Selected topics in Probabilities and Neural Networks: (deeCamp 2018 & Peking University, China)

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Motivations

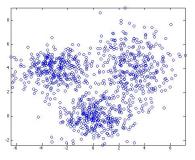
- Probabilities and statistics are a significant part of Machine Learning
- in this class, we discuss about how probabilities and neural networks help each other, under the following four scenarios:
- ▶ This course assumes DeeCamp students are up-to-date with deep learning
- 1. Expectation-Maximization → Matrix Capsule Networks
- 2. Determinantal Point Process → Neural Networks compression
- Kalman Filter ← LSTM
- 4. Probability model estimation \leftarrow Binary classifier
- 5. Probability density re-parameterization
- 6. Stochastic matrices and Monte Carlo Inference



Module one

Expectation-Maximization and Matrix Capsule Networks

A typical motivation of EM - Mixture Densities Model



- When you have data that looks like this figure
- Can you fit them using a single-mode Gaussian distribution, i.e.,:

$$\begin{split} \rho(X) &= \mathcal{N}(X|\mu, \Sigma) \\ &= (2\pi)^{-k/2} |\Sigma|^{-\frac{1}{2}} \exp\left(-\frac{1}{2}(x-\mu)^T \Sigma^{-1}(x-\mu)\right) \end{split}$$

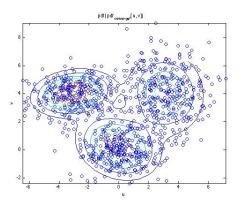
Clearly NOT!

This is typically modeled using Mixture Densities, in the case of Gaussian Mixture Model (k-mixture) (GMM):

$$p(X) = \sum_{l=1}^{k} \alpha_l \mathcal{N}(X|\mu_l, \Sigma_l)$$
 $\sum_{l=1}^{k} \alpha_l = 1$



Gaussian Mixture model result



Let
$$\Theta = \{\alpha_1, \dots \alpha_k, \mu_1, \dots \mu_k, \Sigma_1, \dots \Sigma_k\}$$

$$\begin{split} \Theta_{\mathsf{MLE}} &= \operatorname*{arg\,max}_{\Theta} \mathcal{L}(\Theta|X) \\ &= \operatorname*{arg\,max}_{\Theta} \left(\sum_{l=1}^{n} \log \sum_{l=1}^{k} \alpha_{l} \mathcal{N}(x_{l}|\mu_{l}, \Sigma_{l}) \right) \end{split}$$

- Unlike single mode Gaussian, solving the equation analytically is difficult
- this is where Expectation-Maximization is there to help

The Expectation-Maximization Algorithm

Instead of perform:

$$\theta^{\mathsf{MLE}} = \argmax_{\theta} (\mathcal{L}(\theta)) = \argmax_{\theta} (\log[p(X|\theta)])$$

- ▶ The trick is to assume some "latent" variable Z to the model.
- such that we generate a series of $\Theta = \{\theta^{(1)}, \theta^{(2)}, \dots \theta^{(t)}\}$
- For each iteration of the E-M algorithm, we perform:

$$\Theta^{(g+1)} = \arg\max_{\theta} \left(\int_{Z} \log \left(p(X, Z | \theta) \right) p(Z | X, \Theta^{(g)}) \right) \mathrm{d}z$$

However, we must ensure convergence:

$$\log[p(X|\Theta^{(g+1)})] = \mathcal{L}(\Theta^{(g+1)}) \ge \mathcal{L}(\Theta^{(g)}) \quad \forall i$$



Proof of convergence (1)

$$\begin{split} \mathcal{L}(\theta|X) &= \ln[p(X|\theta)] = \ln[p(Z,X,\theta)] - \ln[p(Z|X,\theta)] \\ &\Longrightarrow \int_{z \in \mathbb{S}} \ln[p(X|\theta)] p(z|X,\Theta^{(g)}) \mathrm{d}z \\ &= \int_{z \in \mathbb{S}} \ln[p(Z,X,\theta)] p(z|X,\Theta^{(g)}) \mathrm{d}z - \int_{z \in \mathbb{S}} \ln[p(Z|X,\theta)] p(z|X,\Theta^{(g)}) \mathrm{d}z \\ &\Longrightarrow \ln[p(X|\theta)] = \underbrace{\int_{z \in \mathbb{S}} \ln[p(Z,X,\theta)] p(z|X,\Theta^{(g)}) \mathrm{d}z}_{Q(\theta,\theta^{(g)})} - \underbrace{\int_{z \in \mathbb{S}} \ln[p(Z|X,\theta)] p(z|X,\Theta^{(g)}) \mathrm{d}z}_{H(\theta,\Theta^{(g)})} \end{split}$$

In E-M, we only maximise, i.e., $\Theta^{(g+1)} = \arg\max_{\theta} Q(\theta, \Theta^{(g)})$. Why? **a trick** If we can prove:

$$\arg\max_{\theta} \left[\int_{z \in \mathbb{S}} \ln[\rho(Z|X,\theta)] \rho(z|X,\Theta^{(g)}) \mathrm{d}z \right] = \Theta^{(g)} \implies H(\Theta^{(g+1)},\Theta^{(g)}) \leq H(\Theta^{(g)},\Theta^{(g)})$$

Then

$$\mathcal{L}(\Theta^{(g+1)}) = \underbrace{Q(\Theta^{(g+1)},\Theta^{(g)})}_{>Q(\Theta^{(g)},\Theta^{(g)})} - \underbrace{H(\Theta^{(g+1)},\Theta^{(g)})}_{< H(\Theta^{(g)},\Theta^{(g)})} \geq Q(\Theta^{(g)},\Theta^{(g)}) - H(\Theta^{(g)},\Theta^{(g)}) = \mathcal{L}(\Theta^{(g)})$$

Proof of convergence (2)

Since $\Phi(.) = -$ In is a convex function:



E-M Example: Gaussian Mixture Model

Gaussian Mixture Model (k-mixture) (GMM):

$$\begin{split} p(X|\Theta) &= \sum_{l=1}^k \alpha_l \mathcal{N}(X|\mu_l, \Sigma_l) \qquad \sum_{l=1}^k \alpha_l = 1 \\ \text{and } \theta &= \{\alpha_1, \dots \alpha_k, \mu_1, \dots \mu_k, \Sigma_1, \dots \Sigma_k\} \end{split}$$

