The Monte Carlo Method in Quantum Field Theory

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- Path integrals in quantum mechanics
- Monte Carlo integration and Markov chains
- Monte Carlo evaluation of path integral in quantum mechanics
- Free Klein-Gordon scalar field theory in 2 + 1 dimensions
- Interacting ϕ^4 scalar field theory in 2 + 1 dimensions
- Applications in quantum chromodynamics
- Topics for future study

Part I

Path integrals in quantum mechanics

Transition amplitudes in quantum mechanics

key quantity in quantum mechanics: transition amplitude

 $Z(b,a) \equiv \langle x_b(t_b) \mid x_a(t_a) \rangle$

- Z(b, a) is probability amplitude for particle to go from point x_a at time t_a to point x_b at time t_b
- in this talk, will work in Heisenberg picture
 - state vectors $|\Psi\rangle$ are stationary
 - operators and their eigenvectors evolve with time

 $\begin{array}{lll} x(t) &=& e^{iHt/\hbar} x(0) \ e^{-iHt/\hbar} \\ |x(t)\rangle &=& e^{iHt/\hbar} \ |x(0)\rangle \end{array}$

often will shift Hamiltonian so ground state energy is zero

 $egin{array}{rcl} H \mid & \phi_n(t)
angle &=& E_n \mid \phi_n(t)
angle, & E_0 = 0 \ & \mid & \phi_0(t)
angle &=& \mid & \phi_0(0)
angle \equiv \mid & 0
angle \end{array}$

Spectral representation of transition amplitude

• insert complete (discrete) set of Heisenberg-picture eigenstates $|\phi_n(t)\rangle$ of Hamiltonian *H* into transition amplitude

 $Z(b,a) \equiv \langle x_b(t_b) \mid x_a(t_a) \rangle = \sum_{n} \langle x_b(t_b) \mid \phi_n(t_b) \rangle \langle \phi_n(t_b) \mid x_a(t_a) \rangle$ • now use $|\phi_n(t)\rangle = e^{iHt/\hbar} |\phi_n(0)\rangle \stackrel{n}{=} e^{iE_nt/\hbar} |\phi_n(0)\rangle$ to obtain $Z(b,a) = \sum_{n} e^{-iE_n(t_b-t_a)/\hbar} \langle x_b(t_b) \mid \phi_n(t_b) \rangle \langle \phi_n(t_a) \mid x_a(t_a) \rangle$

• finally, $\langle x(t) | \phi_n(t) \rangle^n \equiv \varphi_n(x)$ is the wavefunction in coordinate space, so

$$Z(b,a) = \sum_n arphi_n(x_b) arphi_n^*(x_a) \; e^{-iE_n(t_b-t_a)/\hbar t_a}$$

 transition amplitude contains information about all energy levels and all wavefunctions → spectral representation

Vacuum saturation

• take $t_a = -T$ and $t_b = T$ in the limit $T \to (1 - i\epsilon)\infty$ $\langle x_b(T) | \mathbf{x}_a(-T) \rangle = \langle x_b(0) | e^{-iHT/\hbar} e^{iH(-T)/\hbar} | \mathbf{x}_a(0) \rangle$ $= \sum_{n=0}^{\infty} \langle x_b(0) | \phi_n(0) \rangle \langle \phi_n(0) | \mathbf{x}_a(0) \rangle e^{-2iE_nT/\hbar}$ $\to \langle x_b(0) | 0 \rangle \langle 0 | \mathbf{x}_a(0) \rangle$

• insert complete set of energy eigenstates, use $E_{n+1} \ge E_n$, $E_0 = 0$, assume nondegenerate vacuum

• possibility of probing ground state (vacuum) properties

Vacuum expectation values

• now apply limit $T \rightarrow (1 - i\epsilon)\infty$ to more complicated amplitude

$$\begin{aligned} \langle x_b(T) | x(t_2) x(t_1) | x_a(-T) \rangle \\ &= \langle x_b(0) | e^{-iHT/\hbar} x(t_2) x(t_1) e^{-iHT/\hbar} | x_a(0) \rangle \end{aligned}$$

$$= \sum_{n,m} \langle x_b(0) | \phi_n(0) \rangle \langle \phi_n(0) | x(t_2) x(t_1) | \phi_m(0) \rangle \langle \phi_m(0) | x_a(0) \rangle \\ \times e^{-i(E_n + E_m)T/\hbar}$$

 $\rightarrow \langle x_b(0)|0\rangle \langle 0|x(t_2)x(t_1)|0\rangle \langle 0|x_a(0)\rangle$

hence, vacuum expectation values from

$$\langle 0|x(t_2)x(t_1)|0\rangle = \lim_{T \to (1-i\epsilon)\infty} \frac{\langle x_b(T)|x(t_2)x(t_1)|x_a(-T)\rangle}{\langle x_b(T)|x_a(-T)\rangle}$$

result generalizes to higher products of position operator

Observables from correlation functions

- all observables can be extracted from the correlation functions (vacuum expectation values)
- example: energies of the stationary states

$$\begin{aligned} \langle 0|x(t)x(0)|0\rangle &= & \langle 0|e^{iHt/\hbar}x(0)e^{-iHt/\hbar}x(0)|0\rangle \\ &= & \sum_{n} \langle 0|x(0)e^{-iHt/\hbar}|\phi_{n}(0)\rangle\langle\phi_{n}(0)|x(0)|0\rangle \\ &= & \sum_{n} |\langle 0|x(0)|\phi_{n}(0)\rangle|^{2}e^{-iE_{n}t/\hbar} \end{aligned}$$

similarly for more complicated correlation functions

$$\begin{array}{lll} \langle 0|x^{2}(t)x^{2}(0)|0\rangle & = & \langle 0|e^{iHt/\hbar}x^{2}(0)e^{-iHt/\hbar}x^{2}(0)|0\rangle \\ & = & \sum_{n}|\langle 0|x^{2}(0)|\phi_{n}(0)\rangle|^{2}e^{-iE_{n}t/\hbar} \end{array}$$

but difficult to extract energies *E_n* from above oscillatory functions
 → much easier if we had decaying exponentials

• can get decaying exponentials if we rotate from the real to the imaginary axis in time (Wick rotation) $t \rightarrow -i\tau$

$$\begin{array}{lll} \langle 0|x(t)x(0)|0\rangle & = & \sum_{n} |\langle 0|x(0)|\phi_{n}(0)\rangle|^{2}e^{-E_{n}\tau/\hbar} \\ & \xrightarrow[\tau \to \infty]{} & |\langle 0|x(0)|0\rangle|^{2} + |\langle 0|x(0)|\phi_{1}(0)\rangle|^{2}e^{-E_{1}\tau/\hbar} \end{array}$$

• later, will see this imaginary time formalism provides another important advantage for Monte Carlo applications

Quantum mechanics and path integrals

- in the 1940s, Feynman developed an alternative formulation of quantum mechanics (his Ph.D. thesis)
 - Richard Feynman, Rev Mod Phys 20, 367 (1948)
- quantum mechanical law of motion:
 - probability amplitude from sum over histories

 $Z(b,a) \sim \sum_{\substack{\text{all paths } x(t) \\ \text{from } a \text{ to } b}} \exp(iS[x(t)]/\hbar)$



- all paths contribute to probability amplitude, but with different phases determined by the action S[x(t)]
- <u>classical limit</u>: when small changes in path yield changes in action large compared to \hbar , phases cancel out and path of least action $\delta S = 0$ dominates sum over histories

Defining the path integral

action = time integral of Lagrangian (kinetic minus potential energy)

$$S = \int dt \ L(x, \dot{x}) = \int dt \ \left(K - U\right)$$

- divide time into steps of width ε where $N\varepsilon = t_b t_a$
- path integral is defined as

$$Z(b,a) = \lim_{\varepsilon \to 0} \frac{1}{A} \int_{-\infty}^{\infty} \frac{dx_1}{A} \frac{dx_2}{A} \cdots \frac{dx_{N-1}}{A} e^{iS[x(t)]/\hbar}$$

where *A* is a normalization
factor depending on ε chosen
so path integral well-defined
• in nonrelativistic theory, paths
cannot double-back in time

Schrödinger equation

• probability amplitude $\psi(x_b, t_b)$ at time t_b given amplitude $\psi(x_a, t_a)$ at earlier time t_a given by $\psi(x_b,t_b) = \int Z(b,a) \ \psi(x_a,t_a) \ dx_a$ • take $t_a = t$ and $t_b = t + \varepsilon$ one time slice away $\psi(x_b,t+\varepsilon) = \frac{1}{A} \int_{-\infty}^{\infty} \exp\left[\frac{i\varepsilon}{\hbar} L\left(\frac{x_b+x_a}{2},\frac{x_b-x_a}{\varepsilon}\right)\right] \psi(x_a,t) \, dx_a$ • in L, take $\dot{x} = (x_b - x_a)/\varepsilon$ and mid-point prescription $x \to (x_b + x_a)/2$ • particle in potential: $L = \frac{1}{2}m\dot{x}^2 - V(x,t)$, write $x_b = x$, $x_a = x + \eta$ $\psi(x,t+\varepsilon) = \frac{1}{4} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} e^{-i\varepsilon V(x+\eta/2,t)/\hbar} \psi(x+\eta,t) \, d\eta$

Schrödinger equation (continued)

$$\psi(x,t+\varepsilon) = \frac{1}{A} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} e^{-i\varepsilon V(x+\eta/2,t)/\hbar} \psi(x+\eta,t) \, d\eta$$

- rapid oscillation of e^{imη²/(2ħε)} except when η ~ O(√ε) → integral dominated by contributions from η having values of this order
- expand to $O(\varepsilon)$ and $O(\eta^2)$, except $e^{im\eta^2/(2\hbar\varepsilon)}$ (ψ refers to $\psi(x,t)$)

$$\begin{split} \psi + \varepsilon \frac{\partial \psi}{\partial t} &= \frac{1}{A} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} \left[1 - \frac{i\varepsilon}{\hbar} V(x,t) \right] \left[\psi + \eta \frac{\partial \psi}{\partial x} + \frac{\eta^2}{2} \frac{\partial^2 \psi}{\partial x^2} \right] d\eta \\ &= \frac{1}{A} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} \left[\psi - \frac{i\varepsilon}{\hbar} V(x,t) \psi + \eta \frac{\partial \psi}{\partial x} + \frac{\eta^2}{2} \frac{\partial^2 \psi}{\partial x^2} \right] d\eta \end{split}$$

Schrödinger equation (continued)

$$\psi + \varepsilon \frac{\partial \psi}{\partial t} = \frac{1}{A} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} \left[\psi - \frac{i\varepsilon}{\hbar} V(x,t) \psi + \eta \frac{\partial \psi}{\partial x} + \frac{\eta^2}{2} \frac{\partial^2 \psi}{\partial x^2} \right] d\eta$$

• matching leading terms on both sides determines *A* (analytic continuation to evaluate integral)

$$1 = \frac{1}{A} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} d\eta = \frac{1}{A} \left(\frac{2\pi i\hbar\varepsilon}{m} \right)^{1/2} \Rightarrow A = \left(\frac{2\pi i\hbar\varepsilon}{m} \right)^{1/2}$$

• two more integrals:

$$\frac{1}{A} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} \eta \, d\eta = 0, \qquad \frac{1}{A} \int_{-\infty}^{\infty} e^{im\eta^2/(2\hbar\varepsilon)} \eta^2 d\eta = \frac{i\hbar\varepsilon}{m}$$

• *O*(ε) part of equation at top yields

$$-\frac{\hbar}{i} \frac{\partial \psi}{\partial t} = -\frac{\hbar^2}{2m} \frac{\partial^2 \psi}{\partial x^2} + V(x,t)\psi$$

• the Schrödinger equation!

Free particle in one dimension

- Lagrangian of free particle in one dimension $L = \frac{1}{2}m\dot{x}^2$
- amplitude for particle to travel from x_a at time t_a to location x_b at later time t_b is

 $\langle x_b(t_b) | x_a(t_a) \rangle = \int^b \mathcal{D}x(t) \exp(iS[b,a]/\hbar)$

summing over all allowed paths with $x(t_a) = x_a$ and $x(t_b) = x_b$.

• classical path $x_{cl}(t)$ from $\delta S = 0$ and boundary conditions:

$$\ddot{x}_{cl}(t) = 0,$$
 $x_{cl}(t) = x_a + (x_b - x_a) \frac{(t - t_a)}{(t_b - t_a)}$

• classical action is $S_{cl}[b,a] = \int_{t_a}^{t_b} dt \ \frac{1}{2}m\dot{x}_{cl}^2 = \frac{m(x_b - x_a)^2}{2(t_b - t_a)}$

• write $x(t) = x_{cl}(t) + \chi(t)$ where $\chi(t_a) = \chi(t_b) = 0$ then

$$S[b,a] = S_{cl}[b,a] + \int_{t}^{t_{b}} dt \ \frac{1}{2}m\dot{\chi}^{2}$$

where $S_{cl}[b, a]$ is classical action; no terms linear in $\chi(t)$ since S_{cl} is extremum

Path integral for free particle

• amplitude becomes $(T = t_b - t_a)$

$$Z(b,a) = F(T) \exp(iS_{cl}/\hbar)$$

$$F(T) = \int_0^0 \mathcal{D}\chi \exp\left\{\frac{im}{2\hbar}\int_0^T dt \,\dot{\chi}^2\right\}$$

partition time into discrete steps of length ε, use midpoint prescription, and note that χ₀ = χ_N = 0

$$\int_{0}^{0} \mathcal{D}\chi = \frac{1}{A} \int_{-\infty}^{\infty} \left(\prod_{l=1}^{N-1} \frac{d\chi_{l}}{A} \right) \qquad A = \left(\frac{2\pi i\hbar\varepsilon}{m} \right)^{1/2}$$
$$\int_{0}^{T} dt \ \dot{\chi}^{2} = \frac{1}{\varepsilon} \sum_{j=0}^{N-1} (\chi_{j+1} - \chi_{j})^{2}$$
$$F(T) = \left(\frac{m}{2\pi i\hbar\varepsilon} \right)^{N/2} \int_{-\infty}^{\infty} \left(\prod_{l=1}^{N-1} d\chi_{l} \right) \exp\left\{ \frac{im}{2\hbar\varepsilon} \chi_{j} M_{jk} \chi_{k} \right\}$$

Gaussian integration

a multivariate Gaussian integral remains

$$F(T) = \left(\frac{m}{2\pi i \hbar \varepsilon}\right)^{N/2} \int_{-\infty}^{\infty} \left(\prod_{l=1}^{N-1} d\chi_l\right) \exp\left\{\frac{im}{2\hbar \varepsilon} \chi_j M_{jk} \chi_k\right\}$$

where *M* is a symmetric $(N-1) \times (N-1)$ matrix

$$M = \begin{bmatrix} 2 & -1 & 0 & 0 & \cdots \\ -1 & 2 & -1 & 0 & \cdots \\ 0 & -1 & 2 & -1 & \cdots \\ \vdots & \vdots & \vdots & \vdots & \ddots \end{bmatrix}$$

Gaussian integrals of symmetric matrix A easily evaluated

• result:
$$\int_{-\infty}^{\infty} \left(\prod_{i=1}^{n} d\chi_{i}\right) \exp\left(-\chi_{j}A_{jk}\chi_{k}\right) = \left(\frac{\pi^{n}}{\det A}\right)^{1/2}$$
$$F(T) = \left(\frac{m}{2\pi i\hbar\varepsilon \det M}\right)^{1/2}$$

Determinant evaluation

- now need to compute det(M)
- consider $n \times n$ matrix B_n of form

$$B_{n} = \begin{pmatrix} 2b & -b & 0 & 0 & \cdots \\ -b & 2b & -b & 0 & \cdots \\ 0 & -b & 2b & -b & \cdots \\ \vdots & \vdots & \vdots & \vdots & \ddots \end{pmatrix}_{n,n}$$

notice that

$$\det B_n = 2b \det B_{n-1} + b \det \left(\begin{array}{c|c} -b & -b & 0 & \cdots \\ \hline 0 & & \\ \vdots & & B_{n-2} \end{array} \right)$$
$$= 2b \det B_{n-1} - b^2 \det B_{n-2}$$

• define $I_n = \det B_n$ then have recursion relation $I_{n+1} = 2bI_n - b^2 I_{n-1}, \quad I_{-1} = 0, \quad I_0 = 1, \quad n = 0, 1, 2, \dots$

Transition amplitude for free particle

• rewrite $I_{n+1} = 2bI_n - b^2I_{n-1}$, $I_{-1} = 0$, $I_0 = 1$ as $\begin{pmatrix} I_{n+1} \\ I_{n-1} \end{pmatrix} = \begin{pmatrix} 2b & -b^2 \\ 1 & 0 \end{pmatrix} \begin{pmatrix} I_n \\ I_{n-1} \end{pmatrix} = \begin{pmatrix} 2b & -b^2 \\ 1 & 0 \end{pmatrix}^n \begin{pmatrix} I_1 \\ I_0 \end{pmatrix}$ straightforward to show that $\begin{pmatrix} 2b & -b^2 \\ 1 & 0 \end{pmatrix}^n = \begin{pmatrix} (n+1)b^n & -nb^{n+1} \\ nb^{n-1} & -(n-1)b^n \end{pmatrix}$ • so that $\begin{pmatrix} I_{n+1} \\ I_n \end{pmatrix} = \begin{pmatrix} (n+1)b^n & -nb^{n+1} \\ nb^{n-1} & -(n-1)b^n \end{pmatrix} \begin{pmatrix} 2b \\ 1 \end{pmatrix}$ • and thus, $I_n = \det B_n = (n+1)b^n$ • here, b = 1 and n = N - 1 so det M = N and using $N\varepsilon = t_b - t_a$ obtain $F(t_b, t_a) = \left(\frac{m}{2\pi i\hbar(t_b - t_a)}\right)^{1/2}$ Final result:

$$\langle x_b(t_b)|x_a(t_a)\rangle = \left(\frac{m}{2\pi i\hbar(t_b-t_a)}\right)^{1/2} \exp\left\{\frac{im(x_b-x_a)^2}{2\hbar(t_b-t_a)}\right\}$$

Infinite square well

- one of the first systems usually studied when learning quantum mechanics is the infinite square well
- particle moving in one dimension under influence of potential

$$V(x) = \begin{cases} 0 & \text{for } 0 < x < L \\ \infty & \text{for } x \le 0 \text{ and } x \ge L \end{cases}$$

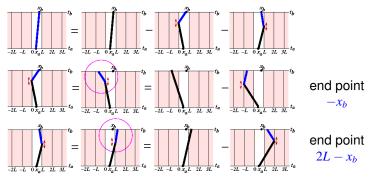
path integral for transition amplitude given by

$$Z(b,a) = \lim_{\varepsilon \to 0} \frac{1}{A} \int_0^L \frac{dx_1}{A} \cdots \int_0^L \frac{dx_{N-1}}{A} \exp\left\{\frac{im}{2\varepsilon\hbar} \sum_{j=0}^{N-1} (x_{j+1}-x_j)^2\right\}$$

