Advanced Theory of Statistics

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1 Probability Theory

1.1 Measure space, measurable function, and integration

Definition 1: A collection of subsets of $\Omega, \mathscr{F}$, is a σ-field (or σ-algebra) if (i) The empty set $\emptyset \in \mathscr{F}$; (ii) If $A \in \mathscr{F}$, then the complement $A^c \in \mathscr{F}$; (iii) If $A_i \in \mathscr{F}, i = 1, 2, \cdots$, then their union $\cup A_i \in \mathscr{F}$. $(\Omega, \mathscr{F})$ is a measurable space if $\mathscr{F}$ is a σ-field on Ω.

Example 1: $\mathscr{C}$ = a collection of subsets of interest. $\sigma(\mathscr{C})$ = the smallest σ-field containing $\mathscr{C}$ (the σ-field generated by $\mathscr{C}$). $\sigma(\mathscr{C}) = \mathscr{C}$ if $\mathscr{C}$ itself is a σ-field. $\sigma(\{A\}) = \{\emptyset, A, A^c, \Omega\}$.

Example 2 (Borel σ-field): $\mathbb{R}^k$: the k-dimensional Euclidean space ($\mathbb{R}^1 = \mathbb{R}$ is the real line). $\mathscr{O}$ = all open sets, $\mathscr{C}$ = all closed sets. $\mathscr{B}^k = \sigma(\mathscr{O}) = \sigma(\mathscr{C})$: the Borel σ-field on $\mathbb{R}^k$. $C \in \mathscr{B}^k, \mathscr{B}_C = \{C \cap B : B \in \mathscr{B}^k\}$ is the Borel σ-field on C.

Definition 2: Let $(\Omega, \mathscr{F})$ be a measurable space. A set function ν defined on $\mathscr{F}$ is a measure if (i) $0 \le \nu(A) \le \infty$ for any $A \in \mathscr{F}$; (ii) $\nu(\emptyset) = 0$; (iii) If $A_i \in \mathscr{F}, i = 1, 2, \cdots$, and A_i's are disjoint, i.e. $A_i \cap A_j = \emptyset$ for any $i \neq j$, then $\nu\left(\cup_{i=1}^{\infty} A_i\right) = \sum_{i=1}^{\infty} \nu(A_i)$. $(\Omega, \mathscr{F}, \nu)$ is a measure if ν is a measure on $\mathscr{F}$ in $(\Omega, \mathscr{F})$.

Convention 1: For any $x \in \mathbb{R}$, $\infty + x = \infty$, $x\infty = \infty$ if $x > 0$, $x\infty = -\infty$ if $x < 0$. $0\infty = 0$, $\infty + \infty = \infty$, $\infty^a = \infty$ for any $a > 0$. $\infty - \infty$ or ∞/∞ is not defined.

Example 3 (Important examples of measures): (a) Let $x \in \Omega$ be a fixed point and $\delta_x(A) = \begin{cases} c & x \in A \\ 0 & x \notin A \end{cases}$. This is called a point mass at x. (b) Let $\mathscr{F}$ = all subsets of Ω and $\nu(A)$ = the number of elements in $A \in \mathscr{F}$ ($\nu(A) = \infty$ if A contains infinitely many elements). Then ν is a measure on $\mathscr{F}$ and is called the counting measure. (c) There is a unique measure m on $(\mathbb{R}, \mathscr{B})$, that satisfies $m([a, b]) = b - a$ for every finite interval $[a, b]$, $-\infty < a \le b < \infty$. This is called the Lebesgue measure.

Proposition 1 (Properties of measures): Let $(\Omega, \mathscr{F}, \nu)$ be a measure space. (1) Monotonicity: If $A \subset B$, then $\nu(A) \subset \nu(B)$. (2) Subadditivity: For any sequence $A_1, A_2, \cdots,$, $\nu\left(\cup_{i=1}^{\infty} A_i\right) \le \sum_{i=1}^{\infty} \nu(A_i)$. (3) Continuity: If $A_1 \subset A_2 \subset A_3 \subset \cdots$ (or $A_1 \supset A_2 \supset A_3 \supset \cdots$ and $\nu(A_1) < \infty$), then $\nu(\lim_{n\to\infty} A_n) = \lim_{n\to\infty} \nu(A_n)$ where $\lim_{n\to\infty} AA_n = \cup_{i=1}^{\infty} A_i$ (or $= \cap_{i=1}^{\infty} A_i$).

Definition 3: Let P be a probability measure on $(\mathbb{R}, \mathscr{B})$. The cumulative distribution function (c.d.f.) of P is defined to be $F(x) = P((-infty, x])$, $x \in \mathbb{R}$.

Proposition 2 (Properties of c.d.f.'s): (i) Let F be a c.d.f. on $\mathbb{R}$. (a) $F(-\infty) = \lim_{x\to-\infty} F(x) = 0$; (b) $F(\infty) = \lim_{x\to\infty} F(x) = 1$; (c) F is nondecreasing, i.e. $F(x) \le F(y)$ if $x \le y$; (d) F is right continuous, i.e. $\lim_{y\to x+0} F(y) = F(x)$. (ii) Suppose a real-valued function F on $\mathbb{R}$ satisfies (a)-(d) in part (i). Then F is the c.d.f. of a unique probability meausre on $(\mathbb{R}, \mathscr{B})$.

Definition 4 (Product space): $\mathscr{I} = \{1, \cdots, k\}$, k is finite or ∞. $\Gamma_i, i \in \mathscr{I}$, are some sets. $\prod_{i\in\mathscr{I}} \Gamma_i = \Gamma_1 \times \cdots \times \Gamma_k = \{(a_1, \cdots, a_k) : a_i \in \Gamma_i, i \in \mathscr{I}\}$. Let $(\Omega_i, \mathscr{F}_i), i \in \mathscr{I}$ be measurable spaces. $\sigma(\prod_{i\in\mathscr{I}} \mathscr{F}_i)$ is called the product σ-field on the product space $\prod_{i\in\mathscr{I}} \Omega_i$. $(\prod_{i\in\mathscr{I}} \Omega_i, \sigma(\prod_{i\in\mathscr{I}} \mathscr{F}_i))$ is denoted by $\prod_{i\in\mathscr{I}}(\Omega_i, \mathscr{F}_i)$.

Definition 5 (σ-finite): A measure ν on $(\Omega, \mathscr{F})$ is said to be σ-finite iff there exists a sequence $\{A_1, A_2, \cdots\}$ such that $\cup A_i = \Omega$ and $\nu(A_i) < \infty$ for all i. Any finite measure is clearly σ-finite. The Lebesgue measure on $\mathscr{F}$ is σ-finite.

Proposition 3 (Product measure theorem): Let $(\Omega_i, \mathscr{F}_i, \nu_i), i = 1, \cdots, k$, be measure spaces with σ-finite measures. There exists a unique σ-finite measure on σ-field $\sigma(\mathscr{F}_1 \times \cdots \times \mathscr{F}_k)$, called the product measure and denoted by $\nu_1 \times \cdots \times \nu_k$, such that $\nu_1 \times \cdots \times \nu_k(A_1 \times \cdots \times A_k) = \nu_1(A_1) \cdots \nu_k(A_k)$ for all $A_i \in \mathscr{F}_i, i = 1, \cdots, k$.

Definition 6 (Measurable function): Let $(\Omega, \mathscr{F})$ and $(\Lambda, \mathscr{G})$ be measurable spaces. Let f be a function from Ω to Λ. f is called a measurable function from $(\Omega, \mathscr{F})$ to $(\Lambda, \mathscr{G})$ iff $f^{-1}(\mathscr{G}) \subset \mathscr{F}$.

Definition 7 (Integration): (a) The integral of a nonnegative simple function ϕ w.r.t.ν is defined as $\int \phi d\nu = \sum_{i=1}^k a_i \nu(A_i)$. (b) Let f be a nonnegative Borel function and let $\mathscr{S}_f$ be the collection of all nonnegative simple functions satisfying $\phi(\omega) \leq f(\omega)$ for any $\omega \in \Omega$. The integral of f w.r.t. ν is defined as $\int f d\nu = \sup\{\int \phi d\nu : \phi \in \mathscr{S}_f\}$ (Hence, for any Borel function $f \geq 0$, there exists aa sequence of simple functions $\phi_1, \phi_2, \cdots$ such that $0 \leq \phi_i \leq f$ for all i and $\lim_{n\to\infty} \int \phi_n d\nu = \int f d\nu$). (c) Let f be a Borel function, $f_+(\omega) = \max\{f(\omega), 0\}$ be the positive part of f, and $f_-(\omega) = \max\{-f(\omega), 0\}$ be the negative part of f. We say that $\int f d\nu$ exists if and only if at least one of $\int f_+ d\nu$ and $\int f_- d\nu$ is finite, in which case $\int f d\nu = \int f_+ d\nu - \int f_- d\nu$. (d) When both $\int f_+ d\nu$ and $\int f_- d\nu$ are finite, we say that f is integrable. Let A be a measurable set and I_A be its indicator function. The integral of f over A is defined as $\int_A f d\nu = \int I_A f d\nu$.

Example 4 (Extended set): For convenience, we define the integral of a measurable f from $(\Omega, \mathscr{F}, \nu)$ to $(\bar{\mathbb{R}}, \bar{\mathscr{B}})$, where $\bar{\mathbb{R}} = \mathbb{R} \cup \{-\infty, +\infty\}, \bar{\mathscr{B}} = \sigma(\mathscr{B} \cup \{\infty, -\infty\})$. Let $A_+ = \{f = \infty\}$ and $A_- = \{f = -\infty\}$. If $\nu(A_+) = 0$, we define $\int f_+ d\nu$ to be $\int I_{A_+^c} f_+ d\nu$; otherwise $\int f_+ d\nu = \infty$. $\int f_- d\nu$ is similarly defined. If at least one of $\int f_+ d\nu$ and $\int f_- d\nu$ is finite, then $\int f d\nu = \int f_+ d\nu - \int f_- d\nu$ is well defined.

1.2 Integration theory and Radon-Nikodym derivative

Proposition 1: $(\Omega, \mathscr{F}, \nu)$ be a measure space and f and g be Borel functions. (i) If $f \leq g$ a.e., then $\int f d\nu \leq \int g d\nu$, provided that the itegrals exist. (ii) If $f \geq 0$ a.e. and $\int f d\nu = 0$, then $f = 0$ a.e.