- For data $X = \{x_1, \dots x_n\}$ we introduce "latent" variable $Z = \{z_1, \dots z_n\}$, each z_i indicates which mixture component x_i belong to.
- Looking at the E-M algorithm:

$$\Theta^{(g+1)} = \operatorname*{arg\,max}_{\Theta} \left[Q(\Theta, \Theta^{(g)}) \right] = \operatorname*{arg\,max}_{\Theta} \left(\int_{\mathcal{Z}} \log \left(p(X, Z | \Theta) \right) p(Z | X, \Theta^{(g)}) \mathrm{d}z \right)$$

▶ We need to define both $p(X, Z|\Theta)$ and $p(Z|X, \Theta)$



Gaussian Mixture Model in action

$$p(X|\Theta) = \sum_{l=1}^{k} \alpha_{l} \mathcal{N}(X|\mu_{l}, \Sigma_{l}) = \prod_{l=1}^{n} \sum_{l=1}^{k} \alpha_{l} \mathcal{N}(X|\mu_{l}, \Sigma_{l})$$

How to define $p(X, Z|\Theta)$

$$p(X,Z|\Theta) = \prod_{i=1}^{n} p(x_i,z_i|\Theta) = \prod_{i=1}^{n} \underbrace{p(x_i|z_i,\Theta)}_{\mathcal{N}(\mu_{Z_i},\Sigma_{Z_i})} \underbrace{p(z_i|\Theta)}_{\alpha_{Z_i}} = \prod_{i=1}^{n} \alpha_{Z_i} \mathcal{N}(\mu_{Z_i},\Sigma_{Z_i})$$

Notice that $p(X, Z|\Theta)$ is actually simpler than $p(X|\Theta)$.

How to define $p(Z|X,\Theta)$

$$p(Z|X,\Theta) = \prod_{i=1}^{n} p(z_i|x_i,\Theta) = \prod_{i=1}^{n} \frac{\alpha_{z_i} \mathcal{N}(\mu_{z_i}, \Sigma_{z_i})}{\sum_{l=1}^{k} \alpha_l \mathcal{N}(\mu_l, \Sigma_l)}$$



The E-Step: (1)

$$Q(\Theta, \Theta^{(g)}) = \int_{z} \ln (p(X, Z|\Theta)) p(Z|X, \Theta^{(g)}) dz$$

$$= \int_{z_{1}} \cdots \int_{z_{n}} \left(\sum_{i=1}^{n} \ln p(z_{i}, x_{i}|\Theta) \prod_{i=1}^{n} p(z_{i}|x_{i}, \Theta^{(g)}) \right) dz_{1}, \dots dz_{n}$$

The E-Step: (2)

Knowing,

$$\int_{y_1} \cdots \int_{y_n} \left(\sum_{i=1}^n f_i(y_i) \right) P(Y) dY = \sum_i^N \left(\int_{y_i} f_i(y_i) P_i(y_i) dy_i \right)$$

$$\begin{split} Q(\Theta,\Theta^{(g)}) &= \int_{z_1} \cdots \int_{z_n} \left(\sum_{i=1}^n \ln p(z_i,x_i|\Theta) \prod_{i=1}^n p(z_i|x_i,\Theta^{(g)}) \right) \mathrm{d}z_1, \dots \mathrm{d}z_n \\ &= \sum_{i=1}^n \left(\int_{z_i} \ln p(z_i,x_i|\Theta) p(z_i|x_i,\Theta^{(g)}) \mathrm{d}z_i \right) \qquad z_i \in \{1,\dots,k\} \\ &= \sum_{z_i=1}^k \sum_{i=1}^n \ln p(z_i,x_i|\Theta) p(z_i|x_i,\Theta^{(g)}) \qquad \text{swap the summation terms} \\ &= \sum_{i=1}^k \sum_{i=1}^n \ln \left[\alpha_i \mathcal{N}(x_i|\mu_I,\Sigma_I) \right] p(I|x_i,\Theta^{(g)}) \qquad \text{substitute Gaussian and replace } z_i \to I \end{split}$$

The M-Step objective function

$$\begin{split} Q(\Theta, \Theta^{(g)}) &= \sum_{l=1}^{k} \sum_{i=1}^{n} \ln[\alpha_{l} \mathcal{N}(x_{i} | \mu_{l}, \Sigma_{l})] p(l | x_{i}, \Theta^{(g)}) \\ &= \sum_{l=1}^{k} \sum_{i=1}^{n} \ln(\alpha_{l}) p(l | x_{i}, \Theta^{(g)}) + \sum_{l=1}^{k} \sum_{i=1}^{n} \ln[\mathcal{N}(x_{i} | \mu_{l}, \Sigma_{l})] p(l | x_{i}, \Theta^{(g)}) \end{split}$$

- \blacktriangleright The first term contains only α
- rightarrow second term contains only μ , Σ.
- So we can maximize both terms independantly.

The M-Step: maximizing α

Maximizing α means that:

$$\frac{\partial \sum_{l=1}^k \sum_{i=1}^n \ln(\alpha_l) p(l|x_i, \Theta^{(g)})}{\partial \alpha_1, \dots, \partial \alpha_k} = [0 \dots 0] \quad \text{ subject to } \sum_{l=1}^k \alpha_l = 1$$