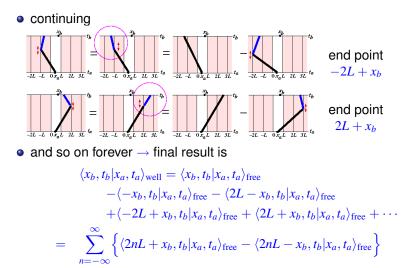
- paths limited to 0 < x < L
- gaussian integrals over bounded domains produce error functions → direct evaluation difficult in closed form
- extend regions of integration to −∞ < x < ∞, but subtract off forbidden paths
 - M. Goodman, Am. Jour. Phys. 49, 9 (1981)

Path cancellations

- black lines: all unbounded paths between end points
- blue lines: paths between end points that do not cross an nL boundary
- no doubling back in time
- magenta circle indicates action preserving reflection



Path cancellations (continued)



Transition amplitude for infinite square well

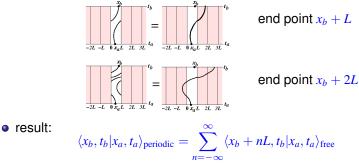
substitute amplitude for free particle

 $\langle x_b(t_b)|x_a(t_a)\rangle = \left(\frac{m}{2\pi i\hbar(t_b-t_a)}\right)^{1/2}$ $\times \sum_{n=1}^{\infty} \left(\exp\left\{ \frac{im(2nL+x_b-x_a)^2}{2\hbar(t_b-t_a)} \right\} - \exp\left\{ \frac{im(2nL-x_b-x_a)^2}{2\hbar(t_b-t_a)} \right\} \right)$ apply Poisson summation and integrate the gaussian $\sum_{n=-\infty}^{\infty} f(n) = \sum_{i=-\infty}^{\infty} \int_{-\infty}^{\infty} ds f(s) e^{2\pi i j s}$ $\int_{-\infty}^{\infty} ds \exp\left(-i\alpha s^2 \pm i\beta s\right) = \sqrt{\frac{\pi}{i\alpha}} \exp\left(\frac{i\beta^2}{4\alpha}\right)$ spectral representation of transition amplitude $\langle x_b(t_b)|x_a(t_a)\rangle = \sum_{a}^{\infty} \varphi_n(x_b)\varphi_n^*(x_a)e^{-iE_n(t_b-t_a)/\hbar}$

$$E_n = \frac{n^2 \pi^2 \hbar^2}{2mL^2} \qquad \varphi_n(x) = \sqrt{\frac{2}{L}} \sin\left(\frac{n\pi x}{L}\right)$$

Free particle in 1D periodic box

- consider particle moving in one-dimension with periodic boundary conditions at x = 0 and x = L
- enforcing boundary conditions on paths difficult
- use trick similar to that used in infinite square well
- express set of allowed paths in terms of equivalent set of unrestricted paths



Transition amplitude for periodic boundary

 substitute amplitude for free particle $\langle x_b(t_b)|x_a(t_a)\rangle = \left(\frac{m}{2\pi i\hbar(t_b - t_a)}\right)^{1/2} \sum_{k=0}^{\infty} \exp\left\{\frac{im(nL + x_b - x_a)^2}{2\hbar(t_b - t_a)}\right\}$ apply Poisson summation and integrate the gaussian $\sum_{n=-\infty}^{\infty} f(n) = \sum_{j=-\infty}^{\infty} \int_{-\infty}^{\infty} ds f(s) e^{2\pi i j s}$ $\int_{-\infty}^{\infty} ds \exp\left(-i\alpha s^2 \pm i\beta s\right) = \sqrt{\frac{\pi}{i\alpha}} \exp\left(\frac{i\beta^2}{4\alpha}\right)$ • spectral representation of transition amplitude $\langle x_b(t_b)|x_a(t_a)
angle = \sum_{a} \varphi_n(x_b)\varphi_n^*(x_a)e^{-iE_n(t_b-t_a)/\hbar}$ $E_n = \frac{p_n^2}{2m}$ $p_n = \frac{2\pi n\hbar}{L}$ $\varphi_n(x) = \frac{1}{\sqrt{L}}e^{ip_n x/\hbar}$ guantization of momenta

The simple harmonic oscillator

 kinetic and potential energy of a simple harmonic oscillator of mass *m* and frequency ω

$$K = \frac{1}{2}m\dot{x}^2 \qquad U = \frac{1}{2}m\omega^2 x^2$$

action is given by

$$S[x(t)] = \int_{t_a}^{t_b} dt \ \left(\frac{1}{2}m\dot{x}^2 - \frac{1}{2}m\omega^2 x^2\right)$$

classical equations of motion

$$\delta S = 0 \quad \Rightarrow \quad \ddot{x}_{\rm cl} + \omega^2 x_{\rm cl} = 0$$

• value of action for the classical path $(T = t_b - t_a)$

$$S_{\rm cl} = \frac{m\omega}{2\sin(\omega T)} \left[(x_a^2 + x_b^2)\cos(\omega T) - 2x_a x_b \right]$$

• to calculate, write path as deviation from classical path $x(t) = x_{cl}(t) + \chi(t)$ $\chi(t_a) = \chi(t_b) = 0$

Path integral of simple harmonic oscillator

amplitude can then be written as

$$Z(b,a) = F(T) \exp(iS_{cl}/\hbar)$$

$$F(T) = \int_{0}^{0} \mathcal{D}\chi \exp\left\{\frac{im}{2\hbar}\int_{0}^{T} dt \left(\dot{\chi}^{2} - \omega^{2}\chi^{2}\right)\right\}$$
• partition time into discrete steps of length ε and use midpoint prescription

$$\int_{0}^{0} \mathcal{D}\chi = \frac{1}{A} \int_{-\infty}^{\infty} \left(\prod_{l=1}^{N-1} \frac{d\chi_{l}}{A} \right) \qquad A = \left(\frac{2\pi i\hbar\varepsilon}{m} \right)^{1/2}$$
$$\int_{0}^{T} dt (\dot{\chi}^{2} - \omega^{2}\chi^{2}) = \frac{1}{\varepsilon} \sum_{j=0}^{N-1} \left[(\chi_{j+1} - \chi_{j})^{2} - \frac{\varepsilon^{2}\omega^{2}}{4} (\chi_{j+1} + \chi_{j})^{2} \right]$$
$$F(T) = \left(\frac{m}{2\pi i\hbar\varepsilon} \right)^{N/2} \int_{-\infty}^{\infty} \left(\prod_{l=1}^{N-1} d\chi_{l} \right) \exp\left\{ \frac{im}{2\hbar\varepsilon} \chi_{j} M_{jk} \chi_{k} \right\}$$

Gaussian integration

• a multivariate Gaussian integral remains

$$F(T) = \left(\frac{m}{2\pi i \hbar \varepsilon}\right)^{N/2} \int_{-\infty}^{\infty} \left(\prod_{l=1}^{N-1} d\chi_l\right) \exp\left\{\frac{im}{2\hbar \varepsilon} \chi_j M_{jk} \chi_k\right\}$$

where *M* is a symmetric $(N-1) \times (N-1)$ matrix

$$M = \begin{bmatrix} 2 & -1 & 0 & 0 & \cdots \\ -1 & 2 & -1 & 0 & \cdots \\ 0 & -1 & 2 & -1 & \cdots \\ \vdots & \vdots & \vdots & \vdots & \ddots \end{bmatrix} - \frac{\varepsilon^2 \omega^2}{4} \begin{bmatrix} 2 & 1 & 0 & 0 & \cdots \\ 1 & 2 & 1 & 0 & \cdots \\ 0 & 1 & 2 & 1 & \cdots \\ \vdots & \vdots & \vdots & \vdots & \ddots \end{bmatrix}$$

• Gaussian integrals are easily evaluated

$$F(T) = \left(\frac{m}{2\pi i\hbar\varepsilon \det M}\right)^{1/2}$$

Evaluating the determinant

- now must compute det M
- consider $det(B_n)$ where $n \times n$ matrix B_n has form

$$B_{n} = \begin{pmatrix} a & b & 0 & 0 & \cdots \\ b & a & b & 0 & \cdots \\ 0 & b & a & b & \cdots \\ \vdots & \vdots & \vdots & \vdots & \ddots \end{pmatrix}_{n},$$

• matches *M* for n = N - 1, $a = 2(1 - \epsilon^2 \omega^2 / 4)$, $b = -(1 + \epsilon^2 \omega^2 / 4)$

notice that

det
$$B_n = a \det B_{n-1} - b \det \begin{pmatrix} b & b & 0 & \cdots \\ 0 & & \\ \vdots & & B_{n-2} \end{pmatrix}$$

= $a \det B_{n-1} - b^2 \det B_{n-2}$

• define $I_n = \det B_n$ to obtain recursion relation

 $I_{n+1} = aI_n - b^2 I_{n-1}, \qquad I_{-1} = 0, \quad I_0 = 1, \qquad n = 0, 1, 2, \dots$

Evaluating the determinant (continued)

rewrite recursion relation as

$$\begin{pmatrix} I_{n+1} \\ I_n \end{pmatrix} = \begin{pmatrix} a & -b^2 \\ 1 & 0 \end{pmatrix} \begin{pmatrix} I_n \\ I_{n-1} \end{pmatrix} = \begin{pmatrix} a & -b^2 \\ 1 & 0 \end{pmatrix}^n \begin{pmatrix} I_1 \\ I_0 \end{pmatrix}$$

• diagonalize

$$\begin{pmatrix} a & -b^2 \\ 1 & 0 \end{pmatrix} = \mathcal{S} \begin{pmatrix} \lambda_+ & 0 \\ 0 & \lambda_- \end{pmatrix} \mathcal{S}^{-1}$$
$$\lambda_{\pm} = \frac{1}{2} \left(a \pm \sqrt{a^2 - 4b^2} \right),$$
$$\mathcal{S} = \begin{pmatrix} \lambda_+ & \lambda_- \\ 1 & 1 \end{pmatrix} \quad \mathcal{S}^{-1} = \frac{1}{\lambda_+ - \lambda_-} \begin{pmatrix} 1 & -\lambda_- \\ -1 & \lambda_+ \end{pmatrix}$$

then we have

$$\begin{pmatrix} I_{n+1} \\ I_n \end{pmatrix} = S \begin{pmatrix} \lambda_+^n & 0 \\ 0 & \lambda_-^n \end{pmatrix} S^{-1} \begin{pmatrix} a \\ 1 \end{pmatrix}$$
$$I_n = \det B_n = \frac{\lambda_+^{n+1} - \lambda_-^{n+1}}{\lambda_+ - \lambda_-} \qquad (\lambda_+ \neq \lambda_-)$$

thus

Amplitude for simple harmonic oscillator

• using $\lambda_{\pm} = 1 \pm i\omega\epsilon + O(\epsilon^2)$ yields $\lim_{\substack{\epsilon \to 0 \\ N \to \infty}} \varepsilon \det M = \lim_{\substack{\epsilon \to 0 \\ N \to \infty}} \varepsilon \frac{1}{2i\omega\varepsilon} \left((1 + i\omega\varepsilon)^N - (1 - i\omega\varepsilon)^N \right)$ $= \lim_{\substack{\epsilon \to 0 \\ N \to \infty}} \frac{1}{2i\omega} \left(\left(1 + \frac{i\omega T}{N} \right)^N - \left(1 - \frac{i\omega T}{N} \right)^N \right)$ $= \frac{1}{2i\omega} \left(e^{i\omega T} - e^{-i\omega T} \right) = \frac{\sin \omega T}{\omega}.$ • final result for the path integral

$$\langle x_b(t_b)|x_a(t_a)\rangle_{\rm sho} = \left(\frac{m\omega}{2\pi i\hbar\sin(\omega(t_b-t_a))}\right)^{1/2} \exp\left\{iS_{\rm cl}/\hbar\right\}$$

Evolution of gaussian wave packet

• for initial wave packet at time $t_a = 0$ with probability dist.

$$|\phi(x_a, t_a)|^2 = \frac{1}{\sigma\sqrt{2\pi}} \exp\left(-\frac{(x_a - \bar{x})^2}{2\sigma^2}\right)$$

probability amplitude at later time

$$\phi(x_b, t_b) = \int_{-\infty}^{\infty} dx_a Z(b, a) \phi(x_a, 0)$$

= $\left(\frac{-im\omega(2\pi)^{-3/2}}{\hbar\sigma\sin(\omega t_b)}\right)^{1/2} \int_{-\infty}^{\infty} dx_a e^{iS_{cl}/\hbar} e^{-(x_a - \bar{x})^2/(4\sigma^2)}$

final result for probability distribution: Gaussian with width s

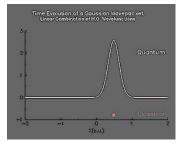
$$|\phi(x_b, t_b)|^2 = \frac{1}{s\sqrt{2\pi}} \exp\left(-\frac{(x_b - \bar{x}\cos(\omega t_b))^2}{2s^2}\right)$$

new width given by

$$s = \sigma \left\{ \cos^2(\omega t_b) + \frac{\hbar^2}{4m^2\omega^2\sigma^4} \sin^2(\omega t_b) \right\}^{1/2}$$

Visualization

 time evolution of a Gaussian wave packet for a simple harmonic oscillator



mass $m = 1 \text{g/mol} = 1.66 \times 10^{-27} \text{kg}$ frequency $\omega = 3 \times 10^{14} \text{radians/sec}$ initial wave packet: center at 0.5 au RMS spread 0.14 au 1 au (atomic unit) = 0.529 angstrom probability distribution shown (in inverse a.u.)

 completely calculated using path integrals → did not use Schrodinger equation • so path integrals give us simple transition amplitudes

$$\langle x_b(t_b)|x_a(t_a)\rangle = \int_a^b \mathcal{D}x \exp\left\{\frac{i}{\hbar}\int_t^{t_b} dt L(x,\dot{x})\right\}$$

 but this important result generalizes to more complicated amplitudes

$$\langle x_b(t_b) | x(t_2) x(t_1) | x_a(t_a) \rangle$$

$$= \int_a^b \mathcal{D}x x(t_2) x(t_1) \exp\left\{\frac{i}{\hbar} \int_{t_a}^{t_b} dt L(x, \dot{x})\right\}$$

for $t_a < t_1 < t_2 < t_b$

 in imaginary time formalism, paths contribute to sum over histories with real exponential weights (not phases)

$$\langle x_b(\tau_b) | x(\tau_2) x(\tau_1) | x_a(\tau_a) \rangle$$

$$= \int_a^b \mathcal{D}x x(\tau_2) x(\tau_1) \exp\left\{-\frac{1}{\hbar} \int_{\tau_a}^{\tau_b} d\tau L(x, \dot{x})\right\}$$

- classical path gets highest weighting
- note that weights are all real and positive since action is real
 - this fact will be crucial for the Monte Carlo method

Vacuum expectation values from path integrals

 obtain correlation functions (vacuum expectation values) from ratios of path integrals

$$\langle 0|x(t_2)x(t_1)|0\rangle = \lim_{T \to \infty} \frac{\langle x_b(T)|x(t_2)x(t_1)|x_a(-T)\rangle}{\langle x_b(T)|x_a(-T)\rangle}$$
$$= \frac{\int_a^b \mathcal{D}x \ x(t_2)x(t_1) \exp\left\{-\frac{1}{\hbar}\int_{-\infty}^\infty d\tau L(x,\dot{x})\right\}}{\int_a^b \mathcal{D}x \ \exp\left\{-\frac{1}{\hbar}\int_{-\infty}^\infty d\tau L(x,\dot{x})\right\}}$$

• generalizes to more complicated correlation functions

• any correlation function can be computed using path integrals

Examples for the simple harmonic oscillator

 evaluating path integrals as before, the following correlation functions can be obtained (τ₁ ≤ τ₂ ≤ τ₃ ≤ τ₄)

$$\langle 0|x(\tau_1)|0\rangle = 0 \langle 0|x(\tau_2)x(\tau_1)|0\rangle = \frac{\hbar}{2m\omega}e^{-\omega(\tau_2-\tau_1)} \langle 0|x(\tau_4)x(\tau_3)x(\tau_2)x(\tau_1)|0\rangle = \left(\frac{\hbar}{2m\omega}\right)^2 e^{-\omega(\tau_4-\tau_1)} \times \left[e^{-\omega(\tau_2-\tau_3)} + 2e^{-\omega(\tau_3-\tau_2)}\right]$$

comparison with spectral representation tells us

$$\langle 0|x(\tau)x(0)|0\rangle = \frac{\hbar}{2m\omega}e^{-\omega\tau} \Rightarrow E_1 - E_0 = \hbar\omega \qquad |\langle 1|x(0)|0\rangle|^2 = \frac{\hbar}{2m\omega}$$

Another example in SHO

• excite vacuum with $x(\tau)^2$ operator

$$\langle 0|x^2(\tau)x^2(0)|0
angle = \left(rac{\hbar}{2m\omega}
ight)^2 \left(1+2e^{-2\omega\tau}
ight)$$

• compare with spectral representation at large time separations

$$\begin{split} \lim_{\tau \to \infty} \langle 0 | x^2(\tau) x^2(0) | 0 \rangle &= |\langle 0 | x^2(0) | 0 \rangle|^2 \\ &+ |\langle 2 | x^2(0) | 0 \rangle|^2 \ e^{-(E_2 - E_0)t/\hbar} + \dots \\ &= \left(\frac{\hbar}{2m\omega}\right)^2 \left(1 + 2e^{-2\omega\tau}\right) \end{split}$$

interpretation:

$$E_2 - E_0 = 2\hbar\omega$$
$$|\langle 0|x^2(0)|0\rangle|^2 = \left(\frac{\hbar}{2m\omega}\right)^2 \qquad |\langle 2|x^2(0)|0\rangle|^2 = 2\left(\frac{\hbar}{2m\omega}\right)^2$$

One last example in SHO

to determine expectation value of x(0)² in first-excited state

$$\langle 0|x(\tau) x^2(\frac{1}{2}\tau) x(0)|0\rangle = 3\left(\frac{\hbar}{2m\omega}\right)^2 e^{-\omega\tau}$$

• compare with spectral interpretation at large times

 $\lim_{\tau \to \infty} \langle 0|x(\tau)x^2(\frac{1}{2}\tau)x(0)|0\rangle$ = $|\langle 0|x(0)|1\rangle|^2 \langle 1|x^2(0)|1\rangle e^{-(E_1-E_0)\tau/\hbar} + \cdots$