Theorem 1: Let $f_1, f_2, \cdot$ be a sequence of Borel functions on $(\Omega, \mathscr{F}, \nu)$. (i) Fatou's lemma: If $f_n \geq 0$, then $\int \liminf_n f_n d\nu \leq \liminf_n \int f_n d\nu$. (ii) Dominated convergence theorem: If $\lim_{n\to\infty} f_n = f$ a.e. and $|f_n| \leq g$ a.e. for integrable g, then $\int \lim_{n\to\infty} f_n d\nu = \lim_{n\to\infty} \int f_n d\nu$. (iii) Monotone convergence theorem: If $0 \leq f_1 \leq f_2 \leq \cdots$ and $\lim_{n\to\infty} f_n = f$ a.e., then $\int \lim_{n\to\infty} f_n d\nu = \lim_{n\to\infty} \int f_n d\nu$.

Example 1 (Interchange of differentiation and integration): Let $(\Omega, \mathscr{F}, \nu)$ be a measure space and, for any fixed $\theta \in \mathbb{R}$, let $f(\omega, \theta)$ be a Borel function on Ω. Suppose that $\partial f(\omega, \theta)/\partial\theta$ exists a.e. for $\theta \in (a, b) \subset \mathbb{R}$ and that $|\partial f(\omega, \theta)/\partial\theta| \leq g(\omega)$ a.e., where g is an integrable function on Ω. Then for each $\theta \in (a, b)$, $\partial f(\omega, \theta)/\partial\theta$ is integrable and, by Theorem 1(ii), $\frac{d}{d\theta} \int f(\omega, \theta) d\nu = \int \frac{\partial f(\omega, \theta)}{\partial\theta} d\nu$.

Theorem 2 (Change of variables): Let f be measurable from $(\Omega, \mathscr{F}, \nu)$ to $(\Lambda, \mathscr{G})$ and g be Borel on $(\Lambda, \mathscr{G})$. Then $\int_\Omega g \circ f d\nu = \int_\Lambda g d(\nu \circ f^{-1})$, i.e., if either integral exists, then so does the other, and the two are the same.

Theorem 3 (Fubini's theorem): Let ν_i be a σ-finite measure on $(\Omega_i, \mathscr{F}_i), i = 1, 2$, and f be a Borel function on $\prod_{i=1}^2 (\Omega_i, \mathscr{F}_i)$ with $f \geq 0$ or $\int |f| d\nu_1 \times \nu_2 < \infty$. Then $g(\omega_2) = \int_{\Omega_1} f(\omega_1, \omega_2) d\nu_1$ exists a.e. ν_2 and defines a Borel function on Ω_2 whose integral w.r.t. ν_2 exists, and $\int_{\Omega\times\Omega} f(\omega_1, \omega_2) d\nu_1 \times \nu_2 = \int_{\Omega_2} [\int_{\Omega_1} f(\omega_1, \omega_2) d\nu_1] d\nu_2$.

Definition 1 (Absolutely continuous): Let λ and ν be two measures on a measurable space $(\Omega, \mathscr{F}, \nu)$. We say λ is absolutely continuous w.r.t. ν and write $\lambda << \nu$ iff $\nu(A) = 0$ implies $\lambda(A) = 0$.

Theorem 4 (Radon-Nikodym theorem): Let ν and λ be two measure on $(\Omega, \mathscr{F})$ and ν be σ-finite. If $\lambda << \nu$, then there exists a nonnegative Borel function f on Ω such that $\lambda(A) = \int_A f d\nu, A \in \mathscr{F}$. Furthermore, f is unique a.e. ν, i.e. if $\lambda(A) = \int_A g d\nu$ for any $A \in \mathscr{F}$, then $f = g$ a.e. ν.

Example 2: A continuous c.d.f. may not have a p.d.f. w.r.t. Lebesgue measure. A necessary and sufficient condition for a c.d.f. F having a p.d.f. w.r.t. Lebesgue measure is that F is absolute continuous in the sense that for any $\epsilon > 0$, there exists a $\delta > 0$ such that for each finite collection of disjoint bounded open intervals (a_i, b_i), $\sum(b_i - a_i) < \delta$ implies $\sum[F(b_i) - F(a_i)] < \epsilon$.

Proposition 2 (Calculus with Radon-Nikodym derivatives): Let ν be a σ-finite measure on a measure space $(\Omega, \mathscr{F})$. (i) If λ is a measure, $\lambda << \nu$, and $f \geq 0$, then $\int f d\lambda = \int f \frac{d\lambda}{d\nu} d\nu$. (ii) If $\lambda_i, i = 1, 2$, are measures and $\lambda_i << \nu$, then $\lambda_1 + \lambda_2 << \nu$ and $\frac{d(\lambda_1 + \lambda_2)}{d\nu} = \frac{d\lambda_1}{d\nu} + \frac{d\lambda_2}{d\nu}$ a.e. ν. (iii) If τ is a measure, λ is a σ-finite measure, and $\tau << \lambda << \nu$, then $\frac{d\tau}{d\nu} = \frac{d\tau}{d\lambda}\frac{d\lambda}{d\nu}$ a.e. ν. In particular, if $\lambda << \nu$ and $\nu << \lambda$ (in which case λ and ν are equivalent), then $\frac{d\lambda}{d\nu} = (\frac{d\nu}{d\lambda})^{-1}$ a.e. ν or λ. (iv) Let $(\Omega_i, \mathscr{F}_i, \nu_i)$ be a measure space and ν_i be σ-finite, $i = 1, 2$. Let λ_i be a σ-finite measure on $(\Omega, \mathscr{F}_i)$ and $\lambda_i << \nu_i, i = 1, 2$. Then $\lambda_1 \times \lambda_2 << \nu_1 \times \nu_2$ and $\frac{d(\lambda_1 \times \lambda_2)}{d(\nu_1 \times \nu_2)}(\omega_1, \omega_2) = \frac{d\lambda_1}{d\nu_1}(\omega_1)\frac{d\lambda_2}{d\nu_2}(\omega_2)$ a.e. $\nu_1 \times \nu_2$.

1.3 Densities, moments, inequalities, and generating functions

Example 1: Let X be a random variable on $(\Omega, \mathscr{F}, P)$ whose c.d.f. F_X has a Lebesgue p.d.f. f_x and $F_x(c) < 1$, where c is a fixed constant. Let $Y = \min\{X, c\}$. Note that $Y^{-1}((-\infty, X]) = \Omega$ if $x \geq c$ and $Y^{-1}((-\infty, x]) = X^{-1}((-\infty, x])$ if $x < c$. Hence Y is a random variable and the c.d.f. of Y is $F_Y(x) = \begin{cases} 1 & x \geq c \\ F_X(x) & x < c \end{cases}$. This c.d.f. is discontinuous at c, since $F_x(c) < 1$. Thus, it does not have a Lebesgue p.d.f. It is not discrete either. Does P_Y, the probability measure corresponding to F_y, have a p.d.f. w.r.t. some measure? Consider the point mass probability measure on $(\mathbb{R}, \mathscr{B})$: $\delta_c(A) = \begin{cases} 1 & c \in A \\ 0 & c \notin A \end{cases}$, $A \in \mathscr{B}$. Then $P_Y << m + \delta_c$, and the p.d.f. of P_Y is $f_Y(x) = \frac{dP_Y}{d(m+\delta_c)}(x) = \begin{cases} 0 & x > c \\ 1 - F_X(c) & x = c \\ f_X(x) & x < c \end{cases}$. To show this, it suffices to show that $\int_{(-\infty, x]} f_Y(t) d(m + \delta_c) = P_Y((-\infty, x])$ for any $x \in \mathscr{B}$.

Proposition 1 (Transformation): Let X be a random k-vector with a Lebesgue p.d.f. f_X and let $Y = g(X)$, where g is a Borel function from $(\mathbb{R}^k, \mathscr{B}^k)$ to $(\mathbb{R}^k, \mathscr{B}^l)$. Let $A_1, \cdots, A_m$ be disjoint sets in $\mathscr{B}^k$ such that $\mathscr{R}^k - (A_1 \cup \cdots \cup A_m)$ has Lebesgue measure 0 and g on A_j is one-to-one with a nonvanishing Jacobian, i.e., the determinant $\mathrm{Det}(\partial g(x)/\partial x) \neq 0$ on $A_j, j = 1, \cdots, m$. Then Y has the following Lebesgue p.d.f.: $f_Y(x) = \sum_{j=1}^m |\mathrm{Det}(\partial h_j(x)/\partial x)| f_X(h_j(x))$, where h_j is the inverse function of g on $A_j, j = 1, \cdots, m$.

Example 2 (F-distribution): Let X_1 and X_2 be independent random variables having the chi-square distributions $\chi^2_{n_1}$ and $\chi^2_{n_2}$, respectively. One can show that the p.d.f. of $Y = (X_1/n_1)/(X_2/n_2)$ is the p.d.f. of the F-distribution F_{n_1,n_2}.

Example 3 (t-distribution): Let U_1 be a random variable having the standard normal distribution $N(0,1)$ and U_2 a random variable having the chi-square distribution χ^2_n. One can show that if U_1 and U_2 are independent, then the distribution of $T = U_1/\sqrt{U_2/n}$ is the t-distribution t_n.

Example 4 (Noncentral chi-square distribution): Let $X_1, \cdots, X_n$ be independent random variables and $X_i \sim N(\mu_i, \sigma^2)$. The distribution of $Y = (X_1^2 + \cdots + X_n^2)/\sigma^2$ is called the noncentral chi-square distribution and denoted by $\chi^2_n(\delta)$, where $\delta = (\mu_1^2 + \cdots + \mu_n^2)/\sigma^2$ is the noncentrality parameter. If $Y_1, \cdots, Y_k$ are independent random variables aand Y_i has the noncentral independent chi-square distribution $\chi^2_{n_i}(\delta_i), i = 1, \cdots, k$, then $Y = Y_1 + \cdots + Y_k$ has the noncentral chi-square distribution $\chi^2_{n_1+\cdots+n_k}(\delta_1 + \cdots + \delta_k)$.