This is to be solved using Lagrange Multiplier

$$\mathbb{LM}(\alpha_1, \dots \alpha_k, \lambda) = \sum_{l=1}^k \ln(\alpha_l) \underbrace{\left(\sum_{i=1}^n p(I|x_i, \Theta^{(g)})\right)}_{\text{contains no } \alpha} - \lambda \left(\sum_{l=1}^k \alpha_l - 1\right)$$

$$\implies \frac{\partial \mathbb{LM}}{\partial \alpha_l} = \frac{1}{\alpha_l} \left(\sum_{i=1}^n p(I|x_i, \Theta^{(g)})\right) - \lambda = 0$$

$$\implies \alpha_l = \frac{1}{N} \sum_{i=1}^n p(I|x_i, \Theta^{(g)})$$

The M-Step: maximizing μ, Σ

second part of
$$Q(\Theta, \Theta^{(g)})$$

$$\begin{split} &\equiv \mathcal{S}(\mu_{l}, \mathbf{\Sigma}_{l}) = \sum_{l=1}^{k} \sum_{i=1}^{n} \ln[\mathcal{N}(x_{i}|\mu_{l}, \mathbf{\Sigma}_{l})] p(l|x_{i}, \Theta^{(g)}) \\ &= \sum_{i=1}^{n} \sum_{l=1}^{k} \ln\left(\frac{1}{\sqrt{(2\pi)^{d}|\mathbf{\Sigma}_{l}|}} \exp\left(-\frac{1}{2}(x_{i} - \boldsymbol{\mu})^{\top} \mathbf{\Sigma}^{-1}(x_{i} - \boldsymbol{\mu})\right)\right) p(l|x_{i}, \Theta^{(g)}) \\ &= \sum_{i=1}^{n} -\frac{1}{2} \ln(|\mathbf{\Sigma}_{l}|) p(l|x_{i}, \Theta^{(g)}) - \sum_{i=1}^{n} \frac{1}{2}(x_{i} - \boldsymbol{\mu}_{l})^{\top} \mathbf{\Sigma}^{-1}(x - \boldsymbol{\mu}_{l}) p(l|x_{i}, \Theta^{(g)}) \end{split}$$

how do we maximize μ , Σ ?

Some formulas to remember, from Matrix Cookbook

derivatives of log of determinant (with determinant)

$$\frac{\partial \ln |\mathbf{X}|}{\partial \mathbf{X}} = (\mathbf{X}^{-1})^{\top}$$

Derivatives of Traces

$$\frac{\partial \operatorname{tr}(F(\mathbf{X}))}{\partial \mathbf{X}} = (f(\mathbf{X}))^{\top}$$

where $f(\cdot)$ is the scalar derivative of $F(\cdot)$

Derivatives of Traces of inverse, fact 1

$$\frac{\partial \text{tr}(\mathbf{AXB})}{\partial \mathbf{X}} = \mathbf{A}^{\top} \mathbf{B}^{\top}$$

Derivatives of Traces of inverse, fact 2

$$\frac{\partial \text{tr}((\mathbf{X} + \mathbf{A})^{-1})}{\partial \mathbf{X}} = -((\mathbf{X} + \mathbf{A})^{-1}(\mathbf{X} + \mathbf{A})^{-1})^{\top}$$

Derivatives of Traces of inverse, fact 3

$$\frac{\partial tr(\mathbf{A}\mathbf{X}^{-1}\mathbf{B})}{\partial \mathbf{X}} = -(\mathbf{X}^{-1}\mathbf{B}\mathbf{A}\mathbf{X}^{-1})^{\top}$$



Maximization μ_I

$$S(\mu_{l}, \mathbf{\Sigma}_{l}) = \sum_{i=1}^{n} -\frac{1}{2} \ln(|\mathbf{\Sigma}_{l}|) p(l|x_{i}, \boldsymbol{\Theta}^{(g)}) - \sum_{i=1}^{n} \frac{1}{2} (x_{i} - \mu_{l})^{T} \mathbf{\Sigma}^{-1} (x - \mu_{l}) p(l|x_{i}, \boldsymbol{\Theta}^{(g)})$$

$$= -\text{Tr} \left(\frac{\sum_{l}^{-1}}{2} \sum_{i=1}^{n} (x_{i} - \mu_{l}) (x - \mu_{l})^{T} p(l|x_{i}, \boldsymbol{\Theta}^{(g)}) \right) + \text{ Constant}$$

$$\Rightarrow \frac{\partial S(\mu_{l}, \mathbf{\Sigma}_{l})}{\partial \mu_{l}} = \frac{\sum_{i=1}^{-1}}{2} \sum_{i=1}^{n} 2(x_{i} - \mu_{l}) p(l|x_{i}, \boldsymbol{\Theta}^{(g)}) = 0$$

$$\Rightarrow \sum_{i=1}^{n} x_{i} p(l|x_{i}, \boldsymbol{\Theta}^{(g)}) = \mu_{l} \sum_{i=1}^{n} p(l|x_{i}, \boldsymbol{\Theta}^{(g)})$$

$$\Rightarrow \mu_{l} = \frac{\sum_{i=1}^{n} x_{i} p(l|x_{i}, \boldsymbol{\Theta}^{(g)})}{\sum_{l=1}^{n} p(l|x_{l}, \boldsymbol{\Theta}^{(g)})}$$

Maximization Σ_I

- let's try something **cool**, and try to obtain the answer using matrix form
- let $\mathcal Y$ be zero-meaned data matrix, where each column of $\mathcal Y$ is $x_i \mu_I$
- let **P** be diagonal matrix in which P_{ii} correspond to $p(I|x_i, \Theta^{(g)})$

$$\begin{split} \mathcal{S}(\boldsymbol{\mu}_{l}, \boldsymbol{\Sigma}_{l}) &= \sum_{i=1}^{n} -\frac{1}{2} \ln(|\boldsymbol{\Sigma}_{l}|) p(\boldsymbol{I}|\boldsymbol{x}_{i}, \boldsymbol{\Theta}^{(g)}) - \sum_{i=1}^{n} \frac{1}{2} (\boldsymbol{x}_{i} - \boldsymbol{\mu}_{l})^{T} \boldsymbol{\Sigma}^{-1} (\boldsymbol{x} - \boldsymbol{\mu}_{l}) p(\boldsymbol{I}|\boldsymbol{x}_{i}, \boldsymbol{\Theta}^{(g)}) \\ &= -\frac{\text{tr}(\boldsymbol{P})}{2} \ln|\boldsymbol{\Sigma}_{l}| - \frac{1}{2} \text{tr}(\boldsymbol{\Sigma}_{l}^{-1} \boldsymbol{\mathcal{Y}} \boldsymbol{P} \boldsymbol{\mathcal{Y}}^{\top}) \end{split}$$

now taking the derivative:

