, ongoing action (mother cooking), the past continuous tense is appropriate for the ongoing action.

Step 4: Evaluating the options

- **(1) cooked (Simple Past):** ”I finished my home-work while my mother cooked the food.” This could imply two sequential completed actions or two actions completed around the same time, but ”was cooking” better emphasizes the ongoing nature of the mother’s activity during which the homework was finished.

- **(2) was cooking (Past Continuous):** ”I finished my home-work while my mother was cooking the food.” This clearly indicates that the mother’s cooking was in progress for some duration, and during that period, the homework was completed. This is the most natural and grammatically fitting choice.

- **(3) had cooked (Past Perfect):** ”I finished my home-work while my mother had cooked the food.” This implies the mother finished cooking *before* the speaker finished their homework, which contradicts the sense of ”while” for simultaneous or overlapping actions here.

- **(4) cook (Simple Present):** This is the wrong tense for a past event.

The best fit is "was cooking" to show an ongoing action in the past interrupted or accompanied by another completed past action.

Quick Tip

When "while" connects two past actions, and one action was ongoing when the other happened, use:

- Simple Past for the shorter, completed action.
- Past Continuous (was/were + verb-ing) for the longer, ongoing action.

Example: "The phone rang **while** I **was taking** a shower." Here: "I finished my homework" (completed) **while** "my mother was cooking" (ongoing).

---

137. Fill in the blanks with correct word :

I prefer milk .... Coffee.

(1) to

(2) than

(3) of

(4) for

Correct Answer: (1) to

Solution: Concept: The verb "prefer" has a specific preposition that is used when comparing two things.

Step 1: Understanding the usage of "prefer" When expressing a preference for one thing over another, the verb "prefer" is followed by the object of preference, then the preposition "to", and then the other item being compared. The structure is: **prefer (something) to (something else)**.

Step 2: Applying the rule to the sentence In the sentence "I prefer milk .... Coffee," milk is the preferred item, and coffee is the item it is being preferred over. According to the rule, the correct preposition to use is "to". So, the sentence becomes: "I prefer milk **to** Coffee."

Step 3: Evaluating the options

- **(1) to:** Correct. This is the standard preposition used with "prefer" for comparisons.
- **(2) than:** "Than" is used with comparative adjectives (e.g., "better than," "taller than") but not typically directly with the verb "prefer" in this structure. While one might say "I like milk more than coffee," with "prefer," "to" is standard.
- **(3) of:** Incorrect preposition in this context.
- **(4) for:** Incorrect preposition in this context.

Quick Tip

When using the verb "prefer" to compare two nouns (things you like), the correct structure is: **prefer X to Y**. Example: "She prefers tea **to** coffee." "He prefers reading **to** watching TV." Don't use "than" with "prefer" in this way.

---

138. Find the correct spelling :

(1) desicion

(2) decision

(3) desision

(4) decicion

Correct Answer: (2) decision

Solution: Concept: This question tests the correct spelling of a common English word.

Step 1: Identify the word intended The options suggest the word is "decision," which means a conclusion or resolution reached after consideration.

Step 2: Analyze the spelling of "decision" The correct spelling of the word is **d-e-c-i-s-i-o-n**. It comes from the verb "decide."

Step 3: Compare with the given options

- **(1) desicion:** Incorrect. The 'c' is missing.
- **(2) decision:** Correct spelling.

- **(3) desision:** Incorrect. The 'c' is missing.
- **(4) decicion:** Incorrect. The 's' is missing.

Therefore, the correct spelling is "decision".

> Quick Tip
>
> The word "decision" is related to the verb "decide." Remember the "ci" combination after "de-" and the "-sion" ending. **de-ci-sion**. Practice spelling common words to improve accuracy.

---

139. The correct narrated form of the given sentence is :
"I will go to Delhi tomorrow," he told me.
(1) He told me that I would go to Delhi tomorrow.
(2) He told me that he will go to Delhi the next day.
(3) He told me that he would go to Delhi the next day.
(4) He told me that he would go to Delhi tomorrow.