Definition 1 (Moments): If $\mathbb{E}X^k$ is finite, where k is a positive integer, $\mathbb{E}X^k$ is called the k-th moment of X or P_x. If $\mathbb{E}|X|^a < \infty$ for some real number a, $\mathbb{E}|X|^a$ is called the a-th absolute moment of X or P_X. If $\mu = \mathbb{E}X$, $\mathbb{E}(X-\mu)^k$ is called the k-th central moment of X or P_X. $\mathrm{Var}(X) = \mathbb{E}(X - \mathbb{E}X)^2$ is called the variance of X or P_X. For random matrix $M = (M_{ij})$, $\mathbb{E}M = (\mathbb{E}M_{ij})$. For random vector X, $\mathrm{Var}(X) = \mathbb{E}(X - \mathbb{E}X)(X - \mathbb{E}X)^T$ is its covariance matrix, whose (i,j)-th element, $i \neq j$, is called the covariance of X_i and X_j and denoted by $\mathrm{Cov}(X_i, X_j)$. If $\mathrm{Cov}(X_i, X_j) = 0$, then X_i and X_j aare said to be uncorrelated. Independence implies uncorelation, not converse. If X is random and c is fixed, then $\mathbb{E}(c^T X) = c^T \mathbb{E}(X)$ and $\mathrm{Var}(c^T X) = c^T \mathrm{Var}(X) c$.

Definition 2 (Moment generating and characteristic functions): Let X be a random k-vector. (i) The moment generating function (m.g.f.) of X or P_X is defined as $\psi_X(t) = \mathbb{E}e^{t^T X}, t \in \mathbb{R}^k$. (ii) The characteristic function (ch.f.) of X or P_X is defined as $\phi_X(t) = \mathbb{E}e^{it^T X} = \mathbb{E}[\cos(t^T X)] + i\mathbb{E}[\sin(t^T X)], t \in \mathbb{R}^k$.

Proposition 2 (Properties of m.g.f. and ch.f.): If the m.g.f. is finite in a neighborhood of $0 \in \mathbb{R}^k$, then (i) moments of X of any order are finite; (ii) $\phi_X(t)$ can be obtained by replacing t in $\psi_X(t)$ by it. If $Y = A^T X + c$, where A is a $k \times m$ matrix and $c \in \mathbb{R}^m$, then $\psi_Y(u) = e^{c^T u}\psi_X(Au)$ and $\phi_Y(u) = e^{ic^T u}\phi_X(Au), u \in \mathbb{R}^m$. For independent $X_1, \cdots, X_k$, $\psi_{\sum_i X_i}(t) = \prod_i \psi_{X_i}(t)$ and $\phi_{\sum_i X_k}(t) = \prod_i \phi_{X_i}(t), t \in \mathbb{R}^k$. For $X = (X_1, \cdots, X_k)$ with m.g.f. ψ_X finite in a neighborhood of 0, $\frac{\partial \psi_X(t)}{\partial t}|_{t=0} = \mathbb{E}X$, $\frac{\partial^2 \psi_X(t)}{\partial t \partial t^T}|_{t=0} = \mathbb{E}(XX^T)$. If $\mathbb{E}|X_1^{r_1} \cdots X_k^{r_k}| < \infty$ for nonnegative integers $r_1, \cdots, r_k$, then $\frac{\partial \phi_X(t)}{\partial t}|_{t=0} = i\mathbb{E}X$, $\frac{\partial^2 \phi_X(t)}{\partial t \partial t^T}|_{t=0} = -\mathbb{E}(XX^T)$.

Theorem 1 (Uniqueness): Let X and Y be random k-vectors. (i) If $\phi_X(t) = \phi_Y(t)$ for all $t \in \mathbb{R}^k$, then $P_X = P_Y$; (2) If $\psi_X(t) = \psi_Y(t) < \infty$ for all t in a neighborhood of 0, then $P_X = P_Y$.

1.4 Conditional expectation and independence

Definition 1: Let X be an integrable random variable on $(\Omega, \mathscr{F}, P)$. (i) The conditional expectation of X given $\mathscr{A}$ (a sub-σ-field of $\mathscr{F}$), denoted by $\mathbb{E}(X|\mathscr{A})$, is the a.s.-unique random variable satisfying the following two conditions: (a) $\mathbb{E}(X|\mathscr{A})$ is a measurable from $(\Omega, \mathscr{A})$ to $(\mathbb{R}, \mathscr{B})$; (b) $\int_A \mathbb{E}(X|\mathscr{A}) dP = \int_A X dP$ for any $\mathscr{A} \in \mathscr{A}$. (ii) The conditional probability of $B \in \mathscr{F}$ given $\mathscr{A}$ is defined to be $P(B|\mathscr{A}) = \mathbb{E}(I_B|\mathscr{A})$. (iii) Let Y be measurable from $(\Omega, \mathscr{F}, P)$ to $(\Lambda, \mathscr{G})$. The conditionala expectation of X given Y is defined to be $\mathbb{E}(X|Y) = \mathbb{E}[X|\sigma(Y)]$.

Theorem 1: Let Y be measurable from $(\Omega, \mathscr{F})$ to $(\Lambda, \mathscr{G})$ and Z a function from $(\Omega, \mathscr{F})$ to $\mathbb{R}^k$. Then Z is measurable from $(\Omega, \sigma(Y))$ to $(\mathbb{R}^k, \mathscr{B}^k)$ iff there is a measurable function h from $(\Lambda, \mathscr{G})$ such that $Z = h \circ Y$.

Example 1: Let X be an integrable random variable on $(\Omega, \mathscr{F}, P)$, $A_1, A_2, \cdots$ be disjoint events on $(\Omega, \mathscr{F}, P)$ such that $\cup A_i = \Omega$ and $P(A_i) > 0$ for all i, and let $a_1, a_2, \cdots$ be distinct real numbers. Define $Y = a_1 I_{A_1} + a_2 I_{A_2} + \cdots$. We can show that $\mathbb{E}(X|Y) = \sum_{i=1}^{\infty} \frac{\int_{A_i} X dP}{P(A_i)} I_{A_i}$.

Proposition 1: Let X be a random n-vector and Y a random m-vector. Suppose thaat (X, Y) has a joint p.d.f. $f(x, y)$ w.r.t. $\nu \times \lambda$, where ν and λ are σ-finite measures on $(\mathbb{R}^n, \mathscr{B}^n)$ and $(\mathbb{R}^m, \mathscr{B}^m)$, respectively. Let $g(x, y)$ be a Borel function on $\mathbb{R}^{n+m}$ for which $\mathbb{E}|g(X, Y)| < \infty$. Then $\mathbb{E}[g(X, Y)|Y] = \frac{\int g(x,Y) f(x,Y) d\nu(x)}{\int f(x,Y) d\nu(x)}$ a.s.

Definition 2 (Conditional p.d.f.): Let (X, Y) be a random vector with a joint p.d.f. $f(x, y)$ w.r.t. $\nu \times \lambda$. The conditional p.d.f. of X given $Y = y$ is defined to be $f_{X|Y}(x|y)/f_Y(y)$ where $f_Y(y) = \int f(x, y) d\nu(x)$ is the marginl p.d.f. of Y w.r.t. λ.

Proposition 2: Let $X, Y, X_1, X_2, \cdots$ be integrable random variables on $(\Omega, \mathscr{F}, P)$ and $\mathscr{A}$ be a sub-σ-field of $\mathscr{F}$. (i) If $X = c$ a.s., $c \in \mathbb{R}$, then $\mathbb{E}(X|\mathscr{A}) = c$ a.s. (ii) If $X \leq Y$ a.s., then $\mathbb{E}(X|\mathscr{A}) \leq \mathbb{E}(Y|\mathscr{A})$ a.s. (iii) If $a, b \in \mathbb{R}$, then $\mathbb{E}(aX + bY|\mathscr{A}) = a\mathbb{E}(X|\mathscr{A}) + b\mathbb{E}(Y|\mathscr{A})$ a.s. (iv) $\mathbb{E}[\mathbb{E}(X|\mathscr{A})] = \mathbb{E}X$. (v) $\mathbb{E}[\mathbb{E}(X|\mathscr{A})|\mathscr{A}_0] = \mathbb{E}(X|\mathscr{A}_0) = \mathbb{E}[\mathbb{E}(X|\mathscr{A}_0)|\mathscr{A}]$ a.s., where $\mathscr{A}_0$ is a sub-σ-field of $\mathscr{A}$. (vi) If $\sigma(Y) \subset \mathscr{A}$ and $\mathbb{E}|XY| < \infty$, then $\mathbb{E}(XY|\mathscr{A}) = Y\mathbb{E}(X|\mathscr{A})$ a.s. (vii) If X and Y are independent and $\mathbb{E}|g(X, Y)| < \infty$ for a Borel function g, then $\mathbb{E}[g(X, Y)|Y = y] = \mathbb{E}[g(X, y)]$ a.s. P_Y. (viii) If $\mathbb{E}X^2 < \infty$, then $[\mathbb{E}(X|\mathscr{A})]^2 \leq \mathbb{E}(X^2|\mathscr{A})$ a.s. (ix) Fatou's lemma: If $X_n \geq 0$ for any n, then $\mathbb{E}(\liminf_n X_n|\mathscr{A}) \leq \liminf_n \mathbb{E}(X_n|\mathscr{A})$ a.s. (x) Dominated convergence theorem: If $|X_n| \leq Y$ for any n and $X_n \to_{\text{a.s.}} X$, then $\mathbb{E}(X_n|\mathscr{A}) \to_{\text{a.s.}} \mathbb{E}(X|\mathscr{A})$.

Definition 3 (Independence): Let $(\Omega, \mathscr{F}, P)$ be a probability space. (i) Let $\mathscr{C}$ be a collection of subsets in $\mathscr{F}$. Events in $\mathscr{C}$ are said to be independent iff for any positive integer n and distinct events $A_1, \cdots, A_n \in \mathscr{C}$, $P(A_1 \cap A_2 \cap \cdots \cap A_n) = P(A_1)P(A_2)\cdots P(A_n)$. (ii) Collections $\mathscr{C}_i \subset \mathscr{F}, i \in \mathscr{I}$ are said to be independent iff events in any collection of the form $\{A_i \in \mathscr{C}_i : i \in \mathscr{I}\}$ are independent. (iii) Random elements $X_i, i \in \mathscr{I}$, are said to be independent iff $\sigma(X_i), i \in \mathscr{I}$ are independent.