• since $\langle 0|x(0)|0
angle = \langle 0|x(au)|0
angle = 0$

by inspection and using previously derived results

 $\langle 1|x^2(0)|1\rangle = \frac{3\hbar}{2m\omega}$

- observables in quantum mechanics can be extracted from the correlation functions (vacuum expectation values)
- imaginary time formalism is a great trick for assisting in such extractions
- correlation functions can be computed via path integrals

$$= \frac{\langle 0|x(t_2)x(t_1)|0\rangle}{\int_a^b \mathcal{D}x \ x(t_2)x(t_1) \exp\left\{-\frac{1}{\hbar}\int_{-\infty}^\infty d\tau L(x,\dot{x})\right\}}{\int_a^b \mathcal{D}x \ \exp\left\{-\frac{1}{\hbar}\int_{-\infty}^\infty d\tau L(x,\dot{x})\right\}}$$



Part II

Monte Carlo integration and Markov chains

• in rare situations, the path integrals can be computed exactly

- simple harmonic oscillator, free particle
- sometimes the action can be written $S = S_0 + gS_I$
 - S₀ describes the free motion of the particles
 - path integrals using S₀ are Gaussian and can be exactly computed
 - S_t describes the interaction of the particles, but the coupling g is small
 - compute in perturbation theory as expansion in g
- however, if interactions are not weak
 - usually must resort to Monte Carlo methods
 - for example, quantum chromodynamics (QCD)

Simple Monte Carlo integration

- trapezoidal/Simpson's rule not feasible for integrals of very large dimension: too many function evaluations
- must start gambling!
- basic theorem of Monte Carlo integration

$$\int_{V} f(\vec{x}) d^{D}x pprox V\langle f
angle \pm V \sqrt{rac{\langle f^{2}
angle - \langle f
angle^{2}}{N}}$$
 $\langle f
angle \equiv rac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i}) \qquad \langle f^{2}
angle \equiv rac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i})^{2}$

- N points x₁,..., x_N chosen independently and randomly with uniform probability dist. throughout D-dimensional volume V
- justified by the law of large numbers/central limit theorem
- in the limit N → ∞, MC estimate tends to normal distribution, uncertainty tends to standard deviation

Quick review of probabilities

- consider an experiment whose outcome depends on chance
- represent an outcome by *X* called a random variable
- sample space Ω of experiment is set of all possible outcomes
- X is discrete if Ω is finite or countably infinite, continuous otherwise
- probability distribution for discrete *X* is real-valued function p_X on domain Ω satisfying $p_X(x) \ge 0$ for all $x \in \Omega$ and $\sum_{x \in \Omega} p_X(x) = 1$
- for any subset *E* of Ω , probability of *E* is $P(E) = \sum_{x \in E} p_X(x)$
- a sequence of random variables X_1, X_2, \ldots, X_N that are mutually independent and have same distribution is called an independent trials process

- for continuous real-valued *X*, real-valued function p_X is a probability density and probability of outcome between real values *a* and *b* is $P(a \le X \le b) = \int_a^b p_X(s) ds$
- cumulative distribution is $F_X(x) = P(X \le x) = \int_{-\infty}^x p_X(s) ds$
- common density: normal $p_X(x) = \frac{1}{\sqrt{2\pi\sigma}} e^{-(x-\mu)^2/(2\sigma^2)}$

Review: expected values

• expected value of X is

 $E(X) = \sum_{x \in \Omega} x p_X(x)$ $\left(= \int_{-\infty}^{\infty} s p_X(s) ds\right)$

- properties: E(X + Y) = E(X) + E(Y) and E(cX) = cE(X)
- for independent random variables X, Y have E(XY) = E(X)E(Y)
- can show E(X) is average of outcomes if repeated many times
- for continuous real-valued function *f*, can show that

$$E(f(X)) = \sum_{x \in \Omega} f(x) p_X(x) \qquad \left(= \int_{-\infty}^{\infty} f(s) p_X(s) ds \right)$$

- Proof: group together terms in $\sum_{x} f(x) p_X(x)$ having same f(x) value
- denote set of different f(x) values by F, and subset of Ω leading to same value of f(x) by Ω_{f(x)}, then

$$\sum_{x \in \Omega} f(x) p_X(x) = \sum_{y \in \mathcal{F}} \sum_{x \in \Omega_{f(x)}} f(x) p_X(x) = \sum_{y \in \mathcal{F}} y \left(\sum_{x \in \Omega_{f(x)}} p_X(x) \right)$$
$$= \sum_{y \in \mathcal{F}} y p(y) = E(f(x))$$

- variance of X is $V(X) = E((X E(X))^2)$
- standard deviation of *X* is $\sigma(X) = \sqrt{V(X)}$
- properties: $V(cX) = c^2 V(X)$ and V(X + c) = V(X)
- for <u>in</u>dependent random variables X, Y have V(X + Y) = V(X) + V(Y) (exercise: prove this)
- let X_1, \ldots, X_N be an independent trials process with $E(X_j) = \mu$ and $V(X_j) = \sigma^2$, and define $A_N = (X_1 + X_2 + \cdots + X_N)/N$, then can show $E(A_N) = \mu$, $V(A_N) = \sigma^2/N$

Chebyshev inequality

• Chebyshev inequality: Let *X* be a discrete random variable with $E(X) = \mu$ and let $\epsilon > 0$ be any positive real number, then

$$P(|X - \mu| \ge \epsilon) \le \frac{V(X)}{\epsilon^2}$$

Proof:

- Let $p_X(x)$ denote distribution of *X*, then probability that *X* differs from μ by at least ϵ is $P(|X - \mu| \ge \epsilon) = \sum_{|x-\mu| \ge \epsilon} p_X(x)$
- considering positive summands and the ranges of summation,

 $V(X) = \sum_{x} (x-\mu)^2 p_X(x) \ge \sum_{|x-\mu| \ge \epsilon} (x-\mu)^2 p_X(x) \ge \sum_{|x-\mu| \ge \epsilon} \epsilon^2 p_X(x)$

• but rightmost expression is $|x-\mu| \ge \epsilon$

$$e^{2}\sum_{|x-\mu|\geq\epsilon}p_{X}(x)=\epsilon^{2}P(|X-\mu|\geq\epsilon)$$

• thus, have shown $V(x) \ge \epsilon^2 P(|X - \mu| \ge \epsilon)$

• Weak law of large numbers: Let $X_1, X_2, ..., X_N$ be an independent trials process with $E(X_j) = \mu$ and $V(X_j) = \sigma^2$, where μ, σ are finite, and let $A_N = (X_1 + X_2 + \cdots + X_N)/N$. Then for any $\epsilon > 0$,

 $\lim_{\substack{N\to\infty\\\mathbf{f}}} P(|A_N-\mu|\geq\epsilon)=0,\qquad \lim_{N\to\infty} P(|A_N-\mu|<\epsilon)=1$

Proof:

- stated two slides ago that $E(A_N) = \mu$ and $V(A_N) = \sigma^2/N$
- from Chebyshev inequality

$$P(|A_N-\mu|\geq\epsilon)\leq rac{V(A_N)}{\epsilon^2}=rac{\sigma^2}{N\epsilon^2}\stackrel{N
ightarrow\infty}{\longrightarrow} 0$$

- also known as the law of averages
- also applies to continuous random variables

Strong law of large numbers

- Strong law of large numbers: Let $X_1, X_2, ..., X_N$ be an independent trials process with $E(X_j) = \mu$ and $E(X_j^4) = K$, where μ, K are finite, then $P\left(\lim_{N \to \infty} (X_1 + X_2 + \dots + X_N)/N = \mu\right) = 1$
- the finiteness of $E(X_i^4)$ is not needed, but simplifies proof
- Proof:
 - define $Y_j = X_j \mu$ so $E(Y_j) = 0$ and set $E(Y_j^4) = C < \infty$
 - define $A_N = (Y_1 + Y_2 + \dots + Y_N)/N$
 - given $E(Y_j) = 0$ and all Y_j are independent,

 $N^{4}E(A_{N}^{4}) = NE(Y_{j}^{4}) + 6\binom{n}{2}E(Y_{i}^{2}Y_{j}^{2}) = NC + 3N(N-1)E(Y_{i}^{2})^{2}$

- since $0 \le V(Y_j^2) = E(Y_j^4) E(Y_j^2)^2$ then $E(Y_j^2)^2 \le E(Y_j^4) = C$
- so $E(A_N^4) \le C/N^3 + 3C/N^2$ which means $E(\sum_{N=1}^{\infty} A_N^4) = \sum_{N=1}^{\infty} E(A_N^4) \le \sum_{N=1}^{\infty} \left(\frac{C}{N^3} + \frac{3C}{N^2}\right) < \infty$
- this implies $\sum_{N=1}^{\infty} A_N^4 < \infty$ with unit probability, and convergence of the series implies $\lim_{N\to\infty} A_N^4 = 0 \implies \lim_{N\to\infty} A_N = 0$
- proves E(X) is average of outcomes for many repetitions

Application to one-dimensional integral

- if *X* is a random variable with probability density $p_X(x)$ and *f* is a well-behaved real-valued function, then Y = f(X) is a random variable
- consider uniform density $p_X(x) = \begin{cases} 1/(b-a) & a \le x \le b \\ 0 & \text{otherwise} \end{cases}$
- use this probability density to obtain N outcomes X_1, X_2, \ldots, X_n
- apply function f to obtain random variables $Y_j = f(X_j)$
- law of large numbers tell us that

$$\frac{1}{N}\sum_{j=1}^{N}Y_{j} \xrightarrow{N \to \infty} E(Y) = E(f(X)) = \frac{1}{(b-a)}\int_{a}^{b}f(s)ds$$

• define $\langle f \rangle \equiv \frac{1}{N}\sum_{j=1}^{N}f(X_{j})$ then $(b-a)\lim_{N \to \infty}\langle f \rangle = \int_{a}^{b}f(s)ds$

- straightforward generalization to multiple dimensions
- how good is estimate for finite *N*?

Central limit theorem

• Central limit theorem: Let $X_1, X_2, ..., X_N$ be independent random variables with common distribution having $E(X_j) = \mu$ and $V(X_j) = \sigma^2$, where μ, σ are finite, and let $A_N = (X_1 + X_2 + \dots + X_N)/N$. Then for a < b,

$$\lim_{N\to\infty} P\left(\frac{a\sigma}{\sqrt{N}} < (A_N - \mu) < \frac{b\sigma}{\sqrt{N}}\right) = \frac{1}{\sqrt{2\pi}} \int_a^b e^{-x^2/2} dx$$

- alternatively: the distribution of $(X_1 + \cdots + X_N N\mu)/(\sigma\sqrt{N})$ tends to the standard normal (zero mean, unit variance)
- for proof, consult the literature
- for large *N*, the central limit theorem tells us that the error one makes in approximating E(X) by A_N is $\sigma/\sqrt{N} = \sqrt{V(X)/N}$
- for Y = f(X) as before, the error in approximating E(f(X)) by $\sum_j f(X_j)/N$ is $\sqrt{V(f(X))/N}$
- use Monte Carlo method to estimate V(f(X))

 $V(Y) = E((Y - E(Y))^2) \approx \langle (f - \langle f \rangle)^2 \rangle = \langle f^2 \rangle - \langle f \rangle^2$

- if *X* is a random variable with probability density $p_X(x)$ and *f* is a well-behaved real-valued function, then Y = f(X) is a random variable
- if p_X(x) can be easily sampled, then use p_X(x) to obtain N outcomes X₁, X₂,..., X_n
- apply function f to obtain random variables $Y_j = f(X_j)$
- law of large numbers tell us that

$$\frac{1}{N}\sum_{j=1}^{N}Y_{j} \xrightarrow{N \to \infty} E(Y) = E(f(X)) = \int_{a}^{b} p_{X}(s) f(s) ds$$

• recap of Monte Carlo integration (uniform sampling):

$$\int_{V} f(\vec{x}) d^{D}x \approx V \langle f \rangle \pm V \sqrt{\frac{\langle f^{2} \rangle - \langle f \rangle^{2}}{N}}$$
$$\langle f \rangle \equiv \frac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i}) \qquad \langle f^{2} \rangle \equiv \frac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i})^{2}$$

- N points x₁,..., x_N chosen independently and randomly with uniform probability dist. throughout D-dimensional volume V
- law of large numbers justifies correctness of estimate
- central limit theorem gives estimate of statistical uncertainty
- in the limit N → ∞, MC estimate tends to normal distribution, uncertainty tends to standard deviation

recap of Monte Carlo integration (non-uniform sampling):

$$\int_{V} p(\vec{x}) f(\vec{x}) d^{D}x \approx \langle f \rangle \pm \sqrt{\frac{\langle f^{2} \rangle - \langle f \rangle^{2}}{N}}$$
$$\langle f \rangle \equiv \frac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i}) \qquad \langle f^{2} \rangle \equiv \frac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i})^{2}$$

- *N* points $\vec{x}_1, \ldots, \vec{x}_N$ chosen independently and randomly with probability dist. $p(\vec{x})$ throughout *D*-dimensional volume *V*
- normalization condition $\int_V p(\vec{x}) d^D x = 1$
- law of large numbers justifies correctness of estimate
- central limit theorem gives estimate of statistical uncertainty

Pseudorandom number generators

- MC integration requires random numbers
- but computers are deterministic!!
- clever algorithms can produce sequences of numbers which appear to be random (pseudorandom)



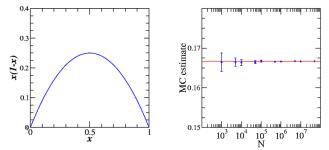
- uniform deviates between 0 and 1
- example: the Mersenne twister
 - http://www.math.sci.hiroshima-u.ac.jp/~m-mat/MT/emt.html
 - currently holds the record for longest period $2^{19937} 1$
 - very fast, passes all standard tests (Diehard) for good RNG
- devising good RNGs is a science in itself
 - most utilize modulus function, bit shifting, shuffling

One-dimensional example

simple example

$$\int_0^1 x(1-x) \, dx = \frac{1}{6} = 0.166666 \cdots$$

plot of integrand and some Monte Carlo estimates



not efficient for 1-dim integrals!

Importance sampling

- Monte Carlo method works best for flat functions, problems when integrand sharply peaked
- importance sampling can greatly improve efficiency of Monte Carlo integration → variance reduction
- recall simple integration

 $\int_{a}^{b} f(x) dx \approx \frac{(b-a)}{N} \sum_{j=1}^{N} f(x_{j}) \quad \begin{array}{l} x_{j} \text{ chosen with uniform probability} \\ \text{between a and b} \end{array}$ • choose function g(x) > 0 with $\int_{a}^{b} g(x) dx = 1$ so $h(x) = \frac{f(x)}{g(x)}$ is as close as possible to a constant $\int_{a}^{b} f(x) dx = \int_{a}^{b} h(x)g(x) dx \approx \frac{(b-a)}{N} \sum_{j=1}^{N} h(x_{j})$ where x_{j} now chosen with probability density g(x)

- must be able to sample with probability density g(x)
- how to choose $g(\vec{x})$ for complicated multi-dimensional integral?

Sampling non-uniform distributions

- random number generators sample the uniform distribution
- to sample other densities, apply transformation method
- random variable U with uniform density $p_U(u) = 1$ for $0 \le x \le 1$
- transform to new random variable $Y = \phi(U)$ where ϕ is a strictly increasing function
 - strictly increasing function ensures inverse function is single-valued
 - also ensures that if u + du > u then y + dy > y for $y = \phi(u)$
- what is density p_Y?
- from conservation of probability

 $p_Y(y)dy = p_U(u)du$ $p_Y(y) = p_U(u)\frac{du}{dy} = p_U(\phi^{-1}(y))\frac{d\phi^{-1}(y)}{dy}$

Sampling non-uniform distributions (continued)

• desired density p_Y is usually known, so must determine ϕ

 $\int_0^u du' = \int_{\phi(0)}^{\phi(u)} p_Y(y) dy \quad \Rightarrow \quad u = F_Y(\phi(u)) \quad \Rightarrow \quad \phi(u) = F_Y^{-1}(u)$

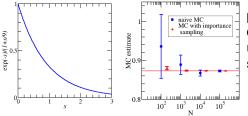
- F^{-1} unique since F is strictly increasing function
- summary: random variable *Y* with density $p_Y(y)$ and cumulative distribution $F_Y(y) = \int_{-\infty}^{y} p_Y(s) ds$ can be sampled by sampling with uniform deviate *U* then applying transformation

 $Y = F_Y^{-1}(U)$

Exponential density

- transformation method requires density whose indefinite integral can be obtained and inverted
- method useful for only a handful of density functions
- one example: the exponential $p_Y(y) = e^{-y}/(1 e^{-b})$ for $0 \le y \le b$
 - cumulative distribution $F_Y(y) = \int_0^y p_Y(s) ds = (1 e^{-y})/(1 e^{-b})$
 - inverse $F_Y^{-1}(u) = -\ln(1 (1 e^{-b})u)$
- example integral:

$$\int_{0}^{r^{3}} \frac{e^{-s} \, ds}{1+s/9} \approx 0.873109$$



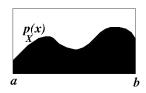
plot of integrand (left); dramatic improvement using importance sampling (right)

Rejection method

- can sample from probability density whose cumulative distribution is not easily calculable and invertible using the rejection method
- sampling from density $p_X(x)$ for $a \le x \le b$

 \Rightarrow equivalent to choosing a random point in two dimensions with uniform probability in the area under curve $p_X(x)$

- simplest method: pick random point with uniform probability in box $a \le x \le b$ horizontally and $0 \le y \le \max(p_X(x))$ vertically
 - accept if below curve
 - reject if above curve, repeat until acceptance



• if $p_X(x)$ sharply peaked, use a comparison function f(x) satisfying $f(x) \ge p_X(x)$ for all $a \le x \le b$ and which can be sampled by transformation method

Integrals of very high dimension

- sampling methods described so far work well in one-dimension
- for multi-dimensional integrals, transformation and rejection methods not feasible
- use of stationary stochastic processes to handle highly multi-dimensional integrals

- stochastic process: a sequence of events X_t, t = 0, 1, 2, ...
 governed by probabilistic laws (limit attention to discrete "time" t)
- consider a system which can be in one of *R* discrete states s_1, s_2, \ldots, s_R (can generalize to continuum of states)
- system moves successively from one state to another
- each move is called a step (discrete "time")
- given previous states of system X₀, X₁,..., X_{t-1}, conditional probability to find system in state X_t at time t is P(X₀,..., X_{t-1}|X_t) which may depend on previous states of system and possibly t
- stochastic processes can be useful for Monte Carlo integration since sequence X_t samples conditional probability distribution $P(X_0, \ldots, X_{t-1}|X_t)$