Correct Answer: (3) He told me that he would go to Delhi the next day.

Solution: Concept: This question requires converting a sentence from direct speech to indirect (reported or narrated) speech. This involves several changes: reporting verb, conjunction, pronouns, tense of the verb in the reported clause, and adverbs of time/place.
Step 1: Identify the components of the direct speech

- Reporting verb: "told me"
- Speaker of reported clause: "I" (refers to "he")
- Verb in reported clause: "will go"
- Adverb of time: "tomorrow"

Step 2: Rules for converting to indirect speech

1. **Reporting Verb:** If the reporting verb is in the past tense (like ”told”), the tense of the verb in the reported speech usually changes. ”told me” remains ”told me” (or could be ”said to me” becoming ”told me”).

2. **Conjunction:** Use ”that” to introduce the reported clause (though it can sometimes be omitted).

3. **Pronouns:** Change pronouns according to the speaker and listener. ”I” (referring to ”he”) becomes ”he”.

4. **Tense Change (if reporting verb is past):**

    - ”will” changes to ”would”.

5. **Adverbs of Time/Place:**

    - ”tomorrow” changes to ”the next day” or ”the following day”.

Step 3: Apply the rules Direct: ”I will go to Delhi tomorrow,” he told me.

- Reporting verb ”told me” remains.
- Add ”that”.
- ”I” (referring to ”he”) becomes ”he”.
- ”will go” becomes ”would go”.
- ”tomorrow” becomes ”the next day”.

Indirect: He told me that he would go to Delhi the next day.

Step 4: Compare with the options

- **(1) He told me that I would go to Delhi tomorrow.** (Pronoun ”I” incorrect; ”tomorrow” not changed)
- **(2) He told me that he will go to Delhi the next day.** (Tense ”will go” incorrect, should be ”would go”)
- **(3) He told me that he would go to Delhi the next day.** (All changes correctly applied)
- **(4) He told me that he would go to Delhi tomorrow.** (”tomorrow” not changed)

Option (3) is the correctly narrated form.

Quick Tip

Key changes for indirect speech (when reporting verb is past):

- Pronouns change (e.g., I $\rightarrow$ he/she).
- Verb tense "backshifts" (e.g., will $\rightarrow$ would, simple present $\rightarrow$ simple past).
- Time/place words change (e.g., tomorrow $\rightarrow$ the next day, now $\rightarrow$ then, here $\rightarrow$ there).

Direct: "I **will** go **tomorrow**." Indirect (He told me that...): "...he **would** go **the next day**."

---

140. Fill in the blank with correct word :

The Principal is giving ....the prizes.

(1) into

(2) away

(3) for

(4) at

Correct Answer: (2) away

Solution: Concept: This question tests the correct use of phrasal verbs. A phrasal verb is a combination of a verb and a preposition or adverb (or both) that creates a meaning different from the original verb.

Step 1: Understand the context The sentence describes the action of a Principal distributing prizes. We need a phrasal verb with "give" that means to distribute or hand out.

Step 2: Meaning of "give away" The phrasal verb **give away** means:

1. To give something to someone for free.
2. To distribute something (like prizes or awards).
3. To reveal a secret or information.

In the context of prizes, ”give away” means to distribute them to the winners.

Step 3: Evaluating the options

- **(1) giving into:** ”Give into” means to yield or surrender to something (e.g., ”He gave into temptation.”). This doesn’t fit the context of distributing prizes.
- **(2) giving away:** ”The Principal is giving away the prizes.” This correctly means the Principal is distributing the prizes.
- **(3) giving for:** ”Give for” is not a standard phrasal verb with a meaning suitable here. One might ”give something for a reason,” but ”giving for the prizes” is awkward.
- **(4) giving at:** ”Give at” is not a standard phrasal verb. One might ”give something at a location,” but it doesn’t fit the action of distribution here.