Theorem 2: Let $\mathscr{C}_i, i \in \mathscr{I}$ be independent collections of events. If each $\mathscr{C}_i$ is a π-system, then $\sigma(\mathscr{C}_i), i \in \mathscr{I}$ are independent.

Proposition 2: Let X be a random variable with $\mathbb{E}|X| < \infty$ and let Y_i be random k_i vectors, $i = 1, 2$. Suppose that (X, Y_1) and Y_2 are independent. Then $\mathbb{E}[X|(Y_1, Y_2)] = \mathbb{E}(X|Y_1)$ a.s.

Definition 4 (Conditional independence): Let X, Y, Z be random vectors. We say that given Z, X and Y are conditionally independent iff $P(A|X, Z) = P(A|Z)$ a.s. for any $A \in \sigma(Y)$.

1.5 Convergence modes and relationships

Definition 1 (Convergence modes): Let $X, X_1, X_2, \cdots$ be a random k-vectors defined on a probability space. (i) We say that the sequence $\{X_n\}$ converges to X almost surely and write $X_n \to_{\text{a.s.}} X$ iff $\lim_{n\to\infty} X_n = X$ a.s. (ii) We say that $\{X_n\}$ converges to X in probability and write $X_n \to_p X$ iff for every fixed $\epsilon > 0$, $\lim_{n\to\infty} P(||X_n - X|| > \epsilon) = 0$. (iii) We say that $\{X_n\}$ converges to X in L_r (or in rth moment) with a fixed $r > 0$ and write $X_n \to_{L_r} X$ iff $\lim_{n\to\infty} \mathbb{E}||X_n - X||_r^r = 0$. (iv)

Let $F, F_n, n = 1, 2, \cdots$ be c.d.f.'s on $\mathbb{R}^k$ and $P, P_n, n = 1, 2, \cdots$ be their corresponding probability measures. We say that $\{F_n\}$ converges to F weakly (or $\{P_n\}$ converges to P weakly) and write $F_n \to_w F$ (or $P_n \to_w P$) iff, for each continuity point x of F, $\lim_{n\to\infty} F_n(x) = F(x)$. We say that $\{X_n\}$ converges to X in distribution (or in law) and write $X_n \to_d X$ iff $F_{X_n} \to_w F_X$.

Proposition 1: If $F_n \to_w F$ and F is continuous on $\mathbb{R}^k$, then $\lim_{n\to\infty} \sup_{x\in\mathbb{R}^k} |F_n(x) - F(x)| = 0$.

Theorem 1: For random k-vectors $X, X_1, X_2, \cdots$ on a probability space, $X_n \to_{\text{a.s.}} X$ iff for every $\epsilon > 0$, $\lim_{n\to\infty} P(\cup_{m=n}^{\infty}\{||X_m - X|| > \epsilon\}) = 0$.

Theorem 2 (Borel-Cantelli lemma): Let A_n be a sequence of events in a probability space and $\limsup_n A_n = \cap_{n=1}^{\infty} \cup_{m=n}^{\infty} A_m$. (i) If $\sum_{n=1}^{\infty} P(A_n) < \infty$, then $P(\liminf_n A_n) = 0$. (ii) If $A_1, A_2, \cdots$ re pairwise independent aaand $\sum_{n=1}^{\infty} P(A_n) = \infty$, then $P(\limsup_n A_n) = 1$.

Definition 2: Let $X_1, X_2, \cdots$ be random vectors and $Y_1, Y_2, \cdots$ be random variables defined on a common probability space. (i) $X_n = O(Y_n)$ a.s. iff $P(||X_n|| = O(|Y_n|)) = 1$. (ii) $X_n = o(Y_n)$ a.s. iff $X_n/Y_n \to_{\text{a.s.}} 0$. (iii) $X_n = O_p(Y_n)$ iff, for any $\epsilon > 0$, there is a constant $C_\epsilon > 0$ such that $\sup_n P(||X_n|| \geq C_\epsilon |Y_n|) < \epsilon$. (iv) $X_n = o_p(Y_n)$ iff $X_n/Y_n \to_p 0$.

Theorem 3: (i) If $X_n \to_{\text{a.s.}} X$, then $X_n \to_p X$. (The converse is not true). (ii) If $X_n \to_{L_r} X$ for an $r > 0$, then $X_n \to_p X$. (The converse is not true). (iii) If $X_n \to_p X$, then $X_n \to_d X$. (The converse is not true). (iv) (Skorohod's theorem). If $X_n \to_d X$, then there are random vectors $Y, Y_1, Y_2, \cdots$ defined on a common probability space such that $P_Y = P_X, P_{Y_n} = P_{X_n}, n = 1, 2, \cdots$ and $Y_n \to_{\text{a.s.}} Y$. (v) If, for every $\epsilon > 0, \sum_{n=1}^{\infty} P(||X_n - X|| \geq \epsilon) < \infty$, then $X_n \to_{\text{a.s.}} X$. (vi) If $X_n \to_p X$, then there ais a subsequence such that $X_{n_j} \to_{\text{a.s.}} X$ as $j \to \infty$. (vii) If $X_n \to_d X$ and $P(X = c) = 1$, where $c \in \mathbb{R}^k$ is a constant vector, then $X_n \to_p c$. (viii) Suppose that $X_n \to_d X$. Then for any $r > 0, \lim_{n\to\infty} \mathbb{E}||X_n||_r^r = \mathbb{E}||X||_r^r < \infty$ if $\{||X_n||_r^r\}$ is uniformly integrable in the sense that $\lim_{t\to\infty} \sup_n \mathbb{E}(||X_n||_r^r I_{\{||X_n||_r > t\}}) = 0$.

Proposition 2 (Sufficient conditions for uniform integrability): $\sup_n \mathbb{E}||X_n||_r^{r+\delta} < \infty$ for a $\delta > 0$.

Proposition 3 (Properties of the quotient random variables): (i) Suppose $X, X_1, X_2, \cdots$ are positive random variables. Then $X_n \to_{\text{a.s.}} X$ iff for every $\epsilon > 0$, $\lim_{n\to\infty} P(\sup_{k\geq n} \frac{X_k}{X} > 1 + \epsilon) = 0$, and $\lim_{n\to\infty} P(\sup_{k\geq n} \frac{X}{X_k} > 1 + \epsilon) = 0$. (ii) Suppose $X, X_1, X_2, \cdots$ are positive random variables. If $\sum_{n=1}^{\infty} P(X_n/X > 1 + \epsilon) < \infty$ and $\sum_{n=1}^{\infty} P(X/X_n > 1 + \epsilon) < \infty$, then $X_n \to_{\text{a.s.}} X$.

1.6 Uniform integrability and weak convergence

Definition 1 (Tightness): A sequence $\{P_n\}$ of probability measure on $(\mathbb{R}^k, \mathscr{B}^k)$ is tight if for every $\epsilon > 0$, there is a compact set $C \subset \mathbb{R}^k$ such that $\inf_n P_n(C) > 1 - \epsilon$. If $\{X_n\}$ is a sequence of random k-vectors, then the tightness of $\{P_{X_n}\}$ is the same as the boundedness of $\{||X_n||\}$ in probability.

Proposition 1: Let $\{P_n\}$ be a sequence of probability measures on $(\mathbb{R}^k, \mathscr{B}^k)$. (i) Tightness of $\{P_n\}$ is a necessary and sufficient condition that for every subsequence $\{P_n\}$ there eixsts a further subsequence $\{P_{n_j}\} \subset \{P_n\}$ and a probability measure P on $(\mathbb{R}^k, \mathscr{B}^k)$ such that $P_{n_j} \to_w P$ as $j \to \infty$. (ii) If $\{P_n\}$ is tight and if each subsequence that converges weakly at all converges to the same probability measure P, then $P_n \to_w P$.

Theorem 1 (Useful sufficient and necessary conditions for convergence in distribution): Let $X, X_1, X_2, \cdots$ be random k-vectors. (i) $X_n \to_d X$ is equivalent to any one of the following conditions:

(a) $\mathbb{E}[h(X_n)] \to \mathbb{E}[h(X)]$ for every bounded continuous function h; (b) $\limsup_n P_{X_n}(C) \leq P_X(C)$ for any closed set $C \subset \mathbb{R}^k$; (c) $\liminf_n P_{X_n}(O) \geq P_X(O)$ for any open set $O \subset \mathbb{R}^k$. (ii) Lévy-Cramér continuity theorem. Let $\phi_X, \phi_{X_1}, \phi_{X_2}$ be the ch.f.'s of $X, X_1, X_2, \cdots$, respectively. $X_n \to_d X$ iff $\lim_{n\to\infty} \phi X_n(t) = \phi_X(t)$ for all $t \in \mathbb{R}^k$. (iii) Cramér-Wold device. $X_n \to_d X$ iff $c^T X_n \to_d c^T X$ for every $c \in \mathbb{R}^k$.

Example 1: Let $X_1, \cdots, X_n$ be independent random variables having a common c.d.f. and $T_n = X_1 + \cdots + X_n, n = 1, 2, \cdots$. Suppose that $\mathbb{E}|X_1| < \infty$. It follows from a result in calculus that the ch.f. of X_1 satisfies $\phi_{X_1}(t) = \phi_{X_1}(0) + \sqrt{-1}\mu t + o(|t|)$ as $|t| \to 0$, where $\mu = \mathbb{E}X_1$. Then, the ch.f. of T_n/n is $\phi_{T_n/n}(t) = [\phi_{X_1}(\frac{t}{n})]^n = [1 + \frac{\sqrt{-1}\mu t}{n} + o(\frac{t}{n})]^n \to e^{\sqrt{-1}\mu t}$ for any $t \in \mathbb{R}$ as $n \to \infty$. $e^{\sqrt{-1}\mu t}$ is the ch.f. of the point mass probability measure at μ. Thus $T_n/n \to_d \mu$ and $T_n/n \to_p \mu$.

Proposition 2 (Scheffé's theorem): Let $\{f_n\}$ be a sequence of p.d.f.'s on $\mathbb{R}^k$ w.r.t. ν. Suppose that $\lim_{n\to\infty} f_n(x) = f(x)$ a.e. and $f(x)$ is a p.d.f. w.r.t. ν. Then $\lim_{n\to\infty} \int |f_n(x) - f(x)| d\nu = 0$.