Stationary stochastic processes

- a stochastic process is stationary when the probabilistic laws remain unchanged through shifts in time
 - joint probability distribution of (X_t, X_{t+j1},..., X_{t+jn}) is same as that of (X_{t+h}, X_{t+h+j1},..., X_{t+h+jn}) for any h
- mean $E(X_t) = \mu$ is independent of t (if it exists)
- variance $E((X_t \mu)^2) = \sigma^2$ independent of t if $E(X_t^2)$ finite
- now X_t are usually not independent random variables
- autocovariance $E((X_t \mu)(X_s \mu)) = R(|t s|)$ depends only on time difference |t s|
- define autocorrelation function ρ(t) = R(t)/R(0) so that ρ(0) = 1 and −1 ≤ ρ(t) ≤ 1 for all t (from Schwartz's inequality)

Law of large numbers for stationary process

- consider a stationary process $X_1, X_2, ...$ with $E(X_k) = \mu$ and autocovariance $R(s) = E((X_k - \mu)(X_{k+s} - \mu))$ satisfying $\sum_{s=0}^{\infty} |R(s)| < \infty$, and define $\overline{X}_N = \frac{1}{N}(X_1 + X_2 + \dots + X_N)$, then for any $\varepsilon > 0$, $\lim_{N \to \infty} P(|\overline{X}_N - \mu| \ge \varepsilon) = 0$
- Proof:

• define $Y_n = X_n - \mu$ and $\overline{Y}_N = \frac{1}{N}(Y_1 + \dots + Y_N)$

$$E(\overline{Y}_{N}^{2}) = \frac{1}{N^{2}}E\left(\sum_{k=1}^{N}Y_{k}^{2}+2\sum_{k
$$= \frac{1}{N^{2}}\left(2\sum_{p=1}^{N}\sum_{h=0}^{p-1}R(h)-NR(0)\right)$$
$$= \frac{1}{N}\left(R(0)+\sum_{k=1}^{N-1}2R(k)(N-k)/N\right)$$$$

Law of large numbers for stationary process (2)

- Proof (continued):
 - continuing,

$$NE(\overline{Y}_{N}^{2}) = |R(0) + \sum_{k=1}^{N-1} 2R(k)(N-k)/N|$$

$$\leq |R(0)| + \sum_{k=1}^{N-1} 2|R(k)| (N-k)/N$$

$$\leq |R(0)| + \sum_{k=1}^{N-1} 2|R(k)|$$

• since $\sum_{j} |R(j)| < \infty$ then $NE(\overline{Y}_N^2) < \infty$ so $\lim_{N \to \infty} E(\overline{Y}_N^2) = 0$

• with Chebyshev inequality $P(|\overline{X}_N - \mu| \ge \varepsilon) \le E((\overline{X}_N - \mu)^2)/\varepsilon^2$ $\lim_{N \to \infty} E((\overline{X}_N - \mu)^2) = 0 \text{ implies } \lim_{\substack{N \to \infty \\ n \to \infty}} P(|\overline{X}_N - \mu| \ge \varepsilon) = 0$

• limiting value $\lim_{N\to\infty} NE((\overline{X}_N - \mu)^2) = \sum_{k=-\infty} R(k)$

• Proof: given absolutely summable autocovariance $\sum_{k} |R(k)| < \infty$, for any $\varepsilon > 0$ there exists a q such that $\sum_{k=1}^{\infty} 2|R(q+k)| < \varepsilon/2$

•
$$\operatorname{so}\left|\sum_{j=-(N-1)}^{N-1} R(j) - NE(\overline{Y}_{N}^{2})\right| = \left|R(0) + 2\sum_{j=1}^{\infty} R(j) - \left(R(0) + \sum_{k=1}^{N-1} 2R(k)(N-k)/N\right)\right|$$

Law of large numbers for stationary process (3)

- limiting value $\lim_{N\to\infty} NE((\overline{X}_N \mu)^2) = \sum_{k=-\infty}^{\infty} R(k)$
 - Proof (continued):

$$\begin{split} \left| \sum_{j=-(N-1)}^{N-1} R(j) - NE(\overline{Y}_{N}^{2}) \right| &= \left| \sum_{k=1}^{N-1} 2kR(k)/N \right| \leq \sum_{k=1}^{N-1} 2k|R(k)| / N \\ &= \sum_{k=1}^{q} 2k|R(k)| / N + \sum_{k=q+1}^{N-1} 2k|R(k)| / N \\ &\leq \sum_{k=1}^{q} 2k|R(k)| / N + \sum_{k=q+1}^{N-1} 2|R(k)| \\ &\leq \sum_{q=1}^{q} 2k|R(k)| / N + \varepsilon/2 \end{split}$$

• since *q* fixed and finite, can always increase *N* so that $\sum_{k=1}^{q} 2k |R(k)| / N < \varepsilon/2 \text{ which holds as } N \to \infty$ • so $\left|\sum_{j=-(N-1)}^{N-1} R(j) - NE(\overline{Y}_{N}^{2})\right| < \varepsilon$ which proves the limit

M-dependent central limit theorem

- *M*-dependent central limit theorem: Let $X_1, X_2, ..., X_N$ be a stationary *M*-dependent sequence of random variables $(X_t \text{ and } X_{t+s} \text{ are independent for } s > M)$ such that $E(X_t) = E(X_1) = \mu$ and $E((X_1 \mu)^2) < \infty$, and define $\overline{X}_N = (X_1 + X_2 + \dots + X_N)/N$ and $\sigma^2 = E((X_1 \mu)^2) + 2\sum_{h=1}^M E((X_1 \mu)(X_{h+1} \mu))$. Then for a < b, $\lim_{N \to \infty} P\left(\frac{a\sigma}{\sqrt{N}} < (\overline{X}_N - \mu) < \frac{b\sigma}{\sqrt{N}}\right) = \frac{1}{\sqrt{2\pi}} \int_a^b e^{-x^2/2} dx$
- distribution of $(X_1 + \cdots + X_N N\mu)/(\sigma\sqrt{N})$ tends to standard normal (zero mean, unit variance)
- for proof, see W. Hoeffding and H. Robbins, Duke Math. Journal 15, 773 (1948) or T. Anderson, The Statistical Analysis of Time Series, Wiley (1971).
- autocovariance $R(h) = R(-h) = E((X_t \mu)(X_{t+|h|} \mu))$

M

• note that
$$\sigma^2 = \sum_{h=-M}^{M} R(h) = NE((\overline{X}_N - \mu)^2)$$
 for $N \gg M$

Monte Carlo integration

Monte Carlo integration using stationary stochastic process:

$$\int_{V} p(\vec{x}) f(\vec{x}) d^{D}x \approx \langle f \rangle \pm \sqrt{\frac{R_{0}(f) + 2\sum_{h \ge 1} R_{h}(f)}{N}}$$
$$\langle f \rangle \equiv \frac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i}), \quad R_{h}(f) \equiv \frac{1}{N-h} \sum_{i=1}^{N-h} \left(f(\vec{x}_{i}) - \langle f \rangle \right) \left(f(\vec{x}_{i+h}) - \langle f \rangle \right)$$

- N points x₁,..., x_N are stationary sequence of random variables with probability dist. p(x) throughout D-dimensional volume V
- normalization condition $\int_V p(\vec{x}) d^D x = 1$
- absolutely summable autocovariance $\sum_{h=0}^{\infty} |R_h(f)| < \infty$
- law of large numbers justifies correctness of estimate
- *m*-dependent central limit theorem gives estimate of statistical uncertainty

Simple stationary stochastic processes

- how can we find a stationary stochastic process that yields the desired probability distribution p(x)?
- use of Markov chains to solve this problem



 Markov chains were introduced by the Russian mathematician Andrei Markov (1856-1922) in 1906

A. A. Mapson (1886).

Markov chains

- <u>discrete Markov chain</u>: stochastic process which generates a sequence of states with probabilities depending only on current state
 - consider a system which can be in one of R states s_1, s_2, \ldots, s_R
 - system moves successively from one state to another
 - each move is called a step (discrete "time")
 - if current state is s_i, then chain moves to state s_j at next step with probability p_{ij} which does not depend on previous states of chain
 - probabilities *p_{ij}* are called transition probabilities
 - the square $R \times R$ real-valued matrix **P** whose elements are p_{ij} is called the transition matrix or the Markov matrix
- time homogeneous if transition probabilities *p_{ij}* independent of "time" or position in chain
- definition generalizes to continuous set of states
 - leads to matrix of transition densities
- will not deal with continuous-time chains here

- transition matrix **P** has non-negative entries $p_{ij} \ge 0$
- since probability of going from s_i to any state must be unity, then matrix elements must satisfy $\sum_{i=1}^{R} p_{ii} = 1$ (rows sum to unity)
- if columns also sum to unity, P is called doubly stochastic matrix
- if P₁ and P₂ are Markov matrices, then the matrix product P₁P₂ is also a Markov matrix
- every eigenvalue λ of a Markov matrix satisfies $|\lambda| \leq 1$
- every Markov matrix has at least one eigenvalue equal to unity

Eigenvalues/eigenvectors of real square matrices

- for a square matrix P, a nonzero column vector v which satisfies Pv = λv for complex scalar λ is known as a right eigenvector corresponding to eigenvalue λ
 - often, "right eigenvectors" are simply called "eigenvectors"
- a nonzero vector **v** satisfying $\mathbf{v}^T \mathbf{P} = \lambda \mathbf{v}^T$, where *T* indicates transpose, is known as a left eigenvector
- every square R × R matrix has R complex eigenvalues, counting multiple roots according to their multiplicity
- for a real square matrix, the eigenvalues are either real or come in complex conjugate pairs
- eigenvectors for distinct eigenvalues are linearly independent
- a degenerate eigenvalue may not have distinct eigenvectors
- *R* linearly independent eigenvectors guaranteed only if all *R* eigenvalues distinct
- a matrix **P** and its transpose \mathbf{P}^T have the same eigenvalues

Properties of Markov matrices (continued)

- every eigenvalue λ of Markov matrix **P** satisfies $|\lambda| \leq 1$
 - Proof: suppose complex number λ is an eigenvalue of **P** with corresponding eigenvector **v** so that $\mathbf{Pv} = \lambda \mathbf{v}$
 - let k be such that $|v_k| \ge |v_j|$ for all j
 - k-th component of eigenvalue equation gives us $\sum_{i} p_{kj} v_j = \lambda v_k$
 - use generalized triangle inequality for complex numbers $|\sum_{i} z_{i}| < \sum_{i} |z_{i}|$ to show

 $\frac{|\sum_{k} z_{k}| \leq \sum_{k} |z_{k}| \text{ to show}}{|\lambda v_{k}| = |\sum_{j} p_{kj} v_{j}| \leq \sum_{j} p_{kj} |v_{j}| \leq \sum_{j} p_{kj} |v_{k}| = |v_{k}|}$

• thus, $|\lambda v_k| = |\lambda| |v_k| \le |v_k| \longrightarrow |\lambda| \le 1$

every Markov matrix P has a least one eigenvalue equal to unity

- Proof: let v be a vector satisfying $v_j = 1$ for all j
- then $\sum_{i} p_{ij} v_j = \sum_{i} p_{ij} = 1 = v_i$
- hence, v is an eigenvector corresponding to eigenvalue 1

Multi-step probabilities

- *n*-step transition probability: *ij*-th element $p_{ij}^{(n)}$ of matrix \mathbf{P}^n is probability that Markov chain, starting in state s_i , will be in state s_j after *n* steps
 - probability to go from s_i to s_j in 2 steps is $\sum_{k=1}^{R} p_{ik} p_{kj}$
 - generalizes to n-steps
- for starting probability vector **u**, probability that chain in state s_j after *n* steps is $u_j^{(n)} = \sum_{i=1}^{R} u_i p_{ij}^{(n)}$

• u_i is probability starting state is s_i , matrix form $\mathbf{u}^{(n)T} = \mathbf{u}^T \mathbf{P}^n$

- first visit probability: the probability that a Markov chain, starting in state s_i , is found for the first time in state s_j after *n* steps \rightarrow denoted by $f_{ij}^{(n)}$
 - define $f_{ij}^{(0)} = 0$ one step, $f_{ij}^{(1)} = p_{ij}$, two steps, $f_{ij}^{(2)} = \sum p_{ik}p_{kj}$
 - generalize $f_{ij}^{(n)} = \sum_{k \neq i} p_{ik} f_{kj}^{(n-1)}$

• important relation for later user: $p_{ij}^{(n)} = \sum_{j=1}^{n} f_{ij}^{(m)} p_{jj}^{(n-m)}$

• total visit probability: probability that, starting from state s_i , chain will ever visit state s_i :

$$f_{ij} = \sum_{n=1}^{\infty} f_{ij}^{(n)}$$

• mean first passage time from s_i to s_j is expected number of steps to reach state s_j in an ergodic Markov chain for the first time, starting from state $s_i \rightarrow$ denoted by m_{ij} (by convention, $m_{ii} = 0$)

$$m_{ij} = \sum_{n=1}^{\infty} n f_{ij}^{(n)}$$

• mean recurrence time μ_i of state s_i is expected number of steps to return to state s_i for the first time in an ergodic Markov chain starting from s_i ∞

$$\mu_i = \sum_{n=1}^{\infty} n f_{ii}^{(n)}$$

Classes

- state s_i is accessible from state s_i if $p_{ij}^{(n)} > 0$ for some finite *n*
 - often denoted by $s_i \rightarrow s_j$
 - if $s_i \rightarrow s_j$ and $s_j \rightarrow s_k$, then $s_i \rightarrow s_k$
- states s_i and s_j communicate if $s_i \rightarrow s_j$ and $s_j \rightarrow s_i$
 - denoted by $s_i \leftrightarrow s_j$

 $-s_i \leftrightarrow s_j$ and $s_j \leftrightarrow s_k$ implies $s_i \leftrightarrow s_k$

class = a set of states that all communicate with one another

- if C_1 and C_2 are communicating classes, then either $C_1 = C_2$ or C_1, C_2 are disjoint
 - if C_1 and C_2 have a common state s_i , then $s_i \leftrightarrow s_{j1}$ for all $s_{j1} \in C_1$ and $s_i \leftrightarrow s_{j2}$ for all $s_{j2} \in C_2$, so $s_{j1} \leftrightarrow s_{j2}$ implying $C_1 = C_2$
- set of all states can be partitioned into separate classes
 - if transition from class C_1 to different class C_2 is possible, then transition from C_2 to C_1 not possible, otherwise $C_1 = C_2$

- a Markov chain is called irreducible if the probability to go from every state to every state (not necessarily in one step) is greater than zero
- all states in irreducible chain are in one single communicating class

Classification of states in Markov chains

states in a Markov chain are

(a) recurrent (persistent) or transient

- recurrent states are either positive or null

(b) periodic (cyclic) or aperiodic

- recurrent or persistent state has $f_{ii} = \sum_{n=1}^{\infty} f_{ii}^{(n)} = 1$
 - unit probability of returning to state after a finite length transient state has $f_{ii} = \sum_{n=1}^{\infty} f_{ii}^{(n)} < 1$
- recurrent state is positive if mean recurrence time finite $\mu_i < \infty$ otherwise, called null
- the period of a state in a Markov chain is the greatest common divisor of all n ≥ 0 for which p_{ii}⁽ⁿ⁾ > 0
 - transition s_i to s_i not possible except for multiples of period d(i)
- periodic state s_i has period d(i) > 1
 aperiodic state s_i has period d(i) = 1

Recurrent and transient states

- for a recurrent state, $\sum_{n=1}^{\infty} p_{ii}^{(n)} = \infty$, whereas for a transient state, $\sum_{n=1}^{\infty} p_{ii}^{(n)} < \infty$
- oproof:
 - we start with the following:

 $\sum_{n=1}^{N} p_{ij}^{(n)} = \sum_{n=1}^{N} \sum_{m=1}^{n} f_{ij}^{(m)} p_{jj}^{(n-m)} = \sum_{m=1}^{N} f_{ij}^{(m)} \sum_{n=0}^{N-m} p_{jj}^{(n)} \le \sum_{m=1}^{N} f_{ij}^{(m)} \sum_{n=0}^{N} p_{jj}^{(n)}$ • but for N > N' we also have $\sum_{n=1}^{N} p_{ij}^{(n)} = \sum_{m=1}^{N} f_{ij}^{(m)} \sum_{n=0}^{N-m} p_{jj}^{(n)} \ge \sum_{m=1}^{N'} f_{ij}^{(m)} \sum_{n=0}^{N-m} p_{jj}^{(n)} \ge \sum_{m=1}^{N'} f_{ij}^{(m)} \sum_{n=0}^{N-N'} p_{jj}^{(n)}$

• putting together above results:

 $\sum_{n=1}^{N'} f_{ij}^{(m)} \sum_{n=1}^{N-N'} p_{jj}^{(n)} \le \sum_{n=1}^{N} p_{ij}^{(n)} \le \sum_{m=1}^{N} f_{ij}^{(m)} \sum_{n=0}^{N} p_{jj}^{(n)}$

• take $N \to \infty$ first, then $N' \to \infty$ to get

$$f_{ij}\sum_{n=0}^{\infty} p_{jj}^{(n)} \leq \sum_{n=1}^{\infty} p_{ij}^{(n)} \leq f_{ij}\sum_{n=0}^{\infty} p_{jj}^{(n)} \quad \Rightarrow \quad f_{ij}\sum_{n=0}^{\infty} p_{jj}^{(n)} = \sum_{n=1}^{\infty} p_{ij}^{(n)}$$

Recurrent and transient states (2)

- for a recurrent state, $\sum_{n=1}^{\infty} p_{ii}^{(n)} = \infty$, whereas for a transient state, $\sum_{n=1}^{\infty} p_{ii}^{(n)} < \infty$
- proof (continued):
 - so far have shown $f_{ij} \sum_{n=0}^{\infty} p_{jj}^{(n)} = \sum_{n=1}^{\infty} p_{ij}^{(n)}$
 - set i = j then $f_{ii}(1 + \sum_{n=1}^{\infty} p_{ii}^{(n)}) = \sum_{n=1}^{\infty} p_{ii}^{(n)}$
 - so finally
 - $f_{ii} = 1$ for a recurrent state and $f_{ii} < 1$ for a transient state, which proves the above statements

 $\sum_{i=1}^{\infty} p_{ii}^{(n)} = \frac{f_{ii}}{1 - f_{ii}}$

note that the above results also imply

$$\sum_{n=1}^{\infty} p_{ij}^{(n)} = \frac{f_{ij}}{1 - f_{ii}}$$

Recurrent and transient states (3)

- a Markov chain returns to a recurrent state infinitely often and returns to a transient state only a finite number of times
- proof:
 - let g_{ij}(m) denote probability that chain enters state s_j at least m times, starting from s_i
 - clearly $g_{ij}(1) = f_{ij}$
 - one also sees $g_{ij}(m+1) = f_{ij}g_{jj}(m)$ so $g_{ij}(m) = (f_{ij})^m$
 - probability of entering s_j infinitely many times is
 - $g_{ij} = \lim_{m \to \infty} g_{ij}(m) = \lim_{m \to \infty} (f_{ij})^m$
 - so starting in s_j then

$$g_{jj} = \lim_{m \to \infty} (f_{jj})^m = \begin{cases} 1 & \text{for recurrent state } f_{jj} = 1\\ 0 & \text{for transient state } f_{jj} < 1 \end{cases}$$