$$\begin{split} \frac{\partial \mathcal{S}(\mu_{l}, \boldsymbol{\Sigma}_{l})}{\partial \boldsymbol{\Sigma}_{l}} &= \boldsymbol{\Sigma}_{l}^{-1} \mathcal{Y} \boldsymbol{P} \mathcal{Y}^{\top} \boldsymbol{\Sigma}_{l}^{-1} - \operatorname{tr}(\boldsymbol{P}) \boldsymbol{\Sigma}_{l}^{-1} = \boldsymbol{0} \\ \Longrightarrow \boldsymbol{\Sigma}_{l}^{-1} \mathcal{Y} \boldsymbol{P} \mathcal{Y}^{\top} \boldsymbol{\Sigma}_{l}^{-1} &= \operatorname{tr}(\boldsymbol{P}) \boldsymbol{\Sigma}_{l}^{-1} \\ \Longrightarrow \boldsymbol{\Sigma}_{l}^{-1} \mathcal{Y} \boldsymbol{P} \mathcal{Y}^{\top} &= \operatorname{tr}(\boldsymbol{P}) \\ \Longrightarrow \boldsymbol{\Sigma}_{l}^{-1} &= \operatorname{tr}(\boldsymbol{P}) (\mathcal{Y} \boldsymbol{P} \mathcal{Y}^{\top})^{-1} \\ \Longrightarrow \boldsymbol{\Sigma}_{l} &= \operatorname{tr}(\boldsymbol{P})^{-1} (\mathcal{Y} \boldsymbol{P} \mathcal{Y}^{\top}) &= \frac{(\mathcal{Y} \boldsymbol{P} \mathcal{Y}^{\top})}{\operatorname{tr}(\boldsymbol{P})} \\ \text{or, } \boldsymbol{\Sigma}_{l} &= \frac{\sum_{i=1}^{n} (x_{i} - \mu_{l})(x - \mu_{l})^{T} \boldsymbol{p}(l|x_{i}, \boldsymbol{\Theta}^{(g)})}{\sum_{i=1}^{n} \boldsymbol{p}(l|x_{i}, \boldsymbol{\Theta}^{(g)})} \end{split}$$

Summary of Gaussian Mixture Model

► Maximizing μ , Σ , α to update Θ ^(g) \rightarrow Θ ^(g+1):

$$\alpha_{I}^{(g+1)} = \frac{1}{N} \sum_{i=1}^{N} p(I|x_{i}, \Theta^{(g)})$$

$$\mu_{l}^{(g+1)} = \frac{\sum_{i=1}^{N} x_{i} p(l|x_{i}, \Theta^{(g)})}{\sum_{i=1}^{N} p(l|x_{i}, \Theta^{(g)})}$$

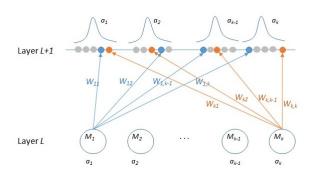
$$\Sigma_{l}^{(g+1)} = \frac{\sum_{i=1}^{N} [x_{i} - \mu_{l}^{(i+1)}][x_{i} - \mu_{l}^{(i+1)}]^{T} \rho(l|x_{i}, \Theta^{(g)})}{\sum_{i=1}^{N} \rho(l|x_{i}, \Theta^{(g)})}$$

responsibility probabilities:

$$p(I|x_i, \Theta^{(g)}) = \frac{\alpha_I \mathcal{N}(x_i|\mu_I, \Sigma_I)}{\sum_{s=1}^k \alpha_s \mathcal{N}(x_i|\mu_s, \Sigma_s)}$$



- Introduction of Capsule Networks, dynamic Routing and Matrix Capsule
- "Hinton., et. al, (2018), Matrix capsules with EM routing



assume both lower layer L and higher layer L+1 have K neurons

- lower layer $\{M_1, \ldots M_k\}$ acts like data source
- ightharpoonup each M_i generates multiple "new data" V_{ij} through its transformation matrix W_{ij} , i.e.,

$$V_{ij} = M_i W_{ij}$$

▶ E-M then tries to fit $\Theta \equiv \{\mu_1, \dots, \mu_k, \sigma_1, \dots, \sigma_k\}$ at layer L + 1 capsules.



M-STEP:

$$\begin{aligned} \forall i \in \Omega_L : R_{ij} \leftarrow R_{ij} \times a_i & \text{apply } a_i \\ \forall h : \mu_j^h \leftarrow \frac{\sum_i R_{ij} V_{ij}^h}{\sum_i R_{ij}} \\ \forall h : \left(\sigma_j^h\right)^2 \leftarrow \frac{\sum_i R_{ij} (V_{ij}^h - \mu_j^h)^2}{\sum_i R_{ij}} \\ & \text{cost}^h \leftarrow \left(\beta_u + \log(\sigma_j^h)\right) \sum_i R_{ij} & a_j \leftarrow \text{sigmoid} \left(\lambda(\beta_a - \sum_h \text{cost}^h)\right) & \text{update } a_j \end{aligned}$$

► E-STEP:

$$\forall j \sim \Omega_{L+1} : \mathbf{p}_j \leftarrow \prod_{h=1}^H \mathcal{N}(V_{ij}^h; \mu_j^h, \sigma_j^h)$$

$$\mathbf{R}_{ij} \leftarrow \frac{a_j p_j}{\sum_{k \in \Omega_{L+1}} a_k p_k} \quad \text{apply } a_j$$

 \triangleright cost \mathcal{C}_i^h used as argument of sigmoid(.) and weighted from all lower capsules

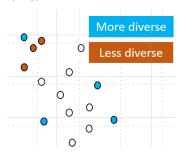
$$\begin{split} \mathcal{C}_{j}^{h}(\mu,\sigma) &= \sum_{i} R_{ij} \mathcal{C}_{ij}^{h} \\ &= \sum_{i} -R_{ij} \ln(\rho_{i|j}^{h}) \\ &= \sum_{i} R_{ij} \left(\frac{(\boldsymbol{v}_{ij}^{h} - \boldsymbol{\mu}_{j}^{h})^{2}}{2(\sigma_{j}^{h})^{2}} + \ln(\sigma_{j}^{h}) + \frac{\ln(2\pi)}{2} \right) \\ &= \sum_{i} R_{ij} \left(\frac{(\overline{\sigma}_{j}^{h})^{2}}{2(\sigma_{j}^{h})^{2}} + \ln(\sigma_{j}^{h}) + \frac{\ln(2\pi)}{2} \right) \\ &\approx \left(\ln(\sigma_{j}^{h}) + k \right) \sum_{i} R_{ij} \quad k \text{ is a constant} \end{split}$$

Module two

Determinantal Point Process and Neural Networks Compression

What is DPP?