1.7 Convergence of transformations and law of large numbers

Theorem 1 (Continuous mapping theorem): Let $X, X_1, X_2, \cdots$ be random k-vectors defined on a probability space and g be a measure function from $(\mathbb{R}^k, \mathscr{B}^k)$ to $(\mathbb{R}^l, \mathscr{B}^l)$. Suppose that g is continuous a.s. P_X. Then (i) $X_n \to_{\text{a.s.}} X$ implies $g(X_n) \to_{\text{a.s.}} g(X)$; (ii) $X_n \to_p X$ implies $g(X_n) \to_p g(X)$; (iii) $X_n \to_d X$ implies $g(X_n) \to_d g(X)$.

Theorem 2 (Slutsky's theorem): Let $X, X_1, X_2, \cdots, Y_1, Y_2, \cdots$ be random variables on a probability space. Suppose that $X_n \to_d X$ and $Y_n \to_p c$, where c is a constant, where c is a constant. Then (i) $X_n + Y_n \to_d X + c$; (ii) $Y_n X_n \to_d cX$; (iii) $X_n/Y_n \to_d X/c$ if $c \neq 0$.

Theorem 3: Let $X_1, X_2, \cdots$ and $Y = (Y_1 + \cdots, Y_k)$ be random k-vectors satisfying $a_n(X_n - c) \to_d Y$, where $c \in \mathbb{R}^k$ and $\{a_n\}$ is a sequence of positive numbers with $\lim_{n\to\infty} a_n = \infty$. Let g be a function from $\mathbb{R}^k \to \mathbb{R}$. (i) If g is differentiable at c, then $a_n[g(X_n) - g(c)] \to_d [\nabla g(c)^T]Y$, where $\nabla g(x)$ denotes the k-vector of partial derivatives of g at x. (ii) Suppose that g has continuous partial derivatives of order $m > 1$ in a neighborhood of c, with all the partial derivatives of order j, $1 \leq j \leq m-1$, vanishing at c, but with the mth-order partial derivatives not all vanishing at c. Then $a_n^m[g(X_n) - g(c)] \to_d \frac{1}{m!} \sum_{i_1=1}^k \cdots \sum_{i_m=1}^k \frac{\partial^m g}{\partial x_{i_1} \cdots \partial x_{i_m}} |_{x=c} Y_{i_1} \cdots Y_{i_m}$.

Theorem 4 (The δ-method): If Y has the $\mathcal{N}_k(0, \Sigma)$ distribution, then $a_n[g(X_n) - g(c)] \to_d \mathcal{N}(0, [\nabla g(c)]^T \Sigma \nabla g(c))$.

Theorem 5: Let $X_1, X_2, \cdots$ be i.i.d. random variables. (i) The WLLN. A necessary and sufficient condition for the existence of a sequence of real numbers $\{a_n\}$ for which $\frac{1}{n}\sum_{i=1}^n X_i - a_n \to_p 0$ is that $nP(|X_1| > n) \to 0$, in which case we may take $a_n = \mathbb{E}(X_1 1_{\{|X_1| \leq n\}})$. (ii) The SLLN. A necessary and sufficient condition for the existence of a constant c for which $\frac{1}{n}\sum_{i=1}^n X_i \to_{\text{a.s.}} c$ is that $\mathbb{E}|X_1| < \infty$, in which case $c = \mathbb{E}X_1$ and $\frac{1}{n}\sum_{i=1}^n c_i(X_i - \mathbb{E}X_1) \to_{\text{a.s.}} 0$ for any bounded sequence of real numbers $\{c_i\}$.

Theorem 6: Let $X_1, X_2, \cdots$ be independent random variables with finite expectations. (i) The SLLN. If there is a constant $p \in [1, 2]$ such that $\sum_{i=1}^i nfty \frac{\mathbb{E}|X_i|^p}{i^p} < \infty$, then $\frac{1}{n}\sum_{i=1}^n (X_i - \mathbb{E}X_i) \to_{\text{a.s.}} 0$. (ii) The WLLN. If there is a constant $p \in [1, 2]$ such that $\lim_{n\to\infty} \frac{1}{n^p} \sum_{i=1}^n \mathbb{E}|X_i|^p = 0$, then $\frac{1}{n}\sum_{i=1}^n (X_i - \mathbb{E}X_i) \to_p 0$.

1.8 The central limit theorem

Theorem 1 (Lindeberg's CLT): Let $\{X_{nj}, j = 1, \cdots, k_n\}$ be independent random variables with $k_n \to \infty$ as $n \to \infty$ and $0 < \sigma_n^2 = \mathrm{var}(\sum_{j=1}^{k_n} X_{nj}) < \infty, n = 1, 2, \cdots$. If $\frac{1}{\sigma_n^2}\sum_{j=1}^{k_n}\mathbb{E}[(X_{nj} - \mathbb{E}X_{nj})^2 I_{\{|X_{nj}-\mathbb{E}X_{nj}|>\epsilon\sigma_n\}}] \to 0$ for any $\epsilon > 0$, then $\frac{1}{\sigma_n}\sum_{j=1}^{k_n}(X_{nj} - \mathbb{E}X_{nj}) \to_d \mathcal{N}(0, 1)$.

Theorem 2 (Multivariate CLT): For i.i.d. random k-vectors $X_1, \cdots, X_n$ with a finite $\Sigma = \mathrm{var}(X_1)$, $\frac{1}{\sqrt{n}}\sum_{i=1}^{n}(X_i - \mathbb{E}X_1) \to_d \mathcal{N}_k(0, \Sigma)$.

Theorem 3 (Berry-Esséen bound): For i.i.d. $\{X_n\}$ and $W_n = \sqrt{n}(\bar{X} - \mu)/\sigma$, $\sup_t |F_{W_n}(t) - \phi(t)| \le \frac{33}{4}\frac{\mathbb{E}|X_1-\mu|^3}{\sigma^3\sqrt{n}}, n = 1, 2, \cdots$. Thus, the convergence speed of F_{W_n} to ϕ is of the order $n^{-1/2}$.

2 Fundamentals of Statistics

2.1 Models, data, statistics, and sampling distributions

Definition 1: A set of probability measures P_θ on $(\Omega, \mathscr{F})$ indexed by a parameter $\theta \in \Theta$ is said to be a parametric family or follow a parametric model iff $\Theta \subset \mathbb{R}^d$ for some fixed positive integer d and each P_θ is a known probability measure when θ is known. The set Θ is called the parameter space and d is called its dimension. $\mathscr{P} = \{P_\theta : \theta \in \Theta\}$ is identifiable iff $\theta_1 \ne \theta_2$ and $\theta_i \in \Theta$ imply $P_{\theta_1} \ne P_{\theta_2}$, which may be achieved through reparameterization.

Definition 2 (Dominated family): A family of populations $\mathscr{P}$ is dominated by ν (a σ-finite measure) if $P << \nu$ for all $P \in \mathscr{P}$, in which case $\mathscr{P}$ can be identified by the family of densities $\{\frac{dP}{d\nu} : P \in \mathscr{P}\}$ or $\{\frac{dP_\theta}{d\nu} : \theta \in \Theta\}$.

Definition 3 (Exponential families): A parametric family $\{P_\theta : \theta :\in \Theta\}$ dominated by a σ-finite measure ν on $(\Omega, \mathscr{F})$ is called on an exponential family iff $\frac{dP_\theta}{d\nu}(\omega) = \exp\{[\eta(\theta)]^T T(\omega) - \xi(\theta)\}h(\omega), \omega \in \Omega$ where $\xi(\theta) = \log\{\int_\omega \exp\{[\eta(\theta)]^T T(\omega)\}h(\omega)d\nu(\omega)\}$. In an exponential family, consider the parameter $\eta = \eta(\theta)$ and $f_\eta(\omega) = \exp\{\eta^T T(\omega) - \zeta(\eta)\}h(\omega), \omega \in \Omega$. This is called the canonical form for the family, and $\Xi = \{\eta : \zeta(\eta)$ is defined$\}$ is called the natural parameter space. An exponential family in canonical form is a natural exponential family. If there is an open set contained in the natural parameter space of an exponential family, then the family is said to be of full rank.

Theorem 1: Let $\mathscr{P}$ be a natural exponential family. (i) Let $T = (Y, U)$ and $\eta = (\theta, \phi)$, Y and θ have the same dimension. Then, Y has the p.d.f. $f_\eta(y) = \exp\{\theta^T y - \zeta(\eta)\}$. In particular, T has a p.d.f. in a natural exponential family. Furthermore, the conditional distribution of Y given $U = u$ has the p.d.f. $f_{\theta,u}(y) = \exp\{\theta^T y - \zeta_u(\theta)\}$ w.r.t. a σ-finite measure depending on ϕ. Furthermore, the conditional distribution of Y given $U = u$ has the p.d.f. $f_{\theta,u}(y) = \exp(\theta^T y - \zeta_u(\theta))$ w.r.t. a σ-finite measure depending on u. (ii) If η_0 is an interior point of the natural parameter space, then the m.g.f. of $P_{\eta_0} \circ T^{-1}$ is finite in a neighbbrhood of 0 and is given by $\psi_{\eta_0}(t) = \exp\{\zeta(\eta_0 + t) - \zeta(\eta_0)\}$.

Definition 4 (Location-scale families): Let P be a known probability measure on $(\mathbb{R}^k, \mathscr{B}^k), \mathscr{V} \subset \mathbb{R}^k$, and $\mathscr{M}_k$ be a collection of $k \times k$ symmetric positive definite matrices. The family $\{P_{(\mu,\Sigma)} : \mu \in \mathscr{V}, \Sigma \in \mathscr{M}_k\}$ is called a location-scale family (on $\mathbb{R}^k$), where $P_{(\mu,\Sigma)}(B) = P(\Sigma^{-1/2}(B - \mu)), B \in \mathscr{B}^k$. The parameters μ and $\Sigma^{1/2}$ are called the location and scale parameters, respectively.

Definition 5 (Statistics and their sampling distributions): Our data set is a realization of a sample (random vector) X from an unknown population P. Statistic $T(X)$: A measurable function T of X; $T(X)$ is a known value whenever X is known. A nontrivial statistic $T(X)$ is usually simpler than X. Finding the form of the distribution of T is one of the major problems in statistical inference and decision theory.

Example 1: Let $X_1, \cdots, X_n$ be i.i.d. random variables having a common distribution P. The sample mean and sample variance $\bar{X} = \frac{1}{n}\sum_{i=1}^n X_i, S^2 = \frac{1}{n-1}\sum_{i=1}^n (X_i - \bar{X})^2$ are two commonly used statistics.