Important result for recurrent states

- if s_i is recurrent and $s_i \rightarrow s_j$, then $f_{ji} = 1$
- proof:
 - let $\alpha > 0$ denote probability to reach s_j from s_i without previously returning to s_i
 - probability of never returning to s_i from s_j is $1 f_{ji}$
 - probability of never returning to s_i from s_i is at least $\alpha(1 f_{ji})$
 - but s_i is recurrent so probability of no return is zero
 - thus, *f_{ji}* = 1
- for two communicating states $s_i \leftrightarrow s_j$ that are each recurrent, it follows that $f_{ij} = f_{ji} = 1$

Similarity of states in a class

- all states in a class of a Markov chain are of the same type, and if periodic, all have the same period
- proof:
 - for any two states s_i and s_j in a class, there exists integers r and s such that p^(r)_{ij} = α > 0 and p^(s)_{ji} = β > 0 so

 $p_{ii}^{(n+r+s)} = \sum_{ik} p_{ik}^{(r)} p_{kl}^{(n)} p_{li}^{(s)} \ge \sum_{ik} p_{ik}^{(r)} p_{ki}^{(n)} p_{ki}^{(s)} \ge p_{ij}^{(r)} p_{jj}^{(n)} p_{ji}^{(s)} = \alpha \beta p_{jj}^{(n)}$

- if s_i is transient, then left-hand side is a term of a convergent series $\sum_k p_{ii}^{(k)} < \infty$, so the same must be true for $p_{jj}^{(k)}$, and if $p_{ii}^{(k)} \to 0$, then $p_{ii}^{(k)} \to 0$
- the same statements remain true if the roles of *i* and *j* are reversed, so either both *s_i* and *s_j* are transient, or neither is
- if s_j is null (infinite mean recurrence time $\mu_j = \sum_{n=1}^{\infty} n f_{jj}^{(n)} = \infty$), then s_i must be null as well
- same statements true if *i*, *j* reversed, so if one is a null state, then so is the other

again, we have

- for any two states s_i and s_j in a class, there exists integers r and s such that p^(r)_{ij} = α > 0 and p^(s)_{ji} = β > 0 so
 - $p_{ii}^{(n+r+s)} = \sum_{ij} p_{ik}^{(r)} p_{kl}^{(n)} p_{li}^{(s)} \ge \sum_{ij} p_{ik}^{(r)} p_{kl}^{(n)} p_{ki}^{(s)} \ge p_{ij}^{(r)} p_{jj}^{(n)} p_{ji}^{(s)} = \alpha \beta p_{jj}^{(n)}$
- suppose s_i has period t, then for n = 0, the right-hand side is positive, so $p_{ii}^{(r+s)} > 0$ which means that r + s must be a multiple of t
- hence, left-hand side vanishes unless *n* is multiple of *t*, so $p_{jj}^{(n)}$ can be nonzero only if *n* is multiple of *t*, so s_i and s_j have same period
- chain aperiodic if p_{ii} > 0 for at least one s_i

Periodic irreducible chains

- states in an irreducible chain with period *d* can be partitioned into *d* mutually exclusive subsets G_0, \dots, G_{d-1} such that if state $s_k \in G_\alpha$, then $p_{1k}^{(n)} = 0$ unless $n = \alpha + \nu d$
- proof:
 - since irreducible, all states have same period d and every state can be reached from every other state
 - there exist for every state sk two integers a and b such that p^(a)_{1k} > 0 and p^(b)_{1k} > 0
 - but $p_{11}^{(a+b)} = \sum_{j} p_{1j}^{(a)} p_{j1}^{(b)} \ge p_{1k}^{(a)} p_{k1}^{(b)} > 0$ so a+b divisible by d
 - thus, a + b = md for integer m, or a = -b + md
 - rewrite as $a = \alpha + \nu d$ for integer ν and $0 \le \alpha < d$
 - α is characteristic of state s_k so all states partitioned into *d* mutually exclusive subsets G_0, G_1, \dots, G_{d-1}
- with proper ordering of G_α subsets, one-step transition from state in G_α always leads to state in G_{α+1}, or from G_{d-1} to G₀
- each subset G_α closed in aperiodic Markov chain with transition matrix P^d

Fact concerning finite Markov chains

- in an irreducible chain having finite number *R* of states, there are no null states and it is impossible that all states are transient
- proof:
 - all rows of the matrix **P**ⁿ must add to unity
 - since each row contains finite number of non-negative elements, it is impossible that p⁽ⁿ⁾_{ii} → 0 for all *i*, *j* pairs
 - thus, impossible that all states are transient
 - so at least one state must be non-null
 - but since irreducible (one class), all states must be non-null
- in an *R*-state irreducible Markov chain, it is possible to go from any state to any other state in at most R 1 steps

A crucial theorem about two sequences

• important theorem: (basic limit theorem of the renewal equation) given a sequence f_0, f_1, f_2, \ldots such that

$$f_0 = 0, \qquad f_n \ge 0, \qquad \sum_{n=0} f_n = 1$$

and greatest common divisor of those *n* for which $f_n > 0$ is $d \ge 1$ and another sequence u_0, u_1, u_2, \ldots defined by

$$u_0 = 1,$$
 $u_n = \sum_{m=1}^n f_m u_{n-m}$ $(n \ge 1)$

then

$$\lim_{n \to \infty} u_{nd} = \begin{cases} d\mu^{-1} & \text{if } \mu = \sum_{n=1}^{\infty} nf_n < \infty \\ 0 & \text{if } \mu = \infty \end{cases}$$

o proof:

- see W. Feller, An Introduction to Probability Theory and Its Applications, Vol. I.
- or S. Karlin and H. Taylor, A First Course in Stochastic Processes.

Basic limit theorem

- we shall only sketch the proof of this theorem
- first, some key properties of these sequences
 - 0 ≤ f_n ≤ 1 for all *n* since f_n ≥ 0 and ∑_{n=0}[∞] f_n = 1
 0 ≤ u_n ≤ 1 for all *n* can be established inductively
 - - $u_0 = 1$, $u_1 = f_1$, $u_2 = f_2 + f_1^2$ satisfy above bounds
 - assume $0 < u_k < 1$ for all 0 < k < n
 - since $f_m \ge 0$ and $\sum_{m=1}^{\infty} f_m = 1$ then $u_{n+1} = \sum_{m=1}^{n+1} f_m u_{n+1-m} \ge 0$ since sum of nonnegative terms, and $u_{n+1} = \sum_{m=1}^{n+1} f_m u_{n+1-m} < \sum_{m=1}^{n+1} f_m < 1$
- next, limit our attention to d = 1 (nonperiodic)
- since u_n is a bounded sequence, $\lambda \equiv \limsup_{n \to \infty} u_n$ is finite and there exists a subsequence $n_1 < n_2 < \cdots$ tending to infinity such that $\lim_{i\to\infty} u_{n_i} = \lambda$
- next step in proof is to show $\lim_{i\to\infty} u_{n_i-q} = \lambda$ for any integer q > 0 when $f_1 > 0$ (we'll skip this)

Basic limit theorem (2)

- define a new sequence $r_n = \sum_{k>n} f_k$
- some important properties of this sequence
 - $r_n > 0$ for all n, and $r_0 = 1$
 - $r_{n-1} r_n = f_n$ for n > 1
 - $\sum_{n=0}^{\infty} r_n = \sum_{n=1}^{\infty} n f_n \equiv \mu$

• one very crucial identity: $\sum_{k=0}^{N} r_k u_{N-k} = 1$ for all N > 0

- define $A_N = \sum_{k=0}^N r_k u_{N-k}$
- start with $u_N = \sum_{m=1}^{N} f_m u_{N-m} = \sum_{m=1}^{N} (r_{m-1} r_m) u_{N-m}$ use $r_0 = 1$ and rearrange $r_0 u_N + \sum_{m=1}^{N} r_m u_{N-m} = \sum_{m=1}^{N} r_{m-1} u_{N-m}$
- take $m \to k+1$ on right: $\sum_{m=0}^{N} r_m u_{N-m} = \sum_{k=0}^{N-1} r_k u_{N-1-k}$
- have shown $A_N = A_{N-1}$ for all N
- $A_N = A_{N-1} = A_{N-2} = \cdots = A_0 = r_0 u_0 = 1$

Basic limit theorem (3)

- recall that $n_1 < n_2 < \cdots$ is subsequence such that $\lim_{j\to\infty} u_{n_j-q} = \lambda$ for any integer $q \ge 0$
- since $\sum_{k=0}^{n_j} r_k u_{n_j-k} = 1$ for all n_j and $r_k \ge 0$, $u_k \ge 0$ for all k, then $\sum_{k=0}^{N} r_k u_{n_j-k} \le 1$ for fixed $N < n_j$
- take limit $j \to \infty$ so $\lim_{j\to\infty} \sum_{k=0}^{N} r_k u_{n_j-k} = \lambda \sum_{k=0}^{N} r_k \le 1$
- already know $\lambda \ge 0$, take $N \to \infty$ to have $0 \le \lambda \le 1/(\sum_{k=0}^{\infty} r_k)$
- if $\sum_{k=0}^{\infty} r_k = \infty$ then $\lim_{n \to \infty} u_n = \lambda = 0$
- if $\mu = \sum_{k=0}^{\infty} r_k$ is finite, $N \to \infty$ gives $\mu \lambda \le 1$
- define $M = \sup_{n>0} u_n$ so $0 \le u_k \le M \le 1$ for all k
- define $g(D) = \sum_{k=D+1}^{\infty} r_k$, note $g(D) \ge 0$ for all D and $\lim_{D\to\infty} g(D) = 0$
- consider $\sum_{k=0}^{D} r_k u_{n_j-k} + \sum_{k=D+1}^{n_j} r_k u_{n_j-k} = 1$ for $D < n_j$
- thus $\sum_{k=0}^{D} r_k u_{n_j-k} + Mg(D) \ge 1$ for $D < n_j$

Basic limit theorem (4)

- again, $\sum_{k=0}^{D} r_k u_{n_j-k} + Mg(D) \ge 1$ for $D < n_j$
- take $j \to \infty$ to conclude $\lambda\left(\sum_{k=0}^{D} r_k\right) + Mg(D) \ge 1$
- take limit $D \to \infty$ to obtain $\lambda \mu \ge 1$
- have now shown $1 \le \mu \lambda \le 1$ so $\mu \lambda = 1$
- proof for nonperiodic (d = 1) case now complete
- when d > 1 then $f_m = 0$ unless m = nd
- can then show $u_m = 0$ unless m = nd
- define new sequences $f'_n = f_{nd}$ and $u'_n = u_{nd}$ for n = 0, 1, 2, ...
- since new sequence aperiodic, know $\lim_{n\to\infty}u_n'=1/\mu'$ where $\mu'=\sum_{n=0}^\infty nf_n'$
- since $f_m = 0$ when $m \neq nd$ then $\mu' = \sum_{n=0}^{\infty} nf_{nd} = d^{-1} \sum_{m=0}^{\infty} mf_m = \mu/d$
- thus, $\lim_{n\to\infty} u_{nd} = d\mu^{-1}$ as required

Asymptotic behavior of $p_{jj}^{(n)}$

• asymptotic behavior of $p_{jj}^{(n)}$ can be summarized as

 $\lim_{n \to \infty} p_{jj}^{(dn)} = \begin{cases} 0 & s_j \text{ transient or null recurrent} \\ \mu_j^{-1} & s_j \text{ aperiodic positive recurrent} \\ d\mu_j^{-1} & s_j \text{ positive recurrent with period } d \end{cases}$

• proof:

- if s_j transient, $\sum_n p_{jj}^{(n)}$ finite (converges) requiring $p_{jj}^{(n)} \rightarrow 0$
- for recurrent s_j , let $f_n = f_{jj}^{(n)}$ and $u_n = p_{jj}^{(n)}$
- sequences f_n , u_n so defined satisfy conditions of basic limit theorem
- basic limit theorem gives $p_{jj}^{(dn)} \to d\mu_j^{-1}$ where $\mu_j = \sum_n n f_{jj}^{(n)}$ is mean recurrence time
- aperiodic case when d = 1
- null recurrent s_j has $\mu_j = \infty$ so $p_{jj}^{(n)} \rightarrow \mu_j^{-1} = 0$

Asymptotic behavior of $p_{ii}^{(n)}$

• asymptotic behavior of $p_{ii}^{(n)}$ can be summarized as

 $\lim_{n \to \infty} p_{ij}^{(n)} = \begin{cases} 0 & s_j \text{ transient or null recurrent} \\ f_{ij}\mu_i^{-1} & s_j \text{ aperiodic positive recurrent} \end{cases}$

- ignore periodic case here
- proof: • $p_{ij}^{(n)} = \sum_{j=1}^{n} f_{ij}^{(m)} p_{jj}^{(n-m)} = \sum_{j=1}^{n'} f_{ij}^{(m)} p_{jj}^{(n-m)} + \sum_{j=1}^{n} f_{ij}^{(m)} p_{jj}^{(n-m)} \quad (n' < n)$ • since $0 \le \sum_{m=n'+1}^{n} f_{ji}^{(m)} p_{ji}^{(n-m)} \le \sum_{m=n'+1}^{n} f_{ji}^{(m)}$ then $0 \le \left(p_{ii}^{(n)} - \sum_{m=1}^{n'} f_{ii}^{(m)} p_{ii}^{(n-m)}\right) \le \sum_{m=n'+1}^{n} f_{ii}^{(m)} \quad (n' < n)$ • take $n \to \infty$, then $n' \to \infty$ above, denote $p_{ii} = \lim_{n \to \infty} p_{ii}^{(n)}$ $0 \leq \left(\lim_{n \to \infty} p_{ij}^{(n)} - p_{jj}f_{ij}\right) \leq 0 \quad \Rightarrow \quad \lim_{n \to \infty} p_{ji}^{(n)} = p_{jj}f_{ij}$
 - for s_i transient or null recurrent, $p_{ii} = 0$ and f_{ii} finite, so $\lim_{n\to\infty} p_{ii}^{(n)} = 0$
 - for s_j aperiod positive recurrent, $p_{ii} = \mu_i^{-1}$ so $p_{ii}^{(n)} \rightarrow f_{ii}\mu_i^{-1}$

Fixed-point or stationary distributions

- a probability vector w is called stationary or invariant or a fixed-point if w^T = w^TP
- clearly, one also has $\mathbf{w}^T = \mathbf{w}^T \mathbf{P}^n$
- the probability vector is always the same (stationary) for the chain
- when this occurs, the Markov chain is said to be in equilibrium

Fatou's lemma

• <u>lemma</u>: let $a_n(t)$ for n = 1, 2, ... be a function on a discrete set $T = \{1, 2, ...\}$, assume $\lim_{n\to\infty} a_n(t)$ exists for each t in T, and suppose $a_n(t) \ge 0$ for all t, n, then

$$\sum_{t\in T} \left(\lim_{n\to\infty} a_n(t)\right) \leq \lim_{n\to\infty} \sum_{t\in T} a_n(t)$$

proof:

• for any integer M

$$\sum_{t=1}^{M} \left(\lim_{n \to \infty} a_n(t) \right) = \lim_{n \to \infty} \sum_{t=1}^{M} a_n(t) \le \lim_{n \to \infty} \sum_{t=1}^{\infty} a_n(t)$$
since all $a_n(t) \ge 0$

• take limit $M \rightarrow \infty$ to obtain required result

• example:
$$a_n(t) = \frac{n}{n^2 + t^2}$$

• for $n > t$ then $\lim_{n \to \infty} a_n(t) = 0$ so $\sum_{t=1}^{\infty} \left(\lim_{n \to \infty} a_n(t)\right) = 0$
• $\sum_{t=1}^{\infty} a_n(t) = \frac{\pi}{2} \coth(n\pi) - \frac{1}{2n}$ so $\lim_{n \to \infty} \sum_{t=1}^{\infty} a_n(t) = \frac{\pi}{2}$

Dominated convergence theorem

• <u>theorem</u>: let $a_n(t)$ for n = 1, 2, ... be a function on a discrete set $T = \{1, 2, ...\}$, assume $\lim_{n\to\infty} a_n(t)$ exists for each t in T, and suppose a function B(t) exists such that $|a_n(t)| \le B(t)$ for all t, n and $\sum_{t\in T} B(t) < \infty$, then

$$\sum_{t\in T} \left(\lim_{n\to\infty} a_n(t)\right) = \lim_{n\to\infty} \sum_{t\in T} a_n(t)$$

• proof:

• let $a(t) = \lim_{n \to \infty} a_n(t)$ and since $|a(t)| \le B(t)$ then $\sum_{t=1} a(t)$ converges • for any integer M

$$\left|\sum_{t=1}^{\infty} a_n(t) - \sum_{t=1}^{\infty} a(t)\right| \le \sum_{t=1}^{M} |a_n(t) - a(t)| + \sum_{t=M+1}^{\infty} \left(|a_n(t)| + |a(t)|\right)$$

• now $\lim_{n \to \infty} \sum_{t=1}^{M} |a_n(t) - a(t)| = \sum_{t=1}^{M} \left(\lim_{n \to \infty} |a_n(t) - a(t)|\right) = 0$
 $\sum_{t=M+1}^{\infty} \left(|a_n(t)| + |a(t)|\right) \le 2\sum_{t=M+1}^{\infty} B(t)$

• so for any integer M

 $\left|\lim_{n\to\infty}\sum_{t=1}^{\infty}a_n(t)-\sum_{t=1}^{\infty}\lim_{n\to\infty}a_n(t)\right|\leq 2\sum_{t=M+1}^{\infty}B(t)$

• right-hand side is remainder of convergent series so equals zero in $M \to \infty$ limit

Fundamental limit theorem for ergodic Markov chains

- Theorem: an irreducible aperiodic Markov chain with transition matrix **P** has a stationary distribution **w** satisfying $w_i > 0$, $\sum_{i} w_{i} = 1$, and $\mathbf{w}^{T} = \mathbf{w}^{T} \mathbf{P}$ if, and only if, all its states are positive recurrent, and this stationary distribution is unique and identical to the limiting distribution $w_i = \lim_{n \to \infty} p_{ii}^{(n)}$ independent of initial state s;
- Proof:
 - o for irreducible aperiodic chain, the following possibilities exist:
 - (a) all states are positive recurrent
 - (b) all states are null recurrent
 - (c) all states are transient
 - if all states transient or null recurrent, $\lim_{n\to\infty} p_{ii}^{(n)} = 0$
 - if all states positive recurrent, then since all states communicate,
 - $f_{ij} = 1$ for all i, j and previous result becomes $\lim_{n \to \infty} p_{ij}^{(n)} = \mu_j^{-1}$ can define $w_j = \lim_{n \to \infty} p_{ij}^{(n)}$ which is independent of initial state s_i
 - for all states positive recurrent, then $0 < \mu_i < \infty$ so $w_i > 0$ for all j