- ▶ in **one sentence**: it's a probability defined on any subsets of *N* data points, such that a diverse subset attracts higher probability
- so, if it's a probability model, what is its parameter?
- we can either use a marginal kernel K, or to use an L-ensemble L
- let's look at marginal kernel first:



How do we define a DPP?

Start with a marginal distribution of subset A

$$\Pr(A \subseteq \mathbf{Y}) = \det(K_A)$$

• An example: given $\mathcal{Y} = \{1, 2, 3, 4, 5\}, A = \{1, 2, 3\}$

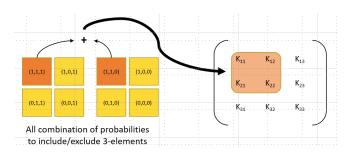
$$\begin{split} \text{Pr}(A \subseteq \textbf{Y}) &\equiv \text{Pr}(A \subseteq \textbf{Y} \subseteq \mathcal{Y}) \equiv \text{Pr}\left(\textbf{Y} \in \{\{1,2,3\}, \{1,2,3,4\}, \{1,2,3,5\}, \{1,2,3,4,5\}\}\right) \\ &= \text{det}(\mathcal{K}_A) \end{split}$$

$$Pr(A \subseteq \mathbf{Y}) \equiv Pr(A \subseteq \mathbf{Y} \subseteq \mathcal{Y}) \equiv Pr(y_1 = 1, y_2 = 1, y_3 = 1)$$

$$= \sum_{t_4=0}^{1} \sum_{t_5=0}^{1} Pr(y_1 = 1, y_2 = 1, y_3 = 1, y_4 = t_4, y_5 = t_5)$$

$$= det(K_A)$$

let's use a diagram!



- sum of probabilities to include 1st and 2nd elements
- ▶ This is defined by $det(K_{\{1,2\}})$

Something about marginal distribution

- ▶ $Pr(A \subseteq Y)$ is marginal, they don't need to add to 1
- ▶ it may be possible that, $Pr(A_1 \subseteq Y) + Pr(A_2 \subseteq Y) > 1$
- ▶ $Pr(\emptyset \subseteq Y) = det(K_{\emptyset}) = 1$ This is obvious, as any Y is a superset of \emptyset .
- $ightharpoonup \Pr(i \subseteq \mathbf{Y}) = \det(K_{ii}) = K_{ii}$
- Look at the two element case: this is the key

$$Pr(i, j \in \mathbf{Y}) = \begin{vmatrix} K_{ii} & K_{ij} \\ K_{ij} & K_{ji} \end{vmatrix}$$
$$= K_{ii}K_{jj} - K_{ij}K_{ij}$$
$$= Pr(i \subseteq \mathbf{Y}) Pr(j \subseteq \mathbf{Y}) - K_{ij}^{2}$$

- Off-diagonal elements determine negative correlations between pairs.
- ► Large values of K_{ij} imply i and j tend **not** co-occur



Example of K does NOT define DPP

- ▶ Any K, $0 \leq K \leq I$ defines a DPP.
- If $K \preceq K'$, that is, K' K is positive semidefinite. Therefore, $\begin{bmatrix} 1 & -0.5 \\ -0.5 & 1 \end{bmatrix}$ can NOT define DPP, as

$$\left| I - \begin{pmatrix} 1 & -0.5 \\ -0.5 & 1 \end{pmatrix} \right| = \begin{pmatrix} 0 & 0.5 \\ 0.5 & 0 \end{pmatrix} \implies \bar{\lambda} = [-0.5, 0.5]^{\top}$$

Another way to see the above is incorrect: $y = \{1, 2\}$

$$\begin{split} \Pr(A = \{1\} \subseteq \textbf{Y}) &\equiv \Pr\left(\textbf{Y} \in \{\{1\}, \{1, 2\}\}\right) \\ &= \det(K_1) = 1 \\ \Pr(A = \{2\} \subseteq \textbf{Y}) &\equiv \Pr\left(\textbf{Y} \in \{\{2\}, \{1, 2\}\}\right) \\ &= \det(K_2) = 1 \end{split}$$

However,

$$Pr(A = \{1, 2\} \subseteq Y) \equiv Pr(Y \in \{\{1, 2\}\})$$

= $det(K_{\{1, 2\}}) = 0.75$

The first two equation says $\{1\}$ and $\{2\}$ must be included; The third equation says both may NOT always be included.



Example of K define DPP

- ▶ Any K, $0 \leq K \leq I$ defines a DPP.
- ▶ If $K \leq K'$, that is, K' K is positive semidefinite.

$$\begin{bmatrix} 0.3 & -0.1 \\ -0.1 & 0.4 \end{bmatrix}$$
 can define DPP:

$$\left| I - \begin{pmatrix} 0.3 & -0.1 \\ -0.1 & 0.4 \end{pmatrix} \right| = \left| \begin{pmatrix} 0.7 & 0.1 \\ 0.1 & 0.6 \end{pmatrix} \right| \implies \bar{\lambda} = [0.5382, 0.7618]^{\top}$$

 $\mathcal{Y} = \{1, 2\}$

$$\begin{aligned} \Pr(A = \{1\} \subseteq \mathbf{Y}) &\equiv \Pr(Y \in \{\{1\}, \{1, 2\}\}) \\ &= \det(K_1) = 0.3 \\ \Pr(A = \{2\} \subseteq \mathbf{Y}) &\equiv \Pr(Y \in \{\{2\}, \{1, 2\}\}) \\ &= \det(K_2) = 0.4 \\ \Pr(A = \{1, 2\} \subseteq \mathbf{Y}) &\equiv \Pr(Y \in \{\{1, 2\}\}) \\ &= \det(K_{\{1, 2\}}) = 0.11 \end{aligned}$$

So where do rest of probabilities go?

$$Pr(A = \emptyset \subseteq Y) \equiv Pr(Y \in \{\emptyset, \{1\}, \{2\}, \{1, 2\}\})$$
$$= det(K_{\emptyset}) = 1$$

Some probabilities mass is assigned to ∅.