Example 2 (Order statistics): Let $X = (X_1, \cdots, X_n)$ with i.i.d. random components. Let $X_{(i)}$ be the ith smallest value of $X_1, \cdots, X_n$. The statistics $X_{(1)}, \cdots, X_{(n)}$ are called the order statistics.

2.2 Sufficiency and minimal sufficiency

Definition 1 (Sufficiency): Let X be a sample from an unknown population $P \in \mathscr{P}$, where $\mathscr{P}$ is a family of populations. A statistic $T(X)$ is said to be sufficient for $P \in \mathscr{P}$ iff conditional distribution of X given T is known.

Theorem 1 (The factorization theorem): Suppose that X is a sample from $P \in \mathscr{P}$ and $\mathscr{P}$ is a family of probability measures on $(\mathbb{R}^n, \mathscr{B}^n)$ dominated by a σ-finite measure ν. Then $T(X)$ is sufficient for $P \in \mathscr{P}$ iff there are nonnegative Borel functions h and g_p on the range of T such that $\frac{dP}{d\nu}(x) = g_p(T(x))h(x)$.

Theorem 2: If a family $\mathscr{P}$ is dominated by a σ-finite measure, then $\mathscr{P}$ is dominated by a probability measure $Q = \sum_{i=1}^\infty c_i P_i$, where c_i's are nonnegative constants with $\sum_{i=1}^\infty c_i = 1$ and $P_i \in \mathscr{P}$.

Convention 1: If a statement holds except for outcomes in an event A satisfying $P(A) = 0$ for all $P \in \mathscr{P}$, then we say that the statement holds a.s. $\mathscr{P}$.

Definition 2 (Minimal sufficiency): Let T be a sufficient statistic for $P \in \mathscr{P}$. T is called a minimal sufficient statistic iff, for any other statistic S sufficient for $P \in \mathscr{P}$, there is a measurable function ψ such that $T = \psi(S)$ a.s. $\mathscr{P}$.

Theorem 3 (Existence and uniqueness): Minimal sufficient statistics exist when $\mathscr{P}$ contains distributions on $\mathbb{R}^k$ dominated by a σ-finite measure. If both T and S are minimal sufficient statistics, then by definition there is one-to-one measurable function ψ such that $T = \psi(S)$ a.s. $\mathscr{P}$.

Theorem 4: Let $\mathscr{P}$ be a family of distributions on $\mathbb{R}^k$. (i) Suppose that $\mathscr{P}_0 \subset \mathscr{P}$ and a.s. $\mathscr{P}_0$ implies a.s. $\mathscr{P}$. If T is sufficient for $P \in \mathscr{P}$ and minimal sufficient for $P \in \mathscr{P}_0$, then T is minimal sufficient for $P \in \mathscr{P}$. (ii) Suppose that $\mathscr{P}$ contains p.d.f.'s $f_0, f_1, f_2, \cdots$ w.r.t. a σ-finite ν. Let $f_\infty(x) = \sum_{i=0}^\infty c_i f_i(x)$, where $c_i > 0$ for all i and $\sum_{i=0}^\infty c_i = 1$, and let $T_i(x) = f_i(x)/f_\infty(x)$ when $f_\infty(x) > 0, i = 0, 1, 2, \cdots$. Then $T(x) = (T_0, T_1, T_2, \cdots)$ is minimal sufficient for $P \in \mathscr{P}$. Furthermore, if $\{x : f_i(x) > 0\} \subset \{x : f_0(x) > 0\}$ for all i, then we may replace $f_\infty(x)$ for $f_0(x)$, in which case $T(x) = (T_1, T_2, \cdots)$ is minimal sufficient for $P \in \mathscr{P}$. (iii) Suppose that $\mathscr{P}$ contains p.d.f.'s f_p w.r.t. a σ-finite measure and that there exists a sufficient statistic $T(x)$ such that, for any possible values x and y of X, $f_p(x) = f_p(y)\phi(x, y)$ for all P implies $T(x) = T(y)$, where ϕ is a measurable function. Then $T(x)$ is minimal sufficient for $P \in \mathscr{P}$.

2.3 Completeness

Definition 1 (Ancillary statistics): A statistic $V(x)$ is ancillary iff its distribution does not depend on any unknown quantity. A statistic $V(X)$ is first-order ancillary iff $\mathbb{E}[V(X)]$ does not depend on any unknown quantity.

Remark 1: If $V(x)$ is a non-trivial ancillary statistic, then $\sigma(V)$ does not contain any information about the unknown population P. If $T(x)$ is a statistic and $V(T(x))$ is a non-trivial ancillary statistic, it indicates that the reduced data set by T contains a non-trivial part that does not contain any information about θ and, hence, a further simplification of T may still be needed.

Definition 2 (Completeness): A statistic $T(x)$ is complete (or boundedly complete) for $P \in \mathscr{P}$ iff, for any Borel f (or bounded Borel f), $\mathbb{E}[f(T)] = 0$ for all $P \in \mathscr{P}$ implies $f = 0$ a.s. $\mathscr{P}$.

Remark 2: If T is complete (or boundedly complete) and $S = \psi(T)$ for a measurable ψ, then S is complete (or boundedly complete). A complete and sufficient statistic should be minimal sufficient. But a minimal sufficient statistic may be not complete.

Proposition 1: If P is in an exponential family of full rank with p.d.f.'s given by $f_\eta(x) = \exp\{\eta^T T(x) - \zeta(\eta)\}h(x)$, then $T(x)$ is complete and sufficient for $\eta \in \Xi$.

Example 1: Suppose that $X_1, \cdots, X_n$ are i.i.d. random variables having the $\mathcal{N}(\mu, \sigma^2)$ distribution, $\mu \in \mathbb{R}$, $\sigma > 0$. The joint p.d.f. of $X_1, \cdots, X_n$ is $(2\pi)^{-n/2}\exp\{\eta_1 T_1 + \eta_2 T_2 - n\zeta(\eta)\}$, where $T_1 = \sum_{i=1}^n X_i, T_2 = -\sum_{i=1}^n X_i^2$ and $\eta = (\eta_1, \eta_2) = (\frac{\mu}{\sigma^2}, \frac{1}{2\sigma^2})$. Hence, the family of distributions for $X = (X_1, \cdots, X_n)$ is a natural exponential family of full rank ($\Xi = \mathbb{R} \times (0, \infty)$). Thus $T(X) = (T_1, T_2)$ is complete and sufficient for η.

Example 2: $T(x) = (X_{(1)}, \cdots, X_{(n)})$ of i.i.d. random variables $X_1, \cdots, X_n$ is sufficient for $P \in \mathscr{P}$, where $\mathscr{P}$ is the family of distributions on $\mathbb{R}$ having Lebesgue p.d.f.'s. We can show that $T(x)$ is also complete for $P \in \mathscr{P}$.

Theorem 1 (Basu's theorem): Let V and T be two statistics of X from a population $P \in \mathscr{P}$. If V is ancillary and T is boundedly complete and sufficient for $P \in \mathscr{P}$, then V and T are independent w.r.t. any $P \in \mathscr{P}$.

Example 3: $X_1, \cdots, X_n$ is a random sample from uniform$(\theta, \theta + 1)$, $\theta \in \mathbb{R}$, and $T = (X_{(1)}, X_{(n)})$ is the minimal sufficient statistic for θ. We can show that T is not complete.

Theorem 2: Suppose that S is a minimal sufficient statistic and T is a complete and sufficient statistic. Then T must be minimal sufficient and S must be complete.

2.4 Statistical decision

Convention 1 (Basic elements): X: a sample from a population $P \in \mathscr{P}$. Decision: an action we take after observing X. $\mathscr{A}$: the set of allowable actions. $(\mathscr{A}, \mathscr{F}_{\mathscr{A}})$: the action space. $\mathscr{X}$: the range of X. Decision rule: a measurable function T from $(\mathscr{X}, \mathscr{F}_{\mathscr{X}})$ to $(\mathscr{A}, \mathscr{F}_{\mathscr{A}})$. If $X = x$ is observed, then we take the action $T(x) \in \mathscr{A}$.

Definition 1 (Loss function): $L(P, a)$: a function from $\mathscr{P} \times \mathscr{A}$ to $[0, \infty)$. $L(P, a)$ is Borel for each P. If $X = x$ is observed and our decision rule is T, then our loss is $L(P, T(x))$.

Definition 2 (Risk): The averaaged loss $R_T(P) := \mathbb{E}[L(P, T(X))] = \int_{\mathscr{X}} L(P, T(X)) dP_X(x)$.

Definition 3 (Comparisons): For decision rules T_1 and T_2, T_1 is as good as T_2 iff $R_{T_1}(P) \leq R_{T_2}(P)$ for any $P \in \mathscr{P}$ and is better than T_2 if, in addition, $R_{T_1}P < R_{T_2}(P)$ for some P. T_1 and T_2 are equivalent iff $R_{T_1}(P) = R_{T_2}(P)$ for all $P \in \mathscr{P}$. Optimal rule: If T^* is as good as any other rule in $\mathscr{E}$, a claass of allowable decision rules, then T^* is $\mathscr{E}$-optimal.

Definition 4 (Randomized decision rules): A function δ on $\mathscr{X} \times \mathscr{F}_{\mathscr{A}}$; for every $A \in \mathscr{F}_{\mathscr{A}}$, $\delta(\cdot, A)$ is a Borel function and, for every $x \in \mathscr{X}$, $\delta(x, \cdot)$ is a probability measure on $(\mathscr{A}, \mathscr{F}_{\mathscr{A}})$. If $X = x$ is observed, we have a distribution of actions: $\delta(x, \cdot)$. A nonrandomized rule T is a special randomized decision rule with $\delta(x, \{a\}) = I_{\{a\}}(T(x)), a \in \mathscr{A}, x \in \mathscr{X}$. The loss function for a randomized rule δ is defined as $L(P, \delta, x) = \int_{\mathscr{A}} L(P, a) d\delta(x, a)$, which reduces to the same loss function when δ is nonrandomized. The risk of a randomized δ is then $R_\delta(P) = \mathbb{E}[L(P, \delta, X)] = \int_{\mathscr{X}} \int_{\mathscr{A}} L(P, a) d\delta(x, a) dP_X(x)$.