The most appropriate phrasal verb is ”giving away.”

Quick Tip

The phrasal verb ”give away” often means to distribute things, especially for free or as awards. Example: ”The store is **giving away** free samples.” ”The teacher will **give away** certificates to students with perfect attendance.” In this sentence, ”The Principal is **giving away** the prizes” means distributing them.

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141. Fill in the blank with correct word :

Can I give you some . . . . ?

(1) advise

(2) advised

(3) advising

(4) advice

Correct Answer: (4) advice

Solution: Concept: This question tests the difference between ”advice” (noun) and ”advise” (verb), which are often confused.

Step 1: Understanding ”advice” vs. ”advise”

- **Advice (noun, ends with -ice, pronounced /dvas/ with an ’s’ sound at the end):** This is an opinion or recommendation offered as a guide to action, conduct, etc. It is something that can be given or received. ”Some advice” implies a quantity of this noun. Example: ”He gave me some good *advice*.” ”I need your *advice*.”
- **Advise (verb, ends with -ise, pronounced /dvaz/ with a ’z’ sound at the end):** This means to offer an opinion or suggestion as worth following; to recommend. It is an action. Example: ”I *advise* you to be cautious.” ”The doctor *advised* him to rest.”

Step 2: Analyzing the sentence structure The sentence ”Can I give you some ....?” requires a noun after ”some.” ”Some” is a determiner that typically precedes a noun (or an adjective + noun). We are looking for the ”thing” that can be given.

Step 3: Evaluating the options

- **(1) advise (verb):** Incorrect. ”Some advise” is grammatically incorrect as ”advise” is a verb.
- **(2) advised (past tense/past participle of the verb ”advise”):** Incorrect. This is a verb form.
- **(3) advising (present participle/gerund of the verb ”advise”):** Incorrect. This is a verb form.
- **(4) advice (noun):** Correct. ”Some advice” means a certain amount of guidance or recommendation. ”Can I give you some advice?” is a grammatically correct and meaningful question.

Quick Tip

Easy way to remember:

- **Advic<u>e</u> (noun):** Think of ”ic<u>e</u>” which is a *thing* (noun). Advice is a thing you give or get.
- **Advis<u>e</u> (verb):** This is an *action*. You advis<u>e</u> someone.

The sentence needs a noun after ”some”: ”some **advice**”.

142. Find the correct auxiliary verb for the following sentence :

One or the other of those fellows ....stolen the car.

(1) has

(2) have

(3) has been

(4) have been

Correct Answer: (1) has

Solution: Concept: This question tests subject-verb agreement, specifically with phrases like "one or the other." The verb must agree with the true subject of the sentence.

Step 1: Identify the subject of the sentence The phrase "One or the other of those fellows" is the subject. When "one" is used in such a construction (e.g., "one of the boys," "one or the other"), the true subject is "one," which is singular. The phrase "of those fellows" is a prepositional phrase modifying "one or the other."

Step 2: Determine the number of the subject Since "one" is singular, the subject "One or the other" is considered singular.

Step 3: Choose the auxiliary verb that agrees with a singular subject in the present perfect tense The main verb is "stolen," which is the past participle of "steal." This indicates the sentence is likely in the perfect tense (e.g., present perfect or past perfect). The options provided suggest present perfect. The present perfect tense is formed with "has/have + past participle."

- For singular subjects (he, she, it, one), use "has."
- For plural subjects (they, we) and for "I" and "you," use "have."

Since the subject "One or the other" is singular, the correct auxiliary verb is "has."

Step 4: Evaluating the options

- **(1) has:** Correct. "One or the other ... **has** stolen the car." (Present Perfect tense, active voice)
- **(2) have:** Incorrect. "have" is used with plural subjects.