L-Ensembles

- Marginal distributions does not define probabilities in terms of a particular set
- ▶ i.e., instead of having $Pr(Y \subseteq Y)$, we want Pr(Y = Y)

$$\Pr_L(\mathbf{Y} = Y) \propto \det(L_Y)$$

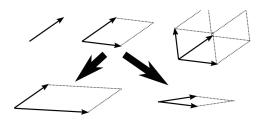
- L must be positive semidefinite.
- Only a statement of proportionality, eigenvalues of L need not < 1</p>

Geometry interpretation

$$X = \begin{bmatrix} x_1 & x_2 & \dots & x_n \end{bmatrix} \Longrightarrow$$

$$L(x_1, \dots, x_n) = X^\top X = \begin{pmatrix} \langle x_1, x_1 \rangle & \langle x_1, x_2 \rangle & \dots & \langle x_1, x_n \rangle \\ \langle x_2, x_1 \rangle & \langle x_2, x_2 \rangle & \dots & \langle x_2, x_n \rangle \\ \vdots & \vdots & \ddots & \vdots \\ \langle x_n, x_1 \rangle & \langle x_n, x_2 \rangle & \dots & \langle x_n, x_n \rangle \end{pmatrix}$$

- Gram determinant is the square of the volume of the parallelotope formed by the vectors
- vectors are linearly independent if and only if the Gram determinant is nonzero



Proof for the Geometry interpretation (1)

- ▶ in **1-element** case: $Vol^2(\mathbf{u}_1) = \mathbf{u}_1^{\top} \mathbf{u}_1$, i.e., length square of a line
- ▶ in **k-element** case: $Vol^2(\mathbf{u}_1 \dots \mathbf{u}_k, \mathbf{u}_{k+1}) = Vol^2(\mathbf{u}_1, \dots, \mathbf{u}_k) \|\tilde{\mathbf{u}}_{k+1}\|^2$
- Let $(\mathbf{u}_1, \dots, \mathbf{u}_k)$ is an $n \times k$ matrix \mathbf{Y} :
- ▶ Then there exists a vector $\mathbf{c} \in \mathbb{R}^k$ such that:

$$\begin{aligned} \mathbf{u}_{k+1} &= \mathbf{u}_{k+1} = \mathbf{YC} + \tilde{\mathbf{u}}_{k+1} \\ &= \underbrace{\begin{bmatrix} | & \vdots & | \\ \mathbf{u}_1 & \vdots & \mathbf{u}_k \\ | & \vdots & | \end{bmatrix}}_{\mathbf{C}_k} \underbrace{\begin{bmatrix} \mathbf{c}_1 \\ \vdots \\ \mathbf{c}_k \end{bmatrix}}_{\mathbf{c}_k} + \tilde{\mathbf{u}}_{k+1} \qquad \text{or } \mathbf{u}_{k+1} = \mathbf{c}_1 \mathbf{u}_1 + \mathbf{c}_2 \mathbf{u}_2 \dots \mathbf{c}_k \mathbf{u}_k + \tilde{\mathbf{u}}_{k+1} \end{aligned}$$

Proof for the Geometry interpretation (2)

extend Y to X:

$$\begin{split} \boldsymbol{X} &= [\boldsymbol{Y} \quad \boldsymbol{u}_{k+1}] = [\boldsymbol{u}_1 \quad \boldsymbol{u}_2 \quad \dots \quad \boldsymbol{u}_k \quad \boldsymbol{u}_{k+1}] = \begin{bmatrix} \boldsymbol{Y} \quad \boldsymbol{Y} \boldsymbol{c} + \tilde{\boldsymbol{u}}_{k+1} \end{bmatrix} \\ \Longrightarrow \boldsymbol{X}^\top \boldsymbol{X} &= \begin{bmatrix} \boldsymbol{Y}^\top \boldsymbol{Y} & \boldsymbol{Y}^\top (\boldsymbol{Y} \boldsymbol{c} + \tilde{\boldsymbol{u}}_{k+1}) \\ (\boldsymbol{Y} \boldsymbol{c} + \tilde{\boldsymbol{u}}_{k+1})^\top \boldsymbol{Y} & (\boldsymbol{Y} \boldsymbol{c} + \tilde{\boldsymbol{u}}_{k+1})^\top (\boldsymbol{Y} \boldsymbol{c} + \tilde{\boldsymbol{u}}_{k+1}) \end{bmatrix} \\ &= \begin{bmatrix} \boldsymbol{Y}^\top \boldsymbol{Y} & \boldsymbol{Y}^\top \boldsymbol{Y} \boldsymbol{c} \\ \boldsymbol{c}^\top \boldsymbol{Y}^\top \boldsymbol{Y} & \boldsymbol{c}^\top \boldsymbol{Y}^\top \boldsymbol{Y} \boldsymbol{c} + \tilde{\boldsymbol{u}}_{k+1}^\top \tilde{\boldsymbol{u}}_{k+1} \end{bmatrix} \quad \text{since } \boldsymbol{Y}^\top \tilde{\boldsymbol{u}}_{k+1} = \boldsymbol{0} \\ &= \begin{bmatrix} \boldsymbol{Y}^\top \boldsymbol{Y} & \boldsymbol{Y}^\top \boldsymbol{Y} \boldsymbol{c} \\ \boldsymbol{c}^\top \boldsymbol{Y}^\top \boldsymbol{Y} & \boldsymbol{c}^\top \boldsymbol{Y}^\top \boldsymbol{Y} \boldsymbol{c} + \|\tilde{\boldsymbol{u}}_{k+1}\|^2 \end{bmatrix} \\ &= \begin{bmatrix} \boldsymbol{Y}^\top \boldsymbol{Y} \\ \boldsymbol{c}^\top \boldsymbol{Y}^\top \boldsymbol{Y} \end{bmatrix} & \begin{pmatrix} \begin{bmatrix} \boldsymbol{Y}^\top \boldsymbol{Y} \boldsymbol{c} \\ \boldsymbol{c}^\top \boldsymbol{Y}^\top \boldsymbol{Y} \boldsymbol{c} \end{bmatrix} + \begin{bmatrix} \boldsymbol{0} \\ \|\tilde{\boldsymbol{u}}_{k+1}\|^2 \end{bmatrix} \end{pmatrix} \end{bmatrix} \end{split}$$