Example 1: $X = (X_1, \cdots, X_n)$ is a vector of i.i.d. measurements for a parameter $\theta \in \mathbb{R}$. We want to estimate θ. Action space: $(\mathscr{A}, \mathscr{F}_{\mathscr{A}}) = (\mathbb{R}, \mathscr{B})$. A common loss function in this problem is the squared error loss $L(P, a) = (\theta - a)^2, a \in \mathscr{A}$. Let $T(X) = \bar{X}$, the sample mean. The loss for $\bar{X}$ is $(\bar{X} - \theta)^2$. If the population has mean μ and variance $\sigma^2 < \infty$, then $R_{\bar{X}}(P) = (\mu - \theta)^2 + \frac{\sigma^2}{n}$. This problem is a special case of a general problem called estimation. In an estimation problem, a decision rule T is called an estimator.

Example 2: Let $\mathscr{P}$ be a family of distributions, $\mathscr{P}_0 \subset \mathscr{P}$, $\mathscr{P}_1 = \{P \in \mathscr{P} : P \notin \mathscr{P}_0\}$. A hypothesis testing problem can be formulated as that of deciding which of the following two statements is true: $H_0 : P \in \mathscr{P}_0$ versus $H_1 : P \in \mathscr{P}_1$. H_0 is called the null hypothesis and H_1 is the alternative hypothesis. The action space for this problem contains only two elements, i.e., $\mathscr{A} = \{0, 1\}$, where 0 is accepting H_0 and 1 is rejecting H_0. This problem is a special case of a general problem called hypothesis testing. A decision rule is called a test, which msut have the form $I_C(X)$, where $C \in \mathscr{F}_{\mathscr{X}}$ is called the rejection or critical region.

Definition 5 (0-1 loss): $L(P, a) = 0$ if a correct decision is made and 1 if an incorrect decision is made, which leads to the risk $R_T(P) = \begin{cases} P(T(X) = 1) = P(X \in C) & P \in \mathscr{P}_0 \\ P(T(X) = 0) = P(X \notin C) & P \in \mathscr{P}_1 \end{cases}$.

Definition 6 (Admissibility): Let $\mathscr{E}$ be a class of decision rules. A decision rule $T \in \mathscr{E}$ is called $\mathscr{E}$-admissible iff there does not exist any $S \in \mathscr{E}$ that is better than T (in terms of the risk).

Remark 1: An admissible decision rule is not necessarily good. For example, in an estimation problem a silly estimator $T(X) \equiv a$ constant may be admissible.

Proposition 1: Let $T(X)$ be a sufficient statistic for $P \in \mathscr{P}$ and let δ_0 be a decision rule. Then $\delta_1(t, A) = \mathbb{E}[\delta_0(X, A)|T = t]$, which is a randomized decision rule depending only on T, is equivalent to δ_0 if $R_{\delta_0}(P) < \infty$ for any $P \in \mathscr{P}$.

Theorem 1: Suppose that $\mathscr{A}$ is a convex subset of $\mathbb{R}^k$ and that for any $P \in \mathscr{P}$, $L(P, a)$ is a convex function of a. (i) Let δ be a randomized rule satisfying $\int_{\mathscr{A}} ||a|| d\delta(x, a) < \infty$ for any $x \in \mathscr{X}$ and let $T_1(x) = \int_{\mathscr{A}} a d\delta(x, a)$. Then $L(P, T_1(x)) \leq L(P, \delta, x)$ (or $L(P, T_1(x)) < L(P, \delta, x)$) if L is strictly convex in a for any $x \in \mathscr{X}$ and $P \in \mathscr{P}$. (ii) Rao-Blackwell theorem. Let T be a sufficient statistic for $P \in \mathscr{P}$, $T_0 \in \mathbb{R}^k$ be a nonrandomized rule satisfying $\mathbb{E}||T_0|| < \infty$, and $T_1 = \mathbb{E}[T_0(X)|T]$. Then $R_{T_1}(P) \leq R_{T_0}(P)$ for any $P \in \mathscr{P}$. If L is strictly convex in a and T_0 is not a function of T, then T_0 is inadmissible.

Definition 7 (Unbiasedness): In an estimation problem, the bias of an estimator $T(X)$ of a parameter θ of the unknown population is defined to be $b_T(P) = \mathbb{E}[T(X)] - \theta$. An estimator $T(X)$ is unbiased for θ iff $b_T(P) = 0$ for any $P \in \mathscr{P}$.

Approach 1: Define a class $\mathscr{E}$ of decision rules that have some desirable properties and then try to find the best rule in $\mathscr{E}$.

Approach 2: Consider some characteristic R_T of $R_T(P)$, for a given decision rule T, and then minimize R_T over $T \in \mathscr{E}$. Methods include the Bayes rule and the minimax rule.

2.5 Statistical inference

Definition 1 (Three components in statistical inference): Point estimators, hypothesis tests, confidence sets.

Definition 2 (Point estimators): Let $T(X)$ be an estimator of $\theta \in \mathbb{R}$. Bias: $b_T(P) = \mathbb{E}[T(X)] - \theta$. Mean squared error (mse): $\mathrm{mse}_T(P) = \mathbb{E}[T(X) - \theta]^2 = [b_T(P)]^2 + \mathrm{Var}(T(X))$. Bias and mse are two common criteria for the performance of point estimators, i.e., instead of considering risk functions, we use bias and mse to evaluate point estimators.

Definition 3 (Hypothesis tests): To test the hypotheses $H_0 : P \in \mathscr{P}_0$ versus $H_1 : P \in \mathscr{P}_1$, there are two types of errors we may commit: rejecting H_0 when H_0 is true (called the type I error) and accepting H_0 when H_0 is wrong (called the type II error). A test T: a statistic from $\mathscr{X}$ to $\{0, 1\}$.

Theorem 1 (Probabilities of making two types of errors): Type I error rate: $\alpha_T(P) = P(T(X) = 1), P \in \mathscr{P}_0$. Type II error rate: $1 - \alpha_T(P) = P(T(X) = 0), P \in \mathscr{P}_1$. $\alpha_T(P)$ is also called the power function of T. Power function is $\alpha_T(\theta)$ if P is in a parametric family indexed by θ.

Example 1: Let $X_1, \cdots, X_n$ be i.i.d. from the $\mathcal{N}(\mu, \sigma^2)$ distribution with an unknown $\mu \in \mathbb{R}$ and a known σ^2. Consider the hypotheses $H_0 : \mu \le \mu_0$ versus $H_1 : \mu > \mu_0$, where μ_0 is a fixed constant. Since the sample mean $\bar{X}$ is sufficient for $\mu \in \mathbb{R}$, it is reasonable to consider the following class of tests: $T_c(X) = I_{(c,\infty)}(\bar{X})$. By the property of the normal distributions, $\alpha_{T_c}(\mu) = P(T_c(X) = 1) = 1 - \phi(\frac{\sqrt{n}(c-\mu)}{\sigma})$. Since $\phi(t)$ is an increasing function of t, $\sup_{P \in \mathscr{P}_0} \alpha_{T_c}(\mu) = 1 - \phi(\frac{\sqrt{n}(c-\mu_0)}{\sigma})$. In fact, it is also true for $\sup_{P \in \mathscr{P}_1}[1 - \alpha_{T_c}(\mu)] = \phi(\frac{\sqrt{n}(c-\mu_0)}{\sigma})$. If we woudl like to use an α as the level of significance, then the most effective way is to choose a c_α such that $\alpha = \sup_{P \in \mathscr{P}_0} \alpha_{T_{c_\alpha}}(\mu)$, in which case c_α must satisify $1 - \phi(\frac{\sqrt{n}(c_\alpha-\mu_0)}{\sigma}) = \alpha$, i.e., $c_\alpha = \sigma z_{1-\alpha}/\sqrt{n} + \mu_0$, where $z_a = \Phi^{-1}(a)$. It can be shown that for any test $T(X)$ satisfying $\sup_{P \in \mathscr{P}_0} \alpha_T(P) \le \alpha$, $1 - \alpha_T(\mu) \ge 1 - \alpha_{T_{c_\alpha}}(\mu), \mu > \mu_0$.

Definition 4 (Significance tests): A common approach of finding an "optimal" test is to assign a small bound α to the type I error rate $\alpha_T(P), P \in \mathscr{P}_0$, and then to attempt to minimize the type II error rate $1 - \alpha_T(P), P \in \mathscr{P}_1$, subject to $\sup_{P \in \mathscr{P}_0} \alpha_T(P) \le \alpha$. The bound α is called the level of significance. The left-hand side is called the size of the test T. The level of significance should be positive, otherwise no test satisfies.

Definition 5 (p-value): It is good practice to determine not only whether H_0 is rejected for a given a and a chosen test T_α, but also the smallest possible level of significance at which H_0 would be rejected for the computed $T_\alpha(x)$, i.e., $\hat{\alpha} = \inf\{\alpha \in (0, 1) : T_\alpha(x) = 1\}$. Such an $\hat{\alpha}$, which depends on x and the chosen test and is a statistic, is called the p-value for the test T_α.

Example 2: Let us calculate the p-value for T_{c_α} in Example 1. Note that $\alpha = 1-\phi(\frac{\sqrt{n}(c_\alpha-\mu_0)}{\sigma}) > 1-\Phi(\frac{\sqrt{n}(\bar{X}-\mu_0)}{\sigma})$ if and only if $\bar{X} > c_\alpha$ (or $T_{c_\alpha}(x)=1$). Hence, $1-\phi(\frac{\sqrt{n}(\bar{X}-\mu_0)}{\sigma}) = \inf\{\alpha \in (0,1) : T_{c_\alpha}(x)=1\} = \hat{\alpha}(X)$ is the p-value for T_{c_α}. It turns out that $T_{c_\alpha}(x) = I_{(0,\alpha)}(\hat{\alpha}(X))$.

Definition 6 (Confidence sets) θ: a k-vector of unknown parameters related to the unknown $P \in \mathscr{P}$. If a Borel set $C(X)$ (in the range of θ) depending only on the sample X such that $\inf_{P\in\mathscr{P}} P(\theta \in C(X)) \geq 1-\alpha$, where α is a fixed constant in $(0,1)$, then $C(X)$ is called a confidence set for θ with level of significance $1-\alpha$. The left-hand side is called the confidence coefficient of $C(X)$, which is the highest possible level of significance for $C(X)$. A confidence set is a random element that covers the unknown θ with certain probability.