- **(3) has been:** This would form "has been stolen." If the car "has been stolen" (by one of them), the sentence structure would be different (passive voice: "The car has been stolen by one or the other..."). Here, "one or the other" is performing the action (active voice).
- **(4) have been:** Incorrect, plural and also suggests passive voice in this context.

The sentence is in the active voice: one of the fellows performed the action of stealing.

Quick Tip

Phrases like:

- "One of the [plural noun]"
- "Each of the [plural noun]"
- "Either of the [plural noun]"
- "Neither of the [plural noun]"
- "One or the other of the [plural noun]"

take a **singular verb** because the true subject is "one," "each," "either," or "neither," all of which are singular. So, "One ... **has** stolen."

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143. Choose the sentence which is grammatically correct :

(1) Politics are not meant for students.

(2) New Delhi is a capital of India.

(3) Many a man has done so.

(4) The novelist and poet are dead.

Correct Answer: (3) Many a man has done so.

Solution: Concept: This question tests various rules of English grammar, including subject-verb agreement and the use of articles.

Step 1: Analyze each sentence for grammatical correctness

- **(1) Politics are not meant for students.** The word "politics," when referring to the art or science of government or political affairs, is generally treated as a singular noun, even though it ends in 's'. Therefore, it should take a singular verb: "Politics **is** not meant for students." (If "politics" refers to specific political beliefs or activities, it can sometimes be plural, but in this general statement, singular is standard.) So, this sentence is likely incorrect.

- **(2) New Delhi is a capital of India.** India has only one capital city, which is New Delhi. When referring to a unique position or title, the definite article "the" is usually used, or no article if it's treated as a proper noun phrase in a specific way. "New Delhi is **the** capital of India" is correct. "New Delhi is **capital** of India" (less common but possible if "capital of India" is seen as a title). "A capital" implies one of several, which is not the case here. So, this sentence is incorrect due to the article "a".

- **(3) Many a man has done so.** The construction "Many a [singular noun]" is a formal and somewhat idiomatic expression. Despite "many" suggesting plurality, this specific phrase takes a **singular verb**. "Many a man" (singular subject) $\rightarrow$ "has done" (singular verb). This sentence is grammatically correct.

- **(4) The novelist and poet are dead.** When two singular nouns are joined by "and":
  - If they refer to *two different people*, and the article "the" (or a possessive pronoun) is used before each noun (e.g., "The novelist and **the** poet..."), a plural verb is used ("are").
  - If they refer to the *same person* who has two roles/titles, and the article "the" is used only before the first noun (e.g., "The novelist and poet..."), a singular verb is used ("is").

  In this sentence, "The novelist and poet" implies one person who is both a novelist and a poet (article "the" is only before "novelist"). Therefore, it should take a singular verb: "The novelist and poet **is** dead." If it meant two different people, it would be "The novelist and *the* poet are dead." So, this sentence, as written, is incorrect.

Step 2: Identifying the grammatically correct sentence Based on the analysis, "Many a man has done so." is the grammatically correct sentence.

Quick Tip

Key grammar points:

- "Politics" (as a subject of study/affairs) is usually singular: "Politics **is**..."
- Unique titles/positions often use "the": "New Delhi is **the** capital..."
- "Many a [singular noun]" takes a **singular verb**: "Many a man **has**..."
- When two nouns joined by "and" refer to the **same person** (article only before the first), use a singular verb: "The novelist and poet **is**..." (meaning one person).

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144. Find the correct antonym to the underlined word :

There was a marked <u>improvement</u> in his condition.

(1) reformation

(2) amendment

(3) deterioration

(4) revision

Correct Answer: (3) deterioration

Solution: Concept: An antonym is a word that has the opposite meaning to another word. The underlined word is "improvement."

Step 1: Understand the meaning of "improvement" "Improvement" means the act of making something better, or a change or development that makes something better. It implies a positive change or progress.