Multi-linearity states:

$$\begin{aligned} \det([a_1 + b_1, a_2, \dots, a_k]) &= \det\left([a_1, a_2, \dots, a_k]\right) + \det\left([b_1, a_2, \dots, a_k]\right) \\ &\Longrightarrow |\mathbf{X}^{\top} \mathbf{X}| = \begin{vmatrix} \mathbf{Y}^{\top} \mathbf{Y} & \mathbf{Y}^{\top} \mathbf{Y} \mathbf{c} \\ \mathbf{c}^{\top} \mathbf{Y}^{\top} \mathbf{Y} & \mathbf{c}^{\top} \mathbf{Y}^{\top} \mathbf{Y} \mathbf{c} \end{vmatrix} + \begin{vmatrix} \mathbf{Y}^{\top} \mathbf{Y} & \mathbf{0} \\ \mathbf{c}^{\top} \mathbf{Y}^{\top} \mathbf{Y} & \|\tilde{\mathbf{u}}_{k+1}\|^2 \end{vmatrix} \\ &= \mathbf{0} + \begin{vmatrix} \mathbf{Y}^{\top} \mathbf{Y} & \mathbf{0} \\ \mathbf{c}^{\top} \mathbf{Y}^{\top} \mathbf{Y} & \|\tilde{\mathbf{u}}_{k+1}\|^2 \end{vmatrix} \\ &= |\mathbf{Y}^{\top} \mathbf{Y}| \|\tilde{\mathbf{u}}_{k+1}\|^2 \end{aligned}$$

Normalization constant

Theorem says,

$$\sum_{A\subset Y\subset \mathcal{Y}} \det(L_Y) = \det(L+I_{\bar{A}})$$

For example,

$$L = \begin{pmatrix} 2.8599 & -0.4936 & -1.8458 \\ -0.4936 & 2.6264 & -1.1437 \\ -1.8458 & -1.1437 & 2.0522 \end{pmatrix}$$

$$A = \{1,2\} \implies \bar{A} = \{3\} \implies I_{\bar{A}} = \begin{pmatrix} 0 & 0 & 0 \\ 0 & 0 & 0 \\ 0 & 0 & 1 \end{pmatrix}$$

Therefore, normalisation constant (or partition function) is:

$$\sum_{\emptyset \subset Y \subset \mathcal{Y}} \det(L_Y) = \sum_{Y \subset \mathcal{Y}} \det(L_Y) = \det(L + I_{\bar{\emptyset}}) = \det(L + I)$$



Conversion to Marginal distribution (1)

$$\Pr_{L}(\mathbf{Y} = Y) \propto \det(L_{Y}) \implies \Pr_{L}(\mathbf{Y} = Y) = \frac{\det(L_{Y})}{\det(L + I)}$$

An L-ensemble is a DPP, and its marginal kernel is:

$$K = L(L+I)^{-1} = I - (L+I)^{-1}$$

 $L(L+I)^{-1} = I - (L+I)^{-1}$ is **true** for any *L* where $(L+I)^{-1}$ exist, think scalar case:

$$1 - \frac{1}{x+1} = \frac{x+1-1}{x+1} = \frac{x}{x+1}$$

Then,

$$\begin{aligned} \mathsf{Pr}_L(A \subseteq \mathbf{Y}) &= \frac{\sum_{A \subseteq Y \subseteq \mathcal{Y}} \mathsf{det}(L_Y)}{\sum_{Y \subseteq \mathcal{Y}} \mathsf{det}(L_Y)} \\ &= \frac{\mathsf{det}(L + I_{\bar{A}})}{\mathsf{det}(L + I)} \\ &= \mathsf{det}\left((L + I_{\bar{A}})(L + I)^{-1}\right) \end{aligned}$$

Since,
$$det(A^{-1}) = \frac{1}{det(A)}$$
 $det(AB) = det(A) det(B)$



Conversion to Marginal distribution (2)

$$\begin{aligned} \Pr_{L}(A \subseteq \mathbf{Y}) &= \det \left((L + I_{\bar{A}})(L + I)^{-1} \right) \\ &= \det \left(\underbrace{L(L + I)^{-1}}_{I - (L + I)^{-1}} + I_{\bar{A}}(L + I)^{-1} \right) \\ &= \det \left(I - (L + I)^{-1} + I_{\bar{A}}(L + I)^{-1} \right) \\ &= \det \left(I - (I - I_{\bar{A}})(L + I)^{-1} \right) \\ &= \det \left(I - I_{\bar{A}}(L + I)^{-1} \right) \\ &= \det \left(\underbrace{I_{\bar{A}} + I_{\bar{A}}}_{I} - I_{\bar{A}}(L + I)^{-1} \right) \\ &= \det \left(I_{\bar{A}} + I_{\bar{A}} - I_{\bar{A}}(L + I)^{-1} \right) \\ &= \det \left(I_{\bar{A}} + I_{\bar{A}} - I_{\bar{A}}(L + I)^{-1} \right) \\ &= \det \left(I_{\bar{A}} + I_{\bar{A}} - I_{\bar{A}}(L + I)^{-1} \right) \end{aligned}$$

Conversion to Marginal distribution (3)

$$\Pr_L(A \subseteq \mathbf{Y}) = \det\left(I_{\bar{A}} + I_A\left(\underbrace{I - (L + I)^{-1}}_{K}\right)\right)$$

left multiplication by I_A zeros out rows of a matrix except those corresponding to A, \Longrightarrow

$$K = \begin{pmatrix} K_{\bar{A}} & K_{\bar{A}A} \\ K_{A\bar{A}} & K_{A} \end{pmatrix} \implies I_A(K) = \begin{pmatrix} \mathbf{0} & \mathbf{0} \\ \mathbf{0} & I_{|A| \times |A|} \end{pmatrix} \begin{pmatrix} K_{\bar{A}} & K_{\bar{A}A} \\ K_{A\bar{A}} & K_{A} \end{pmatrix} = \begin{pmatrix} \mathbf{0} & \mathbf{0} \\ K_{A\bar{A}} & K_{A} \end{pmatrix}$$

Re-organise:

$$\begin{aligned} \mathsf{Pr}_L(A \subseteq \mathbf{Y}) &= \mathsf{det}(I_{\bar{A}} + I_A \mathcal{K}) \\ &= \begin{vmatrix} I_{|\bar{A}| \times |\bar{A}|} & \mathbf{0} \\ \mathcal{K}_{A\bar{A}} & \mathcal{K}_A \end{vmatrix} \\ &= \mathsf{det}(\mathcal{K}_A) \end{aligned}$$

 $K = L(L+I)^{-1} = I - (L+I)^{-1}$ is the conversion formula!