Example 3: Let $X_1,\cdots,X_n$ be i.i.d. from the $\mathcal{N}(\mu,\sigma^2)$ distribution with both $\mu\in\mathbb{R}$ and $\sigma^2>0$ unknown. Let $\theta=(\mu,\sigma^2)$ and $\alpha\in(0,1)$ be given. Let $\bar{X}$ be the sample mean and S^2 be the sample variance. Since $(\bar{X},S^2)$ is sufficient, we focus on $C(X)$ that is a function of $(\bar{X},S^2)$. Since $\sqrt{n}(\bar{X}-\mu)/\sigma$ has the $\mathcal{N}(0,1)$ distribution, $P(-\widetilde{c}_\alpha \leq \frac{\bar{X}-\mu}{\sigma/\sqrt{n}} \leq \widetilde{c}_\alpha) = \sqrt{1-\alpha}$, where $\widetilde{c}_\alpha = \Phi^{-1}(\frac{1+\sqrt{1-\alpha}}{2})$. Since the χ^2 distribution distribution χ^2_{n-1} is a known distribution, we can always find two constants $c_{1\alpha}$ and $c_{2\alpha}$ such that $P(c_{1\alpha} \leq \frac{(n-1)S^2}{\sigma^2} \leq c_{2\alpha}) = \sqrt{1-\alpha}$. Then $P(-\widetilde{c}_\alpha \leq \frac{\bar{X}-\mu}{\sigma/\sqrt{n}} \leq \widetilde{c}_\alpha, c_{1\alpha} \leq \frac{(n-1)S^2}{\sigma^2} \leq c_{2\alpha}) = 1-\alpha$. The LHS defines a set in the range of $\theta=(\mu,\sigma^2)$ bounded by two straight lines,$\sigma^2=(n-1)S^2/c_{i\alpha}, i=1,2$, and a curve $\sigma^2 = n(\bar{X}-\mu)^2/\widetilde{c}_\alpha^2$. This set is a confidence set for θ with confidence coefficient $1-\alpha$.

Definition 7 (Randomized tests): Since the action space contains only two points, 0 and 1, for a hypothesis testing problem, any randomized test $\delta(X,A)$ is equivalent to a statistic $T(X)\in[0,1]$ with $T(x)=\delta(x,\{1\})$ and $1-T(X)=\delta(x,\{0\})$. A nonrandomized test is obviously a special case where $T(x)$ does not take any value in $(0,1)$. For any randomized test $T(X)$, we define the type I error probability to be $\alpha_T(P)=\mathbb{E}[T(X)], P\in\mathscr{P}_0$, and the type II error probability to be $1-\alpha_T(P)=\mathbb{E}[1-T(X)], P\in\mathscr{P}_1$. For a class of randomized tests, we would like to minimize $1-\alpha_T(P)$ subject to $\sup_{P\in\mathscr{P}_0}\alpha_T(P)=\alpha$.

Definition 8 (Consistency of point estimators): Let $X=(X_1,\cdots,X_n)$ be a sample from $P\in\mathscr{P}$, $T_n(X)$ be an estimator of θ for every n, and $\{a_n\}$ be a sequence of positive constants, $a_n\to\infty$. (i) $T_n(x)$ is consistent for θ iff $T_n(x)\to_p\theta$ w.r.t. any P. (ii) $T_n(x)$ is a_n-consistent for θ iff $a_n[T_n(X)-\theta]=O_p(1)$ w.r.t. any P. (iii) $T_n(x)$ is strongly consistent for θ iff $T_n(x)\to_{\text{a.s.}}\theta$ w.r.t. any P. (iv) $T_n(X)$ is L_r-consistent for θ iff $T_n(x)\to_{L_r}\theta$ w.r.t. for any P for some fixed $r>0$; if $r=2$, L_2-consistency is called consistency in mse.

Remark 1 (Consistency is an essential requirement): Like the admissibility, consistency is an essential requirement: any inconsistent estimators should not be used, but there are many consistent estimators and some may not be good. Thus, consistency should be used together with other criteria.

Remark 2 (Approximate and asymptotic bias): Unbiasedness is a criterion for point estimator. In some cases, however, there is no unbiased estimator. Furthermore, having a "slight" bias in some cases may not be a bad idea.

Definition 9: (i) Let $\xi,\xi_1,\xi_2,\cdots$ be random variables and $\{a_n\}$ be a sequence of positive numbers satisfying $a_n\to\infty$ or $a_n\to a>0$. If $a_n\xi_n\to_d\xi$ and $\mathbb{E}|\xi|<\infty$, then $\mathbb{E}\xi/a_n$ is called an asymptotic expectation of ξ_n. (ii) For a point estimator T_n of θ, an asymptotic expectation of $T_n-\theta$, if it exists, is called an asymptotic bias of T_n and denoted by $\widetilde{b}_{T_n}(P)$. If $\lim_{n\to\infty}\widetilde{b}_{T_n}(P)=0$ for any P, then T_n is asymptotically unbiased.

Proposition 1 (Asymptotic expectation is essentially unique): For a sequence of random variables $\{\xi_n\}$, suppose both $\mathbb{E}\xi/a_n$ and $\mathbb{E}\eta/b_n$ are asymptotic expectations of ξ_n. Then, one of the following three must hold: (a) $\mathbb{E}\xi=\mathbb{E}\eta=0$; (b) $\mathbb{E}\xi\neq 0, \mathbb{E}\eta=0$, and $b_n/a_n\to 0$; (c) $\mathbb{E}\xi\neq 0, \mathbb{E}\eta\neq 0$, and $(\mathbb{E}\xi/a_n)/(\mathbb{E}\eta/b_n)\to 1$.

Example 4 (Functions of sample means): We consider the case where $X_1,\cdots,X_n$ are i.i.d. random k-vectors with finite $\Sigma=\mathrm{Var}(X_1), T_n=g(\bar{X})$, where g is a function on $\mathbb{R}^k$ that is second-order differentiable at $\mu=\mathbb{E}X_1$. Consider T_n as an estimator of $\theta=g(\mu)$. By Taylor's expansion, $T_n-\theta=[\nabla g(\mu)]^T(\bar{X}-\mu)+2^{-1}(\bar{X}-\mu)^T\nabla^2 g(\mu)(\bar{X}-\mu)+o_p(n^{-1})$. By the CLT, $2^{-1}n(\bar{X}-\mu)\nabla^2 g(\mu)(\bar{X}-\mu)\to_d 2^{-1}Z_\Sigma^T\nabla^2 g(\mu)Z_\Sigma$, where $Z_\Sigma=\mathcal{N}_k(0,\Sigma)$. Thus, $\frac{\mathbb{E}[Z_\Sigma^T\nabla^2 g(\mu)Z_\Sigma]}{2n}=\frac{\mathrm{tr}(\nabla^2 g(\mu)\Sigma)}{2n}$ is the n^{-1} order asymptotic bias of $T_n=g(\bar{X})$.

Definition 10 (Asymptotic variance and amse): Let T_n be an estimator of θ for every n and $\{a_n\}$ be a sequence of positive numbers satisfying $a_n\to\infty$ or $a_n\to a>0$. Assume that $a_n(T_n-\theta)\to_d Y$ with $0<\mathbb{E}Y^2<\infty$. (i) The asymptotic mean squared error of T_n, denoted by $\mathrm{amse}_{T_n}(P)$, is defined as the asymptotic expectation of $(T_n-\theta)^2$, $\mathrm{amse}_{T_n}(P)=\mathbb{E}Y^2/a_n^2$. The asymptotic variance of T_n is defined as $\sigma^2_{T_n}(P)=\mathrm{Var}(Y)/a_n^2$. (ii) Let T_n' be another estimator of θ. The asymptotic relative efficiency of T_n' w.r.t. T_n is defined as $e_{T_n',T_n}=\mathrm{amse}_{T_n}(P)/\mathrm{amse}_{T_n'}(P)$. (iii) T_n is said to be asymptotically more efficient than T_n' iff $\limsup_n e_{T_n',T_n}(P)\le 1$ for any P and <1 for some P.

Proposition 2: Let T_n be an estimator of θ for every n and $\{a_n\}$ be a sequence of positive numbers satisfying $a_n\to\infty$ or $a_n\to a>0$. If $a_n(T_n-\theta)\to_d Y$ with $0<\mathbb{E}Y^2<\infty$, then (i) $\mathbb{E}Y^2\le\liminf_n\mathbb{E}[a_n^2(T_n-\theta)^2]$ and (ii) $\mathbb{E}Y^2=\lim_{n\to\infty}\mathbb{E}[a_n^2(T_n-\theta)^2]$ if and only if $\{a_n^2(T_n-\theta)^2\}$ is uniformly integrable.

Example 5: Let $X_1,\cdots,X_n$ be i.i.d. from the Poisson distribution $P(\theta)$ with an unknown $\theta>0$. Consider the estimation of $\theta=P(X_i=0)=e^{-\theta}$. Let $T_{1n}=F_n(0)$, where F_n is the empirical c.d.f. Then T_{1n} is unbiased and has $\mathrm{mse}_{T_{1n}}(\theta)=e^{-\theta}(1-e^{-\theta})/n$. Also, $\sqrt{n}(T_{1n}-\theta)\to_d\mathcal{N}(0,e^{-\theta}(1-e^{-\theta}))$ by the CLT. Thus, in the case $\mathrm{amse}_{T_{1n}}(\theta)=\mathrm{mse}_{T_{1n}}(\theta)$. Consider $T_{2n}=e^{-\bar{X}}$. Note that $\mathbb{E}T_{2n}=e^{n\theta(e^{-1/n}-1)}$, hence $nb_{T_{2n}}(\theta)\to\theta e^{-\theta}/2$. Using the CLT, we can show that $\sqrt{n}(T_{2n}-\theta)\to_d\mathcal{N}(0,e^{-2\theta}\theta)$. Then $\mathrm{amse}_{T_{2n}}(\theta)=e^{-2\theta}\theta/n$. Thus, the asymptotic relative efficiency of T_{1n} w.r.t. T_{2n} is $e_{T_{1n},T_{2n}}=\theta/(e^\theta-1)<1$. This shows that T_{2n} is asymptotically more efficient than T_{1n}.