Step 2: Look for a word with the opposite meaning We need a word that means the act of becoming worse, or a negative change.

Step 3: Analyzing the options

- **(1) reformation:** The act of reforming; improvement or amendment of what is wrong, corrupt, unsatisfactory, etc. This is a synonym or closely related to improvement, not an antonym.
- **(2) amendment:** A minor change or addition designed to improve a text, piece of

legislation, etc. It generally implies an improvement or correction. Not an antonym.

- **(3) deterioration:** The process of becoming progressively worse. This is the direct opposite of improvement. If a condition deteriorates, it gets worse.
- **(4) revision:** The act of revising; a corrected or improved version. This often implies an improvement or update. Not an antonym.

Step 4: Identifying the correct antonym The word that means the opposite of ”improvement” (getting better) is ”deterioration” (getting worse).

Quick Tip

Improvement = Getting better. **Antonym** = Opposite meaning. We need a word for ”getting worse.”

- Reformation, amendment, revision: All suggest making things better or changing them, often for the better.
- **Deterioration:** Means becoming worse. This is the clear antonym.

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145. Choose the correct narrated form for the following sentence :
He said to me, ”Where are you going?”

(1) He says where I am going.

(2) He told me where I was going.

(3) He asked me where they are going.

(4) He asked me where I was going.

Correct Answer: (4) He asked me where I was going.

Solution: Concept: This question requires converting an interrogative (question) sentence from direct speech to indirect (reported) speech.

Step 1: Identify the components of the direct speech

- Reporting verb: ”said to me”
- Type of sentence: Interrogative (Wh-question: ”Where...”)

- Verb in reported clause: ”are going” (Present Continuous)
- Pronoun in reported clause: ”you” (refers to ”me,” the listener)

Step 2: Rules for converting interrogative sentences to indirect speech

1. **Reporting Verb:** ”said to” changes to ”asked” (or ”enquired of,” ”wondered,” etc., but ”asked” is most common).
2. **Conjunction:**
   - For Wh-questions (who, what, where, when, why, how), the Wh-word itself acts as the conjunction. Do not use ”that.”
   - For Yes/No questions, use ”if” or ”whether.”
3. **Sentence Structure:** The interrogative form of the reported speech changes to an assertive (statement) form. This means the subject comes before the verb. (Question: ”Where <u>are you</u> going?” $\rightarrow$ Statement form: ”...where <u>I was</u> going.”)
4. **Pronouns:** Change according to the speaker and listener. ”you” (referring to ”me”) becomes ”I”.
5. **Tense Change (if reporting verb is past, like ”said to”):**
   - Present Continuous (”are going”) changes to Past Continuous (”was going”).
6. **Question Mark:** The question mark is removed and replaced with a full stop (period).

Step 3: Apply the rules Direct: He said to me, ”Where are you going?”

- ”said to me” becomes ”asked me”.
- Wh-word ”Where” acts as the conjunction.
- ”you” (referring to ”me”) becomes ”I”.
- ”are going” (Present Continuous) becomes ”was going” (Past Continuous).
- The structure becomes assertive: ”where I was going”.

Indirect: He asked me where I was going.

Step 4: Compare with the options

- **(1) He says where I am going.** (Reporting verb "says" incorrect tense; verb "am going" incorrect tense)
- **(2) He told me where I was going.** (Reporting verb "told" is usually for statements, "asked" is better for questions)
- **(3) He asked me where they are going.** (Pronoun "they" incorrect; verb "are going" incorrect tense)
- **(4) He asked me where I was going.** (All changes correctly applied)

Option (4) is the correctly narrated form.

Quick Tip

For Wh-questions in indirect speech (when reporting verb is past): 1. Reporting verb $\rightarrow$ "asked". 2. Wh-word (Where, What, etc.) stays and acts as a connector. 3. Change the question part into a statement (subject before verb). 4. Pronouns change (e.g., you $\rightarrow$ I/he/she/they). 5. Verb tense "backshifts" (e.g., present continuous $\rightarrow$ past continuous). Direct: "Where **are you** going?" Indirect (He asked me...): "...where **I was** going."