Fundamental limit theorem (2)

- <u>Theorem</u>: an irreducible aperiodic Markov chain with transition matrix **P** has a stationary distribution **w** satisfying $w_j > 0$, $\sum_j w_j = 1$, and $\mathbf{w}^T = \mathbf{w}^T \mathbf{P}$ if, and only if, all its states are positive recurrent, and this stationary distribution is unique and identical to the limiting distribution $w_j = \lim_{n \to \infty} p_{ij}^{(n)}$ independent of initial state s_i
- Proof (continued):
 - we have $p_{ij}^{(m+n)} = \sum_{k=1}^{\infty} p_{ik}^{(n)} p_{kj}^{(m)}$ so using Fatou's lemma: $\lim_{n\to\infty} p_{ij}^{(m+n)} = \lim_{n\to\infty} \sum_{k=1}^{\infty} p_{ik}^{(n)} p_{kj}^{(m)} \ge \sum_{k=1}^{\infty} \lim_{n\to\infty} p_{ik}^{(n)} p_{kj}^{(m)}$
 - taking the limit $n \to \infty$ yields $w_j \ge \sum_{k=1}^{\infty} w_k p_{kj}^{(m)}$
 - define $s \equiv \sum_{k=1}^{\infty} w_k$ then sum above equation over *j*: $s = \sum_{j=1}^{\infty} w_j \ge \sum_{j=1}^{\infty} \sum_{k=1}^{\infty} w_k p_{kj}^{(m)} = \sum_{k=1}^{\infty} w_k \sum_{j=1}^{\infty} p_{kj}^{(m)} = \sum_{k=1}^{\infty} w_k = s$

interchanging order of the two infinite summations is possible since all summands non-negative (Fubini's theorem) $\sim \sim$

since s ≥ s, equality must hold for all j:

$$w_j = \sum_{k=1}^{\infty} w_k p_{kj}^{(m)}$$

Fundamental limit theorem (3)

- Theorem: an irreducible aperiodic Markov chain with transition matrix **P** has a stationary distribution **w** satisfying $w_i > 0$, $\sum_{j} w_{j} = 1$, and $\mathbf{w}^{T} = \mathbf{w}^{T} \mathbf{P}$ if, and only if, all its states are positive recurrent, and this stationary distribution is unique and identical to the limiting distribution $w_i = \lim_{n \to \infty} p_{ii}^{(n)}$ independent of initial state s;
- Proof (continued):
 - have shown $w_i = \sum_{k=1}^{\infty} w_k p_{k_i}^{(m)}$
 - for m = 1, we see the limiting vector w is stationary!!

 - next, from $\sum_{j=1}^{\infty} p_{ij}^{(n)} = 1$ then use Fatou: $1 = \lim_{n \to \infty} \sum_{j=1}^{\infty} p_{ij}^{(n)} \ge \sum_{j=1}^{\infty} \lim_{n \to \infty} p_{ij}^{(n)} = \sum_{j=1}^{\infty} w_j$ given $\sum_j w_j \le 1$ then consider the limit $m \to \infty$ of $w_j = \lim_{m \to \infty} \sum_{k=1}^{\infty} w_k p_{kj}^{(m)}$
 - since $0 \le p_{ki}^{(m)} \le 1$ then $|w_k p_{ki}^{(m)}| \le w_k$ and $\sum_{k=1}^{N} w_k < \infty$ so the dominated convergence theorem can be applied

$$w_j = \lim_{m \to \infty} \sum_{k=1}^{\infty} w_k \, p_{kj}^{(m)} = \sum_{k=1}^{\infty} w_k \lim_{m \to \infty} p_{kj}^{(m)} = \left(\sum_{k=1}^{\infty} w_k\right) w_j$$

• can at last conclude $\sum_{j=1}^{\infty} w_j = 1$

Fundamental limit theorem (4)

- <u>Theorem</u>: an irreducible aperiodic Markov chain with transition matrix **P** has a stationary distribution **w** satisfying $w_j > 0$, $\sum_j w_j = 1$, and $\mathbf{w}^T = \mathbf{w}^T \mathbf{P}$ if, and only if, all its states are positive recurrent, and this stationary distribution is unique and identical to the limiting distribution $w_j = \lim_{n \to \infty} p_{ij}^{(n)}$ independent of initial state s_i
- Proof (continued):
 - only uniqueness of stationary state to show
 - if another stationary vector **v** existed, it would have to satisfy $v_j > 0$, $\sum_{j=1}^{\infty} v_j = 1$, and $v_j = \sum_{i=1}^{\infty} v_i p_{ij}^{(n)}$
 - conditions for dominated convergence theorem again apply, so taking $n \to \infty$ limit gives

$$v_j = \lim_{n \to \infty} \sum_{i=1}^{\infty} v_i p_{ij}^{(n)} = \sum_{i=1}^{\infty} v_i \lim_{n \to \infty} p_{ij}^{(n)} = \left(\sum_{i=1}^{\infty} v_i\right) w_j = w_j$$

• since $\mathbf{v} = \mathbf{w}$, then \mathbf{w} is unique

- consider the following transition matrix $\mathbf{P} = \begin{bmatrix} \frac{2}{4} & \frac{1}{4} & \mathbf{0} \\ 0 & \frac{2}{3} & \frac{1}{3} \\ \frac{1}{4} & \frac{1}{4} & \frac{1}{2} \end{bmatrix}$
- **P**² all positive entries, so chain is irreducible
- eigenvalues are $1, \frac{1}{2}, \frac{5}{12}$
- right and left eigenvectors (unnormalized) are

 $1 \quad \frac{1}{2} \quad \frac{5}{12} \quad 1 \quad \frac{1}{2} \quad \frac{5}{12}$ right: $\begin{bmatrix} 1\\1\\1\\1 \end{bmatrix} \begin{bmatrix} 2\\-2\\-2\\1 \end{bmatrix} \begin{bmatrix} 3\\-4\\-4\\2 \end{bmatrix}$ left: $\begin{bmatrix} 2\\-1\\-1\\-1\\-1\\-4 \end{bmatrix} \begin{bmatrix} -3\\-1\\-1\\-1\\-1\\-1 \end{bmatrix}$ left fixed-point probability vector $\mathbf{w} = \frac{1}{7} \begin{bmatrix} 2 \\ 3 \\ 2 \end{bmatrix} \qquad \lim_{n \to \infty} \mathbf{P}^n = \mathbf{W} = \frac{1}{7} \begin{bmatrix} 2 & 3 & 2 \\ 2 & 3 & 2 \\ 2 & 2 & 3 & 2 \end{bmatrix}$

- positive recurrent chain guarantees existence of at least one invariant probability vector
- irreducibility guarantees uniqueness of invariant probability vector
- aperiodicity guarantees limit distribution coincides with invariant distribution

Equilibrium in Markov chains

- suppose a Markov chain is started with probability vector given by w, the left fixed-point vector of the transition matrix P
- this means the probability of starting in state *s_i* is *w_i*
- the probability of being in state s_j after *n* steps is $(\mathbf{w}^T \mathbf{P}^n)_j$, but $\mathbf{w}^T \mathbf{P}^n = \mathbf{w}^T$, so this probability is w_j
- thus, the probability vector is always the same, that is, it is stationary or invariant
- when this occurs, the Markov chain is said to be in equilibrium
- recall that an ergodic Markov chain which starts in any probability vector y eventually tends to equilibrium
- the process of bringing the chain into equilibrium from a random starting probability vector in known as thermalization

Reversibility in Markov chains

- an ergodic Markov chain is reversible if the probability of going from state s_i to s_j is the same as that for going from state s_j to s_i once the chain is in equilibrium
- the probability that a transition from s_i to s_j occurs is the probability w_i of finding the chain in state s_i in equilibrium times the transition probability p_{ij}
- reversibility occurs when $w_i p_{ij} = w_j p_{ji}$
- the above condition is often referred to as detailed balance
- note that detailed balance guarantees the fixed-point condition: since ∑_i p_{ij} = 1 then

$$\sum_{j} w_{j} p_{ji} = \sum_{j} w_{i} p_{ij} = w_{i}$$

Law of large numbers for Markov chains

- consider an *R*-state ergodic Markov chain which starts in state s_i
- define $X_j^{(m)} = \begin{cases} 1 & \text{if chain in state } s_j \text{ after } m \text{ steps} \\ 0 & \text{otherwise} \end{cases}$ define $N_j^{(n)}$ as number of times chain in state s_j in first n steps

$$N_j^{(n)} = X_j^{(1)} + X_j^{(2)} + \dots + X_j^{(n)}$$

- often called occupation times
- expected value $E(X_i^{(m)}) = p_{ii}^{(m)}$ so $E(N_{i}^{(n)}) = \sum_{h=1}^{n} p_{ii}^{(h)}$
- it can be shown that

$$\lim_{n\to\infty} E(N_j^{(n)})/n = w_j$$

• can show law of large numbers for ergodic Markov chain:

$$P(|N_j^{(n)}/n-w_j|>arepsilon) o 0$$
 as $n o\infty$

Central limit and ergodic theorem for Markov chains

can show a central limit holds

$$\lim_{n \to \infty} P\left(\frac{a\sigma_j}{\sqrt{n}} < \left(\frac{N_j^{(n)}}{n} - w_j\right) < \frac{b\sigma_j}{\sqrt{n}}\right) = \frac{1}{\sqrt{2\pi}} \int_a^b e^{-x^2/2} dx$$

where σ_j depends on w_j

- distributions of random variables $N_i^{(n)}$ tend to normal distributions
- let X₁, X₂,..., X_n be the actual outcomes that make up an ergodic *R*-state Markov chain
- from the definition of $X_j^{(n)}$, it follows that $\sum_{j=1}^R X_j^{(n)} = 1$ so

$$\frac{1}{n}\sum_{h=0}^{n-1}f(X_h) = \frac{1}{n}\sum_{h=0}^{n-1}\sum_{j=1}^{R}X_j^{(h)}f(s_j) = \sum_{j=1}^{R}N_j^{(n)}f(s_j) \to \sum_{j=1}^{R}w_jf(s_j)$$

- Markov-chain "time"-average approaches required ensemble average!!
- already knew this for stationary stochastic processes

Monte Carlo integration

Monte Carlo integration using Markov chain in equilibrium:

$$\int_{V} p(\vec{x}) f(\vec{x}) d^{D}x \approx \langle f \rangle \pm \sqrt{\frac{R_{0}(f) + 2\sum_{h \ge 1} R_{h}(f)}{N}}$$
$$\langle f \rangle \equiv \frac{1}{N} \sum_{i=1}^{N} f(\vec{x}_{i}), \quad R_{h}(f) \equiv \frac{1}{N-h} \sum_{i=1}^{N-h} \left(f(\vec{x}_{i}) - \langle f \rangle \right) \left(f(\vec{x}_{i+h}) - \langle f \rangle \right)$$

• each point in *D*-dim. volume *V* is a state of a Markov chain

- *N* points $\vec{x}_1, \ldots, \vec{x}_N$ are elements of an irreducible aperiodic Markov chain with positive recurrent states and stationary or limiting probability dist. $p(\vec{x})$ throughout *D*-dimensional volume *V*
- Markov chain should be in equilibrium
- normalization condition $\int_V p(\vec{x}) d^D x = 1$
- absolutely summable autocovariance $\sum_{h=0}^{\infty} |R_h(f)| < \infty$

Autocorrelations

- configurations generated by Markov process depend on previous elements in the chain
- this dependence known as autocorrelation
- this autocorrelation can actually be measured!
- for any observable (integrand) O_i , autocorrelation $\varrho(\tau)$ defined by

 $rac{\langle O_i O_{i+ au}
angle - \langle O_i
angle^2}{\langle O_i^2
angle - \langle O_i
angle^2}$

- highly correlated → value near 1
- independent \rightarrow value near 0
- decreasing autocorrelations decreases Monte Carlo error
- dependence decreases as distance between elements in chain increases
 - o do not use every element in chain for "measurements"
 - skip some number of elements between measurements

Constructing the transition probability

- generally know probability density $\pi(\phi)$ we need to sample
- for our path integrals, we need to generate paths with probability distribution $e^{-S[\phi]/\hbar}$

$$\pi(\phi) = \frac{e^{-S[\phi']/\hbar}}{\int_a^b \mathcal{D}\phi' \ e^{-S[\phi']/\hbar}}$$

- in imaginary time formalism, path integral weight is real and positive → probability interpretation for Monte Carlo
- how do we construct the Markov transition matrix $P(\tilde{\phi} \leftarrow \phi)$?
 - change to quantum mechanical notation of putting earlier states on right, later states on left
- simplest answer to this question is

the Metropolis-Hastings method

- useful for local updating so changes to action are small
- probability normalization never enters in the calculation!

The Metropolis-Hastings algorithm

- this method uses an auxiliary proposal density $R(\phi \leftarrow \phi)$ which
 - must be normalized
 - can be evaluated for all $\phi, \widetilde{\phi}$
 - can be easily sampled
 - no relationship to the fixed-point probability density $\pi(\phi)$ needed
- given this proposal density, the Metropolis-Hastings method updates $\phi \to \widetilde{\phi}$ as follows:
 - use $R(\widetilde{\phi} \leftarrow \phi)$ to propose new value $\widetilde{\phi}$ from current value ϕ
 - accept the new value with probability

$$P_{\rm acc}(\widetilde{\phi} \leftarrow \phi) = \min\left(1, \frac{R(\phi \leftarrow \widetilde{\phi})\pi(\widetilde{\phi})}{R(\widetilde{\phi} \leftarrow \phi)\pi(\phi)}\right)$$

(a) if rejected, the original value ϕ is retained

- if proposal density satisfies reversibility R(φ̃ ← φ) = R(φ ← φ̃), then acceptance probability reduces to min(1, π(φ̃)/π(φ))
 - original Metropolis method

Detailed balance in Metropolis-Hastings

- Metropolis-Hastings satisfies detailed balance
- proof:
 - (normalized) transition probability density is

$$W(\widetilde{\phi} \leftarrow \phi) = P_{\rm acc}(\widetilde{\phi} \leftarrow \phi)R(\widetilde{\phi} \leftarrow \phi) + \delta(\widetilde{\phi} - \phi) \left(1 - \int \mathcal{D}\overline{\phi} P_{\rm acc}(\overline{\phi} \leftarrow \phi)R(\overline{\phi} \leftarrow \phi)\right)$$

define

$$\begin{split} A(\widetilde{\phi} \leftarrow \phi) &\equiv P_{\rm acc}(\widetilde{\phi} \leftarrow \phi) R(\widetilde{\phi} \leftarrow \phi) \pi(\phi) \\ &= \min\left(1, \frac{R(\phi \leftarrow \widetilde{\phi})\pi(\widetilde{\phi})}{R(\widetilde{\phi} \leftarrow \phi)\pi(\phi)}\right) R(\widetilde{\phi} \leftarrow \phi)\pi(\phi) \\ &= \min\left(R(\widetilde{\phi} \leftarrow \phi)\pi(\phi), R(\phi \leftarrow \widetilde{\phi})\pi(\widetilde{\phi})\right) \\ \end{split}$$
where last line follows from $R(\widetilde{\phi} \leftarrow \phi)\pi(\phi) \ge 0$

• symmetric:
$$A(\widetilde{\phi} \leftarrow \phi) = A(\phi \leftarrow \widetilde{\phi}).$$

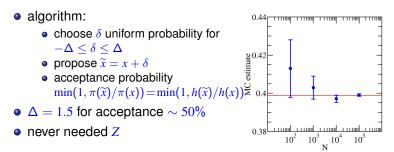
Detailed balance in Metropolis-Hastings (continued)

• so we have $W(\widetilde{\phi} \leftarrow \phi)\pi(\phi) = P_{acc}(\widetilde{\phi} \leftarrow \phi)R(\widetilde{\phi} \leftarrow \phi)\pi(\phi) + \delta(\widetilde{\phi} - \phi)\left(1 - \int \mathcal{D}\overline{\phi} P_{acc}(\overline{\phi} \leftarrow \phi)R(\overline{\phi} \leftarrow \phi)\right)\pi(\phi) = A(\widetilde{\phi} \leftarrow \phi) + \delta(\widetilde{\phi} - \phi)\left(\pi(\phi) - \int \mathcal{D}\overline{\phi} A(\overline{\phi} \leftarrow \phi)\right) = A(\widetilde{\phi} \leftarrow \phi) + \delta(\widetilde{\phi} - \phi) K(\phi)$ where $W(\widetilde{\phi} = \pi(\phi) - \int \mathcal{D}\overline{\phi}A(\overline{\phi} \leftarrow \phi)$

given symmetry of A and Dirac δ-function, then detailed balance holds
 W(φ̃ ← φ)π(φ) = W(φ ← φ̃)π(φ̃)

A one dimensional example

- o does this really work?
- let $g(x) = \cos(\sqrt{1+x^2})$ and $h(x) = \frac{e^{-x^2}}{x^2+2}$
- g(x) changes sign, $h(x) \ge 0$
- consider ratio of integrals $I = \frac{\int_{-\infty}^{\infty} g(x)h(x)dx}{\int_{-\infty}^{\infty} h(x)dx} = 0.3987452$
- sampling density $\pi(x) = Z^{-1}h(x)$ where $Z = \int_{-\infty}^{\infty} h(x)dx$



Part III

Monte Carlo study of the simple harmonic oscillator

Discretization of SHO action

action of harmonic oscillator (imaginary time formalism)

$$S[x(\tau)] = \int_{\tau_a}^{\tau_b} d\tau \left(\frac{1}{2}m\dot{x}^2 + \frac{1}{2}m\omega^2 x^2\right)$$

• discretize time $N\varepsilon = \tau_b - \tau_a$ for Monte Carlo evaluation

$$\frac{S}{\hbar} = \frac{m\varepsilon}{2\hbar} \sum_{j=0}^{N-1} \left[\left(\frac{x_{j+1} - x_j}{\varepsilon} \right)^2 + \omega^2 \left(\frac{x_{j+1} + x_j}{2} \right)^2 \right]$$

- choose ε so discretization errors sufficiently small
- introduce dimensionless parameters

$$x_k = d_k \sqrt{\frac{\varepsilon \hbar}{m}} \qquad \kappa = \frac{1}{4} \varepsilon^2 \omega^2$$
$$\frac{S}{\hbar} = \frac{1}{2} \sum_{j=0}^{N-1} \left[(d_{j+1} - d_j)^2 + \kappa (d_{j+1} + d_j)^2 \right]$$