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146. Choose the right option for the synonym of the underlined word :

We must <u>eradicate</u> corruption.

(1) minimise

(2) uproot

(3) condemn

(4) control

Correct Answer: (2) uproot

Solution: Concept: A synonym is a word or phrase that means exactly or nearly the same as another word or phrase in the same language. The underlined word is "eradicate."

Step 1: Understand the meaning of "eradicate" "Eradicate" means to destroy completely; to put an end to; to eliminate or get rid of something, especially something bad or undesirable, as if by pulling it up by the roots. The origin of the word is from Latin "eradicare," meaning "to root out" (e- "out" + radix "root").

Step 2: Look for a word with a similar meaning We need a word that conveys the idea of complete removal or destruction.

Step 3: Analyzing the options

- **(1) minimise:** To reduce to the smallest possible amount or degree. This implies reduction, not complete elimination.
- **(2) uproot:** To pull (a plant, etc.) up by the roots; to remove or destroy completely. This meaning is very close to "eradicate," especially considering its etymology.
- **(3) condemn:** To express complete disapproval of, typically in public; to censure. This is about expressing disapproval, not about removal or destruction.
- **(4) control:** To determine the behaviour or supervise the running of; to maintain influence or authority over. This implies management or limitation, not necessarily complete elimination.

Step 4: Identifying the best synonym "Uproot" shares the core meaning of complete removal from the very foundation, similar to "eradicate." Other strong synonyms for eradicate include eliminate, abolish, extinguish, wipe out. Among the choices, "uproot" is the closest in meaning and imagery.

Quick Tip

To **eradicate** something means to get rid of it completely, as if pulling it out by its roots.

- **Uproot** literally means to pull up by the roots, and figuratively means to remove completely. This is a strong synonym.
- Minimise = make smaller.
- Condemn = strongly disapprove.
- Control = manage or limit.

"Eradicate" and "uproot" both imply a thorough and complete removal.

147. Choose the correct passive form of the following sentence :
I have already seen this movie.
(1) This movie is alredy seen by me.
(2) This movie has already seen by me.
(3) This movie was already seen by me.
(4) This movie has already been seen by me.

Correct Answer: (4) This movie has already been seen by me.

Solution: Concept: Converting a sentence from active voice to passive voice involves changing the focus from the doer of the action (subject) to the recipient of the action (object). The verb form also changes.
Step 1: Identify components of the active sentence

- Subject (doer): I
- Verb: have seen (Present Perfect tense)
- Adverb: already
- Object (recipient): this movie

Active sentence: Subject + Verb (have + V3) + Object $\rightarrow$ I + have already seen + this movie.
Step 2: Rules for converting to Passive Voice (Present Perfect Tense)

1. The object of the active sentence becomes the subject of the passive sentence. (Active object: ”this movie” $\rightarrow$ Passive subject: ”This movie”)
2. The active verb form ”have/has + V3 (past participle)” changes to ”have/has + been + V3”. (”have seen” $\rightarrow$ ”has been seen” - ”has” because ”This movie” is singular)
3. The adverb ”already” is usually placed between the auxiliary verb ”has/have” and ”been”.
4. The subject of the active sentence becomes the agent in the passive sentence, usually introduced by ”by”. (Active subject: ”I” $\rightarrow$ Passive agent: ”by me”)

Passive structure: New Subject + has/have + (adverb) + been + V3 + by + Agent.

Step 3: Apply the rules Active: I have already seen this movie.

- New subject: This movie
- Auxiliary for singular subject ("This movie"): has
- Adverb: already
- Add "been".
- Past participle of "see": seen
- Agent: by me

Passive: This movie has already been seen by me.