• a few more manipulations produce

$$rac{S}{\hbar} = rac{1}{2}(1+\kappa)(d_0^2+d_N^2) + (1+\kappa)\left[\sum_{j=1}^{N-1}d_j^2
ight] - (1-\kappa)\left[\sum_{j=0}^{N-1}d_jd_{j+1}
ight]$$

• first constant irrelevant (set to zero), then one last rescaling

$$u_j = d_j \sqrt{1+\kappa}$$
 $g = \frac{1-\kappa}{1+\kappa}$ $d_0 = d_N = 0$

final result for action

$$\frac{S}{\hbar} = \left[\sum_{j=1}^{N-1} u_j^2\right] - g\left[\sum_{j=0}^{N-1} u_j u_{j+1}\right]$$

Metropolis updating of path

to update location (at a single time)

- propose random shift $-\Delta \leq \delta \leq \Delta$ with uniform probability
- calculate change to the action

 $\delta S/\hbar = \delta \left(\delta + 2u_j - g(u_{j-1} + u_{j+1}) \right)$

- accept $u_j^{\text{new}} = u_j + \delta$ with probability $\min(1, e^{-\delta S/\hbar})$
- rule of thumb: fix △ for about 50% acceptance rate
 - lower rate = wasting too much time with rejections
 - higher rate = moving through phase space too slowly
- repeat for each u_j for j = 1, ..., N 1 (this is called one sweep)
- repeat for certain number of sweeps
 - until autocorrelations sufficiently small

Actual C++ code

here is actual C++ code which does the updating

```
void markov::update()
 double shift.deltaS:
 for (int i=1;i<=Nsweeps;i++)</pre>
   for (int t=1;t<Ntimesteps;t++) {</pre>
            // propose shift in location[t]
    shift=2.0*max shift per instance*(rng.generate()-0.5);
            // compute change in action
    deltaS=shift*(shift+2.0*locations[t]
                -hop_param*(locations[t-1]+locations[t+1]));
            // Metropolis accept or reject
    if (deltaS<0.0) accept=1;
    else accept=(rng.generate()<=exp(-deltaS));</pre>
    if (accept) locations[t]+=shift;
```

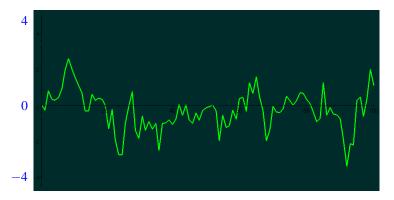
Simulation guidelines

to start Markov chain

- choose a random path (hot start)
- or choose $u_j = 0$ for all j (cold start)
- update *N*_{therm} sweeps until fixed point of chain achieved (thermalization) → check some simple observable
- once thermalized, begin "measurements"
- must choose
 - ε so discretization errors sufficiently small
 - ▲ for adequate acceptance rate
 - N_{sweeps} for sufficiently small autocorrelations
 - N_{meas} for desired precision of results

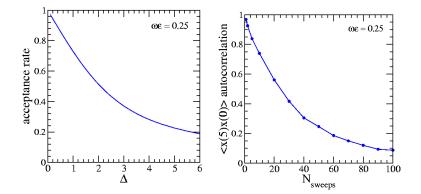
Path animation

• animation of first 100 time slices of *u_j* path



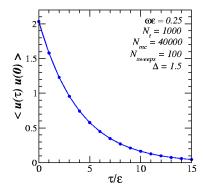
Acceptance rate and autocorrelations

- choose △ so acceptance rate near 0.5
- choose N_{sweeps} so autocorrelations near 0.1



Correlation function

comparison of final Monte Carlo estimates with exact results



- exact result shown as curve
- Monte Carlo estimates shown by circles (statistical uncertainties too small to see)

Part IV

Monte Carlo calculations in real scalar field theory in 2+1 dimensions

Action in continuous space-time

 action in continuous Euclidean *D*-dimensional space-time (imaginary time formalism) given by

$$S = \int d^D x \, \left(\frac{1}{2} \partial_\mu \varphi(x) \partial_\mu \varphi(x) + \frac{1}{2} m^2 \varphi(x)^2 + \frac{g}{4!} \varphi(x)^4 \right).$$

- action must be dimensionless (natural units $\hbar = c = 1$)
- *m* has units of a derivative ∂_{μ} , that is, of a mass
- units of field $[\phi] = [m]^{\frac{1}{2}D-1}$
- coupling g has units $[g] = [m]^{4-D}$
 - coupling dimensionless in 4 space-time dimensions
 - has units of mass in 3 space-time dimensions so g/m dimensionless

- quantization using path integrals
- generalize notion of "path": a path here is a field configuration
- path integral is now integrations over all field configurations
- for real scalar field, integral −∞ ≤ φ(x) ≤ ∞ at every space-time point x
- time-ordered two-point function given by

$$\langle T\phi(x_1)\phi(x_2)\rangle = \frac{\int \mathcal{D}\phi \ \phi(x_1)\phi(x_2) \exp(-S[\phi])}{\int \mathcal{D}\phi \ \exp(-S[\phi])}$$

generalizes to n-point functions: time-ordered product of n fields

Discretization of action

- Monte Carlo study requires action on a space-time lattice
- use anisotropic cubic lattice with temporal lattice spacing a_t and spatial lattice spacing a_s
- use simplest finite difference for the field derivatives
- action is given by

$$S = a_s^{D-1} a_t \sum_{x} \left(\sum_{\mu} \frac{(\varphi(x + a_{\mu}\hat{\mu}) - \varphi(x))^2}{2a_{\mu}^2} + \frac{1}{2}m^2\varphi(x)^2 + \frac{g}{4!}\varphi(x)^4 \right)$$

= $a_s^{D-1} a_t \sum_{x} \left(-\sum_{\mu} \frac{\varphi(x + a_{\mu}\hat{\mu})\varphi(x)}{a_{\mu}^2} + \frac{1}{2} \left(m^2 + \sum_{\nu} \frac{2}{a_{\nu}^2} \right) \varphi(x)^2 + \frac{g}{4!}\varphi(x)^4 \right)$

• redefine the field: $\sqrt{a_s^{D-3}a_t} \varphi(x) = \sqrt{2\kappa_s} \phi(x)$ where κ_s is dimensionless number, new field $\phi(x)$ is dimensionless • a few more dimensionless parameters:

$$a_s/a_t = \zeta, \quad \lambda = \frac{g\zeta\kappa_s^2}{6a_s^{D-4}},$$

$$\kappa_s(a_s^2m^2 + 2\zeta^2 + 2D - 2) = 1 - 2\lambda, \quad \kappa = \zeta\kappa_s$$

final form for lattice action

$$S = \sum_{x} \left(-\frac{2\kappa}{\zeta} \sum_{j=1}^{D-1} \phi(x)\phi(x+a_x\hat{j}) - 2\kappa\zeta \phi(x)\phi(x+a_t\hat{t}) + (1-2\lambda)\phi(x)^2 + \lambda\phi(x)^4 \right)$$

• hopping parameter κ essentially sets mass parameter, λ is interaction coupling

Exact results in free field theory

- the free field theory $\lambda = 0$ is exactly soluable
- path integrals are multivariate gaussians
- free action can be written in form

$$S[\phi] = \frac{1}{2} \sum \phi(x) M(x, y) \phi(y)$$

- for *N* lattice sites, *M* is real and symmetric $N \times N$ matrix having positive eigenvalues
- this matrix given by

$$M(x,y) = -\frac{2\kappa}{\zeta} \sum_{j=1}^{D-1} (\delta(y, x+a_s\hat{j}) + \delta(x, y+a_s\hat{j})) -2\kappa\zeta \left(\delta(y, x+a_t\hat{t}) + \delta(x, y+a_t\hat{t})\right) + 2\delta(x, y)$$

Gaussian integrals in free theory

• N-dimensional multivariate Gaussian integral of form

$$\prod_{i=1}^{N} \left(\int_{-\infty}^{\infty} d\phi_i \right) \exp\left(-\frac{1}{2}\phi_j M_{jk}\phi_k + J_n\phi_n\right)$$
$$= \left(\det\left(\frac{M}{2\pi}\right) \right)^{-1/2} \exp\left(\frac{1}{2}J_j M_{jk}^{-1}J_k\right)$$

J-trick: use derivatives wrt to *J_k*, followed by *J_k* → 0 to evaluate all integrals involving any number of products of the fields

$$\prod_{i=1}^{N} \left(\int_{-\infty}^{\infty} d\phi_i \right) \phi_{m_1} \phi_{m_2} \dots \phi_{m_r} \exp\left(-\frac{1}{2} \phi_j M_{jk} \phi_k\right)$$
$$= \frac{\delta}{\delta J_{m_1}} \cdots \frac{\delta}{\delta J_{m_r}} \prod_{i=1}^{N} \left(\int_{-\infty}^{\infty} d\phi_i \right) \exp\left(-\frac{1}{2} \phi_j M_{jk} \phi_k + J_n \phi_n\right)$$

o does Wick contractions automagically!

Two-point function

- two-point function given by $\langle T\phi(x_1)\phi(x_2)\rangle = M^{-1}(x_1,x_2)$
- invert M by method of Green functions and use Fourier series
- for $L_x \times L_y \times L_t$ lattice, result is

$$M^{-1}(x,y) = \frac{\zeta}{2\kappa L_x L_y L_t} \sum_{k_{\mu}} \frac{\cos(k \cdot (x-y))}{(a_s^2 m^2 + 4\sum_{j=1}^2 \sin^2(\frac{1}{2}k_j) + 4\zeta^2 \sin^2(\frac{1}{2}k_t))}$$

where $k_{\mu} = 2\pi n_{\mu}/L_{\mu}$ for $n_{\mu} = 0, 1, 2, \dots, L_{\mu} - 1$

• pole gives energy $a_t E_p$ of single particle of momentum $a_s p$

$$a_t E_p = 2\sinh^{-1}\left(\frac{1}{2\zeta}\sqrt{a_s^2 m^2 + 4\sin^2(\frac{1}{2}a_s p_x) + 4\sin^2(\frac{1}{2}a_s p_y)}\right)$$

• for small a_t, a_s this becomes $E_p = \sqrt{m^2 + p_x^2 + p_y^2}$

spectrum is sum of free particle energies

- Metropolis-Hastings method needs acceptable acceptance rate
- changing all field values at once generally leads to large changes in action → near zero acceptance rate
- reasonable acceptance rate achieved by updating field at a single lattice site at any given time
- ergodicity ensured by sweeping through lattice, updating each and every site one at a time
- in battle against autocorrelations, expect
 - small wavelength modes updated well
 - long wavelength modes updated not so well

δS for single-site update

recall action is

$$S = \sum_{x} \left(-\frac{2\kappa}{\zeta} \sum_{j=1}^{D-1} \phi(x)\phi(x+a_{x}\hat{j}) - 2\kappa\zeta \phi(x)\phi(x+a_{t}\hat{t}) + (1-2\lambda)\phi(x)^{2} + \lambda\phi(x)^{4} \right)$$

• for $\tilde{\phi} \leftarrow \phi$, change in action is $\delta S = S[\tilde{\phi}] - S[\phi]'$

define neighborhood

$$N(x) = -\frac{2\kappa}{\zeta} \sum_{j=1}^{D-1} \left(\phi(x+a_s\hat{j}) + \phi(x-a_s\hat{j}) \right) - 2\kappa\zeta \left(\phi(x+a_t\hat{t}) + \phi(x-a_t\hat{t}) \right)$$

• if field at one site *x* changed $\phi(x) \rightarrow \phi(x) + \Delta$, then

$$\delta S = \Delta \left(N(x) + (\Delta + 2\phi(x)) \left(1 + \lambda \left((\Delta + 2\phi(x)) \Delta + 2(\phi(x)^2 - 1) \right) \right) \right)$$

• change in action can also be written

$$\delta S = \Delta (a_0 + a_1 \Delta + a_2 \Delta^2 + a_3 \Delta^3),$$

$$a_0 = N(x) + 2\phi(x)(1 + 2\lambda(\phi(x)^2 - 1)),$$

$$a_1 = 1 + 2\lambda(3\phi(x)^2 - 1),$$

$$a_2 = 4\lambda\phi(x),$$

$$a_3 = \lambda$$

Metropolis sweeps

- single-site updates involve a single continuous real variable ϕ
- use simplest proposal density

$$R(\widetilde{\phi} \leftarrow \phi) = \left\{ egin{array}{cc} rac{1}{\Delta_0} & -rac{1}{2}\Delta_0 \leq (\widetilde{\phi} - \phi) \leq rac{1}{2}\Delta_0 \ 0 & |\widetilde{\phi} - \phi| > rac{1}{2}\Delta_0 \end{array}
ight.$$

- width Δ_0 chosen for acceptance probability around 50%
- proposed new value accepted with probability $\min(1, \exp(-\delta S))$
- if rejected, keep current field value
- sweeping through lattice ensures ergodicity
- in sweeping through the lattice in predetermined order, detailed balance no longer holds
 - not a problem since the fixed-point stability condition still holds
 - detailed balance maintained by updating sites in random order

- when the single particle mass $a_t m_{gap}$ is small, the coherence length $\xi = 1/(a_t m_{gap})$ becomes large
- $\xi \to \infty$ signals continuum limit
- $\xi \rightarrow \infty$ occurs near critical point (2nd order phase transition)
- we will see that autocorrelations with Metropolis updating become long ranged as ξ becomes large

 \rightarrow known as critical slowing down

- autocorrelations problematic even for $\xi \approx 5$ with Metropolis
- need help to better update long wavelength modes

Microcanonical updating

- long wavelength modes are associated with lower frequencies, lower energies
- in other words, long-wavelength modes associated with very small changes to the action
- possible way to improve autocorrelations:
 - \rightarrow make large but action preserving $\delta S = 0$ changes

to field at one site

- call this a microcanonical update
 - often referred to as overrelaxation
- Iocal updating is so easy, don't want to give up on it yet!
- must still update in such a way to satisfy detailed balance
- not ergodic, so microcanonical sweeps must be used in combination with ergodic scheme, such as Metropolis sweeps

Microcanonical updating (2)

- we know Metropolis-Hasting method satisfies detailed balance
- choose proposal density strongly peaked about action-preserving value of field, then carefully take δ -function limit
- revisit Metropolis-Hastings with sharply-peaked Breit-Wigner proposal probability density

$$R_f(\widetilde{\phi} \leftarrow \phi) = rac{1}{\pi} rac{arepsilon}{\left(\widetilde{\phi} - f(\phi)
ight)^2 + arepsilon^2}$$

where ε is a constant and $f(\phi)$ is well-behaved, single-valued, invertible function

acceptance probability

$$P_{\rm acc}(\widetilde{\phi} \leftarrow \phi) = \min\left(1, \frac{R_f(\phi \leftarrow \widetilde{\phi})\pi(\widetilde{\phi})}{R_f(\widetilde{\phi} \leftarrow \phi)\pi(\phi)}\right) = \min\left(1, \frac{\left((\widetilde{\phi} - f(\phi))^2 + \varepsilon^2\right)\pi(\widetilde{\phi})}{\left((\phi - f(\widetilde{\phi}))^2 + \varepsilon^2\right)\pi(\phi)}\right)$$

Microcanonical updating (3)

- carefully take $\varepsilon \to 0$ limit: $R_f(\phi \leftarrow \phi) \to \delta(\phi f(\phi))$
- determining acceptance probability is tricky
- probability of proposing a value between

 $f(\phi) - \sqrt{arepsilon} \leq \phi \leq f(\phi) + \sqrt{arepsilon}$ is

$$\int_{f(\phi)-\sqrt{\varepsilon}}^{f(\phi)+\sqrt{\varepsilon}} d\widetilde{\phi} \, R_f(\widetilde{\phi} \leftarrow \phi) = \frac{2}{\pi} \tan^{-1}\left(\frac{1}{\sqrt{\varepsilon}}\right)$$

which does tends to unity as $\varepsilon \rightarrow 0$

if f(φ) more than √ε away from φ, probability transition is actually made is

$$\int_{f(\phi)-\sqrt{\varepsilon}}^{f(\phi)+\sqrt{\varepsilon}} d\widetilde{\phi} \ W_f(\widetilde{\phi} \leftarrow \phi) = \int_{f(\phi)-\sqrt{\varepsilon}}^{f(\phi)+\sqrt{\varepsilon}} d\widetilde{\phi} \ P_{\rm acc}(\widetilde{\phi} \leftarrow \phi) R_f(\widetilde{\phi} \leftarrow \phi)$$
$$= \min\left(\frac{2}{\pi} \tan^{-1}\left(\frac{1}{\sqrt{\varepsilon}}\right), \frac{1}{\pi} \int_{f(\phi)-\sqrt{\varepsilon}}^{f(\phi)+\sqrt{\varepsilon}} \frac{\varepsilon \ \pi(\widetilde{\phi})}{\left((\phi-f(\widetilde{\phi}))^2 + \varepsilon^2\right)\pi(\phi)}\right)$$

Microcanonical updating (4)

• write $\tilde{\phi} = f(\phi) + y$, then remaining integral becomes

$$\frac{1}{\pi} \int_{-\sqrt{\varepsilon}}^{\sqrt{\varepsilon}} dy \frac{\varepsilon \ \pi(f(\phi) + y)}{\left((\phi - f(f(\phi) + y))^2 + \varepsilon^2\right) \pi(\phi)}$$

• if $f(f(\phi)) \neq \phi$, can show this integral goes to zero as $\varepsilon \to 0$

 for self-inverse function f(f(φ)) = φ, expansion about y = 0 must be carefully done, integral has form

$$\frac{\varepsilon}{\pi} \int_{-\sqrt{\varepsilon}}^{\sqrt{\varepsilon}} dy \frac{(a_0 + a_1y + a_2y^2 + \dots)}{(\varepsilon^2 + b_2y^2 + b_3y^3 + b_4y^4 \dots}$$

• must retain $b_2 y^2$ in denominator, expand rest about y = 0:

$$\frac{\varepsilon}{\pi} \int_{-\sqrt{\varepsilon}}^{\sqrt{\varepsilon}} dy \frac{a_0}{(\varepsilon^2 + b_2 y^2)} \left\{ 1 + \frac{a_1}{a_0} y + \frac{a_2}{a_0} y^2 + \left(\frac{a_3}{a_0} - \frac{b_3}{\varepsilon^2}\right) y^3 \dots \right\}$$

• for $b_2 > 0$, result of integration is

$$\frac{2a_0}{\pi\sqrt{b_2}}\tan^{-1}\left(\sqrt{\frac{b_2}{\varepsilon}}\right)\left\{1+d_1\sqrt{\varepsilon}+d_2\varepsilon+d_3\varepsilon^{3/2}+\cdots\right\}$$

Microcanonical updating (5)