Step 4: Compare with the options

- **(1) This movie is already seen by me.** (Incorrect tense. "is seen" is simple present passive.)
- **(2) This movie has already seen by me.** (Missing "been" for present perfect passive.)
- **(3) This movie was already seen by me.** (Incorrect tense. "was seen" is simple past passive.)
- **(4) This movie has already been seen by me.** (Correct present perfect passive form.)

Quick Tip

Active (Present Perfect): Subject + has/have + V3 + Object. Passive (Present Perfect): Object (as new Subject) + has/have + been + V3 + by + Agent. Active: I **have seen** this movie. Passive: This movie **has been seen** by me. The adverb "already" fits in nicely: "This movie **has already been seen** by me."

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148. Fill in the blank with the correct word :

Security arrangements have been tightened up in all .... areas.

(1) sensible

(2) sensual

(3) sensitive

(4) sensational

Correct Answer: (3) sensitive

Solution: Concept: This question requires choosing the adjective that best fits the context of ”security arrangements” and ”areas” where they might be tightened.

Step 1: Understand the context The sentence talks about tightening security arrangements. This is usually done in areas that are vulnerable, important, or prone to issues.

Step 2: Meanings of the adjectives in the options

- **(1) sensible:** Having or showing good sense or judgment; practical and functional rather than decorative. (e.g., ”a sensible decision,” ”sensible shoes”).
- **(2) sensual:** Relating to or gratifying the physical senses, especially sexual senses. (e.g., ”sensual pleasure,” ”sensual lips”).
- **(3) sensitive:**
  1. Quick to detect or respond to slight changes, signals, or influences.
  2. Easily damaged, injured, or distressed by slight changes.
  3. Requiring careful handling due to its delicate, secret, or controversial nature; potentially volatile. (e.g., ”sensitive information,” ”a sensitive topic,” ”sensitive equipment,” ”a sensitive area/zone”).
- **(4) sensational:** Causing or trying to cause great public interest and excitement, often by exaggerating or distorting details. (e.g., ”sensational headlines,” ”a sensational trial”).

Step 3: Choose the best fit for ”areas” requiring tightened security

- ”Sensible areas”: Doesn't fit well. Areas are not usually described as sensible in this context.
- ”Sensual areas”: Completely inappropriate for security context.

- **”Sensitive areas”:** This fits perfectly. ”Sensitive areas” are locations that are vulnerable, important, or require special protection due to potential risks (e.g., border areas, government buildings, areas prone to conflict). Security is often tightened in such areas.
- ”Sensational areas”: Doesn’t fit. ”Sensational” relates to causing excitement or shock, not directly to the need for security in this manner.

Although both ”sensitive” and ”sensational” were circled in the image, ”sensitive” is the standard and most appropriate adjective in this context. Security is tightened in areas that are considered ”sensitive” due to their nature or potential for problems.

Quick Tip

When talking about security and places that need protection, the word ”sensitive” is often used. A ”sensitive area” is a place that is vulnerable, important, or potentially dangerous, thus requiring heightened security.

- **Sensitive documents/information** need protection.
- **Sensitive equipment** needs careful handling.
- **Sensitive areas/zones** need tight security.

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149. Choose the correct option for the following sentence :

If I had worked hard, I .... very high marks in the examination.

(1) would have scored

(2) scored

(3) would score

(4) could score

Correct Answer: (1) would have scored

Solution: Concept: This sentence is a conditional sentence, specifically a **Type 3 Conditional** (also known as the third conditional or past unreal conditional). It describes a hypothetical situation in the past that did not happen, and its hypothetical past result.

Step 1: Structure of a Type 3 Conditional Sentence A Type 3 conditional sentence has two clauses:

1. **The 'if' clause (condition):** Describes the unreal past condition. Structure: **If + Past Perfect tense (had + past participle)** Example: "If I *had worked* hard..."

2. **The main clause (result):** Describes the unreal past result. Structure: **Subject + would have + past participle** (Other modal verbs like "could have" or "might have" can also be used depending on the nuance).

Step 2: Apply the structure to the given sentence The 'if' clause is "If I had worked hard," which is in the Past Perfect tense (*had worked*). This confirms it's a Type 3 conditional. Therefore, the main clause should use "would have + past participle" (or "could have / might have + past participle"). The verb is "score," and its past participle is "scored." So, the main clause should be "I **would have scored** very high marks..."

Step 3: Evaluating the options

- **(1) would have scored:** Correct. This fits the structure of the main clause in a Type 3 conditional.

- **(2) scored (Simple Past):** Incorrect. This doesn't fit the unreal past conditional structure.

- **(3) would score (would + base verb):** Incorrect. This is used in Type 2 conditionals (unreal present/future).

- **(4) could score (could + base verb):** Incorrect. This is used in Type 2 conditionals. For Type 3, it would be "could have scored."

The correct completion is "would have scored."

Quick Tip

Type 3 Conditional sentences talk about **unreal past situations** and their **unreal past results**. Structure: **If + had + V3 (past participle)**, ... **subject + would have + V3 (past participle)**. Example: "If she **had studied** (Past Perfect), she **would have passed** (would have + V3) the exam." (Meaning: She didn't study, so she didn't pass.) In your sentence: "If I **had worked** hard, I **would have scored**..."

150. Choose the right option which best expresses the meaning of the underlined idiom : The luxury car that they bought turned out to be a white elephant.

(1) a rare article

(2) useful mode of transport

(3) costly or trouble some possession

(4) a proud possession

Correct Answer: (3) costly or troublesome possession

Solution: Concept: An idiom is a phrase or expression whose meaning cannot be deduced from the literal meanings of its constituent words. "A white elephant" is a common English idiom.

Step 1: Understanding the idiom "a white elephant" The idiom "a white elephant" refers to a possession that is:

- Expensive to maintain or upkeep.
- Difficult or troublesome to manage.
- Ultimately of little use or value, or its cost and trouble outweigh its benefits.
- Often, it's something that is not easy to get rid of.

The origin is said to come from the practice of kings of Siam (now Thailand) supposedly giving actual white elephants (which were considered sacred and could not be put to work) as "gifts" to courtiers they disliked. The recipient would be financially ruined by the cost of maintaining the sacred animal.

Step 2: Applying the meaning to the sentence "The luxury car that they bought turned out to be a white elephant." This implies that the luxury car, while perhaps initially desirable, became a burden – possibly expensive to run, maintain, insure, or repair, and perhaps not as useful as expected, making its overall cost and trouble greater than its worth to them.

Step 3: Analyzing the options

- **(1) a rare article:** While a white elephant itself is rare, the idiom doesn't primarily mean rarity. It means something burdensome.

- **(2) useful mode of transport:** This is the opposite of what a white elephant usually implies. If it were useful without being a burden, it wouldn't be a white elephant.

- **(3) costly or troublesome possession:** This perfectly captures the meaning of ”a white elephant.” It's a possession that costs a lot (either to acquire or maintain) and/or causes a lot of trouble, and may not be worth it.

- **(4) a proud possession:** While one might initially be proud of a luxury car, if it becomes a ”white elephant,” the pride is likely overshadowed by the burden. The idiom focuses on the negative aspects of the possession.

Step 4: Identifying the best expression of meaning Option (3) ”costly or troublesome possession” best expresses the meaning of the idiom ”a white elephant.” (Note: ”trouble some” in the option should ideally be ”troublesome”).

Quick Tip

An idiom is a phrase with a figurative meaning. ”A white elephant” is something that is:

- Expensive to own or maintain.
- More trouble than it's worth.
- Often not very useful despite its cost or grandeur.

Think of a gift that looks impressive but costs you a lot to keep and you can't really use it much – that's a white elephant.