• acceptance probability in limit $\varepsilon \to 0$ given by

$$P_{\rm acc} = \min\left(1, \frac{a_0}{\sqrt{b_2}}\right)$$

• here $a_0 = \pi(f(\phi))/\pi(\phi)$ and $b_2 = (f'(f(\phi)))^2$

differentiate both sides of f(f(φ)) = φ with respect to φ, so for self-inverse function

$$1 = \frac{d}{d\phi} \left(f(f(\phi)) \right) = f'(f(\phi)) f'(\phi)$$
$$\frac{1}{(f'(f(\phi)))^2} = \left| \frac{f'(\phi)}{f'(f(\phi))} \right| \quad \text{(self-inverse full)}$$

• take limit $\varepsilon \rightarrow 0$ acceptance probability goes to

$$P_{\rm acc}(\widetilde{\phi} \leftarrow \phi) = \min\left(1, \frac{\sqrt{|f'(\phi)|} \ \pi(\widetilde{\phi})}{\sqrt{|f'(\widetilde{\phi})|} \ \pi(\phi)}\right)$$

nction)

Microcanonical updating (6)

- specialize to action preserving function $f(\phi)$
- for infinitesimal change $\phi \rightarrow \phi + \delta \phi$

$$S(\phi + \delta \phi) = S(f(\phi + \delta \phi))$$

expand both sides

 $S(\phi) + S'(\phi)\delta\phi + O(\delta\phi^2) = S(f(\phi) + f'(\phi)\delta\phi + O(\delta\phi^2))$ = $S(f(\phi)) + S'(f(\phi))f'(\phi)\delta\phi + O(\delta\phi^2)$ = $S(\phi) + S'(f(\phi))f'(\phi)\delta\phi + O(\delta\phi^2).$

- solve order by order in $\delta\phi$ $S'(\phi) = S'(f(\phi))f'(\phi) \rightarrow f'(\phi) = \frac{S'(\phi)}{S'(f(\phi))}, \quad f'(f(\phi)) = \frac{S'(f(\phi))}{S'(\phi)}$
- proposal and acceptance probability densities are

 $R_{f}(\widetilde{\phi} \leftarrow \phi) = \delta\left(\widetilde{\phi} - f(\phi)\right), \quad f(f(\phi)) = \phi, \quad S(f(\phi)) = S(\phi),$ $P_{\text{acc}}(\widetilde{\phi} \leftarrow \phi) = \min\left(1, \left|\frac{S'(\phi)}{S'(\widetilde{\phi})}\right|\right), \quad \pi(\phi) = \frac{\exp(-S[\phi])}{\int \mathcal{D}\widetilde{\phi} \exp(-S[\widetilde{\phi}])}$

Microcanonical updating (7)

- generalize to multiple self-inverse functions
 - for ϕ^4 at most four field values with same local action
- generalize to probability μ of proposing a change
 - sometimes need $\mu < 1$ to prevent (damped) oscillations in autocorrelation function
- summary of microcanonical updating process:
 - **1** decide to propose new field value with probability μ (skip steps below if no proposal)
 - 2 solve $\delta S(\phi) = 0$, let ϕ_i denote real solutions different from ϕ
 - these are roots of a cubic polynomial



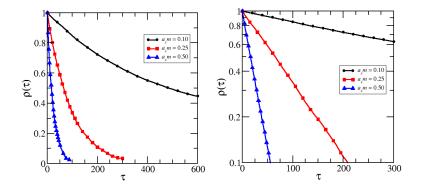
3 randomly choose one of the ϕ_i with equal probability, let $\overline{\phi}$ denote the chosen value

accept with probability

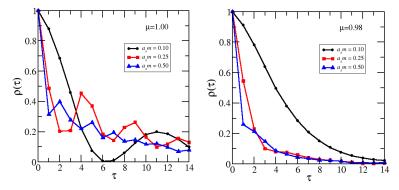
accept with probability

$$P_{\text{acc}}(\widetilde{\phi} \leftarrow \phi) = \min\left(1, \left|\frac{S'(\phi)}{S'(\widetilde{\phi})}\right|\right)$$
if rejected, original value ϕ retained

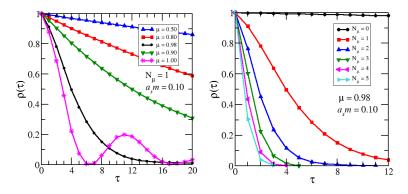
- studied autocorrelation function $\rho(\tau)$ of $\langle \Phi(t)\Phi(0) \rangle$ for $t = 1/(2a_sm)$ and $\Phi(t) = \sum_{xy} \phi(x, y, t)$
- au is number of Metropolis sweeps in plots below
- $a_s m = 0.10, 0.25, 0.50$ for $\lambda = 0$ on 24^3 isotropic lattice
- 2200 sweeps to reduce autocorrelations to 0.1 for $a_s m = 0.10$



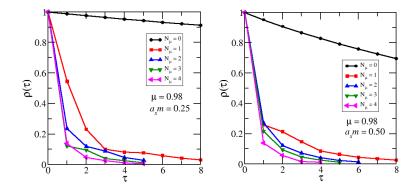
- autocorrelations $\rho(\tau)$ of $\langle \Phi(t)\Phi(0)\rangle$ for $t = 1/(2a_sm)$
- au is number of compound sweeps,
- compound sweep = 1 Metropolis + 1 microcanonical sweep
- μ is probability of proposing change in microcanonical updates
- $a_s m = 0.10, 0.25, 0.50$ for $\lambda = 0$ on 24^3 isotropic lattice
- undesirable oscillations on left removed using $\mu = 0.98$ or updating sites in random order



- autocorrelations $\rho(\tau)$ of $\langle \Phi(t)\Phi(0)\rangle$ for $t = 1/(2a_sm)$
- au is number of compound sweeps
- compound sweep = 1 Metropolis + N_{μ} microcanonical sweeps
- μ is probability of proposing change in microcanonical updates
- $a_s m = 0.10$ for $\lambda = 0$ on 24^3 isotropic lattice
- left-hand plot, $N_{\mu} = 1$ and μ is varied
- right-hand plot, $\mu = 0.98$ and N_{μ} is varied



- autocorrelations $\rho(\tau)$ of $\langle \Phi(t)\Phi(0) \rangle$ for $t = 1/(2a_sm)$
- au is number of compound sweeps
- compound sweep = 1 Metropolis + N_{μ} microcanonical sweeps
- $\mu = 0.98$ probability of proposing change in microcanonical
- $a_s m = 0.25, 0.50$ for $\lambda = 0$ on 24^3 isotropic lattice



Calculating the spectrum

- stationary-state energies extracted from asymptotic decay rates of temporal correlations of the fields
- temporal evolution of field as Heisenberg-picture quantum operator $\phi(t) = e^{Ht}\phi(0)e^{-Ht}$

$$\begin{aligned} \langle 0|\phi(t)\phi(0)|0\rangle &= \sum_{n} \langle 0|e^{Ht}\phi(0)e^{-Ht}|n\rangle\langle n|\phi(0)|0\rangle, \\ &= \sum_{n}^{n} \left|\langle n|\phi(0)|0\rangle\right|^{2} e^{-(E_{n}-E_{0})t} = \sum_{n} A_{n}e^{-(E_{n}-E_{0})t}, \end{aligned}$$

- where complete set of (discrete) eigenstates of H satisfying $H|n\rangle = E_n|n\rangle$ inserted
- if (1|φ(0)|0) ≠ 0, then A₁ and E₁ − E₀ can be extracted as t becomes large, assuming (0|φ(0)|0) = 0
- can use any operator O(t) which is a function of the field φ(t) only on a time slice t

Calculating the spectrum (2)

• extraction of A_1 and $E_1 - E_0$ done using correlated $-\chi^2$

$$\chi^2 = \sum_{tt'} \left(C(t) - M(t,\alpha) \right) \sigma_{tt'}^{-1} \left(C(t') - M(t',\alpha) \right)$$

where C(t) represents Monte Carlo estimates of correlation function with covariance matrix $\sigma_{tt'}$ and model function is $M(t, \alpha) = \alpha_1 e^{-\alpha_0 t}$.

- minimize expression with respect to the model parameters α_0, α_1
- uncertainties in the best-fit parameters $\alpha_0 = E_1 E_0$ and $\alpha_1 = A_1$ are obtained by a jackknife or bootstrap procedure
- fit must be done for a time range t_{min} ≤ t ≤ t_{max} such that an acceptable fit quality is obtained, that is, χ²/dof ≈ 1
- sum of two-exponentials as model function can be used to minimize sensitivity to t_{min}
 - but fit parameters associated with faster-decaying exponential generally *not* good estimates of gap to next energy level and should be discarded

Jackknife resampling

- return to independent trials process X_1, X_2, \ldots, X_N
- expected value E(f(X)) estimated using $\langle f \rangle = \frac{1}{N} \sum_{k=1}^{N} f(X_k)$
- sometimes *f* is a very complicated function, or it could be a function of the expected value!
- $\bullet\,$ propagation of errors often not possible \rightarrow resampling schemes
- let $\langle f \rangle$ denote Monte Carlo estimate of some quantity *f* using all X_k for k = 1, 2, ..., N
- let ⟨f⟩_J denote Monte Carlo estimate of f omitting X_J (so use the other N − 1 values X_k)
- jackknife error estimate given by

$$\sigma^{(J)} = \left(\frac{N-1}{N}\sum_{J=1}^{N}(\langle f \rangle_J - \langle f \rangle)^2\right)^{1/2}$$

- Monte Carlo error formula can be used to determine covariance matrix σ_{tt} for correlation function itself in χ²
- jackknife gives errors in model fit parameters

Bootstrap resampling

- another resampling scheme is the bootstrap
- again, let ⟨f⟩ denote Monte Carlo estimate of some quantity f using all X_k for k = 1, 2, ..., N
- let $\langle f \rangle_b$ denote Monte Carlo estimate of f using a new set \widehat{X}_k for k = 1, 2, ..., N where each \widehat{X}_k is one of the original X_j chosen randomly with equal probability (a bootstrap sample)
- a given X_j can occur multiple times in the bootstrap sample
- obtain large number *B* of such estimates
- let $\langle \widehat{f} \rangle = (1/B) \sum_{b=1} \langle f \rangle_b$
- bootstrap error given by

$$\sigma^{(B)} = \left(\frac{1}{B-1}\sum_{b=1}^{B} (\langle f \rangle_b - \langle \widehat{f} \rangle)^2\right)^{1/2}$$

plot of probability distribution from bootstrap estimates

The effective mass

- particularly good visual tool to see how well energy extracted is so-called effective mass
- for correlator C(t), effective mass defined by

$$m_{\rm eff}(t) = \ln\left(\frac{C(t)}{C(t+a_t)}\right)$$

• function which tends to $E_1 - E_0$ as t becomes large

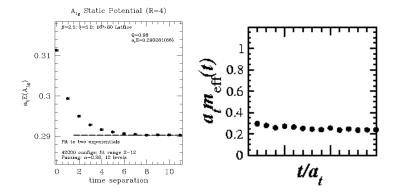
$$\lim_{t \to \infty} m_{\text{eff}}(t) = \lim_{t \to \infty} \ln \left(\frac{A_1 e^{-(E_1 - E_0)t} \left(1 + (A_2/A_1) e^{-(E_2 - E_1)t} + \dots \right)}{A_1 e^{-(E_1 - E_0)(t + a_t)} \left(1 + (A_2/A_1) e^{-(E_2 - E_1)(t + a_t)} + \dots \right)} \right)$$
$$= \ln \left(e^{(E_1 - E_0)a_t} \right) = a_t (E_1 - E_0).$$

• value $E_1 - E_0$ seen as large-time plateau in effective mass

- contributions from faster-decaying exponentials seen as deviations of the effective mass from its asymptotic plateau value
- "good" operator with little coupling to higher-lying states = rapid onset of plateau
- statistically noise generally grows with t

The effective mass (continued)

- two examples of effective masses
- left: static quark-antiquark potential for separation 0.5 fm
- right: nucleon



Excited states from correlation matrices

- extracting more than just the lowest energy in a symmetry channel requires a hermitiam matrix of correlation functions C_{ii}(t)
- let λ_n(t, t₀) denote eigenvalues of C(t₀)^{-1/2} C(t) C(t₀)^{-1/2}, for t₀ some fixed reference time
- these eigenvalues can be viewed as principal correlators
- ordered such that $\lambda_0 \geq \lambda_1 \geq \cdots$ as *t* becomes large
- can show that

$$\lim_{t\to\infty}\lambda_n(t,t_0) = e^{-E_n(t-t_0)}\Big(1+O(e^{-\Delta_n(t-t_0)})\Big),$$
$$\Delta_n = \min_{k\neq n}|E_k-E_n|.$$

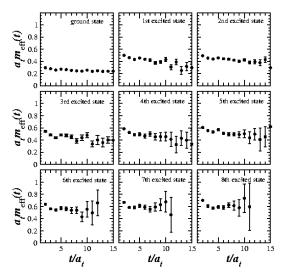
principal effective masses associated with principal correlators

$$m_{ ext{eff}}^{(n)}(t) = \ln\left(rac{\lambda_n(t,t_0)}{\lambda_n(t+a_t,t_0)}
ight)$$

• for $N \times N$ correlation matrix, these functions plateau to N lowest lying energies

Principal effective masses

 LHPC currently holds world record for most energy levels extracted in any lattice QCD computation: 9 in nucleon channel



Spectrum for free scalar field theory

• for free-field case on $N_x \times N_y \times N_t$ lattice, define

$$\Phi(t, n_x, n_y) = \sum_{x,y} \phi(x, y, t) e^{2\pi i x n_x/N_x + 2\pi i n_y/N_y}$$

 lowest six levels having total zero momentum can be extracted using the following set of six operators:

$$O_0(t) = \Phi(t, 0, 0)$$

$$O_1(t) = \Phi(t, 0, 0) \Phi(t, 0, 0)$$

$$O_2(t) = \Phi(t, 1, 0) \Phi(t, -1, 0)$$

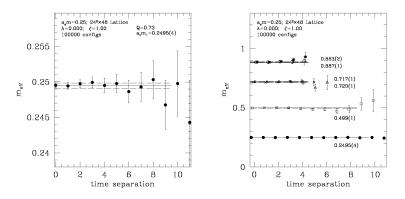
$$O_3(t) = \Phi(t, 0, 1) \Phi(t, 0, -1)$$

$$O_4(t) = \Phi(t, 1, 1) \Phi(t, -1, -1)$$

$$O_5(t) = \Phi(t, 1, -1) \Phi(t, -1, 1)$$

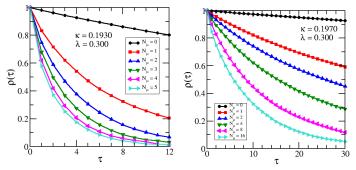
Spectrum for $\lambda = 0$

- extracted six lowest-lying levels in $\lambda = 0$ scalar field theory
- $24^2 \times 48$ isotropic lattice with $a_s m = 0.25$
- exact results: 0.24935 for the mass, 0.49871 for twice the mass, 0.71903 for the two states having minimal relative momenta, and 0.88451 for the next two states



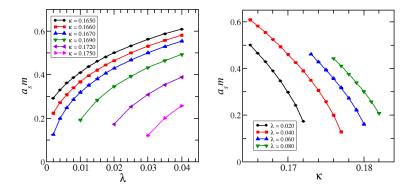
Autocorrelations in the interacting theory

- autocorrelations $\rho(\tau)$ of $\langle \Phi(t)\Phi(0)\rangle$ for $t \sim 1/(2a_s m_{gap})$
- compound sweep = 1 Metropolis + N_{μ} microcanonical sweep
- $\mu = 1$ is probability of proposing change in microcanonical
- left plot: $t = 2a_t$ used with $\kappa = 0.1930$ and $\lambda = 0.300$ on $24^2 \times 48$ isotropic lattice and $a_s m_{gap} \sim 0.25$
- right plot: $t = 5a_t$ used with $\kappa = 0.1970$ and $\lambda = 0.300$ on $32^2 \times 96$ isotropic lattice and $a_s m_{gap} \sim 0.10$
- microcanonical acceptance rate about 80% in both cases



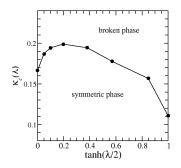
Mass gaps

• various single particle masses on 24³ isotropic lattice



Phase structure

- theory has two phases separated by a line of critical points
- for each value of λ, there exists a critical value κ_c(λ) at which mass gap goes to zero
- symmetric phase for $\kappa < \kappa_c(\lambda)$
 - $\phi
 ightarrow -\phi$ symmetry holds, $\langle \phi
 angle = 0$
- broken phase for $\kappa > \kappa_c(\lambda)$
 - $\phi
 ightarrow -\phi$ spontaneously broken, $\langle \phi
 angle
 eq 0$

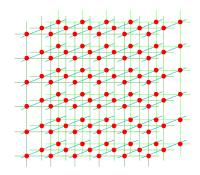


Part V

Monte Carlo calculations in lattice Quantum Chromodynamics

Lattice QCD

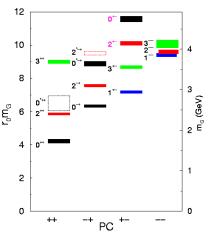
- hypercubic space-time lattice
- quarks reside on sites, gluons reside on links between sites
- for gluons, 8 dimensional integral on each link
- path integral has dimension 32N_xN_yN_zN_t
 - 10.6 million for 24⁴ lattice
- more sophisticated updating algorithms
- systematic errors
 - discretization
 - finite volume



Glueball spectrum in pure gauge theory

- gluons can bind to form glueballs
 - e.m. analogue: massive globules of pure light!
- states labeled by JPC
- scale set by $r_0^{-1} = 410(20) \text{ MeV}$
- computed using pseudo-heatbath and microcanonical
- 24 × 24 correlation matrix in each symmetry channel
- spin identification
- mass gap with a bounty
 - Clay mathematics institute will pay \$ 1 million

C. Morningstar and M. Peardon, Phys. Rev. D 60, 034509 (1999)



- observables in quantum mechanical systems can be extracted from the correlation functions of the theory
- correlation functions can be computed using path integrals
- path integrals in the imaginary time formalism can be evaluated using the Monte Carlo method
- importance sampling from Markov chains
- Metropolis-Hastings method
- microcanonical updating
- 1-dimensional simple harmonic oscillator was first example
- calculations in real scalar ϕ^4 theory in 2 + 1 dimensions

For Further Reading



C.M. Grinstead and J.L. Snell, Introduction to Probability



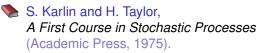
E. Parzen, Stochastic Processes (Holden-Day, San Francisco, 1962).



N.U. Prabhu, Stochastic Processes N.U. (Macmillan, New York, 1965).



📚 I. Montvay and G. Münster Quantum Fields on a Lattice (Cambridge Press, 1994).





S. Hamilton, Time Series Analysis (Princeton University Press, 1994).