Eigen-value conversion

$$K = L(L+I)^{-1} = I - (L+I)^{-1}$$

Properties

$$\lambda_n \in \operatorname{eig}(A) \implies \lambda_n + 1 \in \operatorname{eig}(A + I)$$

 $\implies (\lambda_n)^{-1} \in \operatorname{eig}(A^{-1})$

Apply it to $K = I - (L + I)^{-1}$:

$$(\lambda_n + 1) \in eig(L + I) \implies \frac{1}{\lambda_n + 1} \in eig((L + I)^{-1})$$
$$\implies 1 - \frac{1}{\lambda_n + 1} \in eig(I - (L + I)^{-1})$$

$$1 - \frac{1}{\lambda_n + 1} = \frac{\lambda_n + 1 - 1}{\lambda_n + 1} = \frac{\lambda_n}{\lambda_n + 1}$$

Therefore,

$$L = \sum_{n=1}^{N} \lambda_n \mathbf{v}_n \mathbf{v}_n^{\top} \implies K = \sum_{n=1}^{N} \frac{\lambda_n}{\lambda_n + 1} \mathbf{v}_n \mathbf{v}_n^{\top}$$



Conversions from K to L

$$K = L(L+I)^{-1} = I - (L+I)^{-1}$$

$$K = I - (L+I)^{-1} \implies I - K = (L+I)^{-1}$$

$$\implies (L+I)(I-K) = I$$

$$\implies L+I-LK-K=I$$

$$\implies L(I-K) = K$$

$$\implies L = K(I-K)^{-1}$$

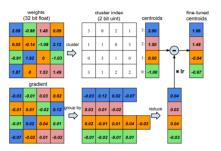
Many interesting things about DPP

- although it's a discrete distribution, but you can just sample it like softmax output, why?
- there is so much to study about this family of model and many interesting research, including neural networks, e.g., for mini-batch diversification
 - "Zhang, Kjellstrm, Mandt, Determinantal Point Processes for Mini-Batch Diversification"
- as well as a lot of applications such as diverse subset of news collection portal
- today, we talk about an example of it it used in neural network compression
 - "Mariet et., al, Diversity Networks: Neural Network Compression Using Determinantal Point Processes"

firstly, what is Neural Networks compression?

- very important topic, think about how we can put all VGG parameters in a mobile device!
- many works are started on neural network compression
- one seminal research:

"Han et. al, (2015) deep compression: Compression Deep Neural Networks with Pruning, Trained Quantization and Huffman coding"



- examples as such performs compression on parameters
- question is, can we use DPP to remove redundantly-performing neurons?



Apply DPP to Neural Networks compression

- let $\tau = \{I_1, \dots, I_J\}$ be the total number of J training data
- $ightharpoonup a_{ii}$ is activation of neuron *i* on image *j*, then
- $ightharpoonup \mathbf{v}_i$ contains a vector of all activation $[a_{i,1}, a_{i,1}, \dots, a_{i,J}]^{\top}$

A possible solution is to use DPP

- it is logical to retain a set of most "diversely performed" neurons
- \triangleright a particular neural layer has n_i neurons, each has some v_i responses to dataset,
- which we have a data matrix:

$$\mathbf{V} = \begin{bmatrix} | & \vdots & | \\ \mathbf{v}_1 & \vdots & \mathbf{v}_{n_l} \\ | & \vdots & | \end{bmatrix}$$

b to specify the *L* ensemble, we defined the following:

$$\mathbf{L}'_{ij} = \exp(-\beta \|\mathbf{v}_i - \mathbf{v}_j\|^2) \quad 1 \le i, j \le n_l$$

then we can sample the most diverse subset of neurons using DPP



Module Three

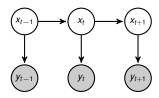
LSTM to help Kalman Filter

Kalman Filter

- why we need filtering?
- for example
- ▶ the objective is to estimate the probability of the current state x_t , given all the measurements so far: $y_1, y_2, ..., y_t$
- insert some pictures of application of Kalman Filter

State Space Models

graphical model for state space model:



using markov property of probabilistic graphical model:

$$p(x_t|x_1,...,x_{t-1},y_1,...,y_{t-1}) = p(x_t|x_{t-1})$$

$$p(y_t|x_1,...,x_{t-1},x_t,y_1,...,y_{t-1}) = p(y_t|x_t)$$

▶ looks familiar to Recurrent Neural Networks?



What do we want to compute?

$$\begin{aligned} \text{Prediction}: & & & & & & & & & & & & & & & & \\ & & & & & & & & & & & & & \\ & & & & & & & & & & \\ & & & & & & & & & & \\ & & & & & & & & \\ & & & & & & & & \\ & & & & & & & \\ & & & & & & & \\ & & & & & & & \\ & & & & & & \\ & & & & & & \\ & & & & & & \\ & & & & & & \\ & & & & & & \\ & & & & & \\ & & & & & \\ & & & & & \\ & & & & & \\ & & & & & \\ & & & & \\ & & & & \\ & & & & \\ & & & & \\ & & & & \\ & & & & \\ & & & & \\ & & \\ & & & \\ & & \\ & & & \\ & & \\ & & & \\ & &$$

This is because:

$$p(x_t|\mathbf{y}_{1:t}) \propto p(x_t, \mathbf{y}_{1:t})$$

$$\propto p(y_t|x_t)p(x_t|\mathbf{y}_{1:t-1})$$

$$= \frac{p(y_t|x_t)p(x_t|\mathbf{y}_{1:t-1})}{\int_{s_t} (y_t|s_t)p(ds_t|\mathbf{y}_{1:t-1})}$$

Linear Gaussian Dynamic Model

$$\begin{aligned} \mathbf{x}_t &= A\mathbf{x}_{t-1} + B + w_t & w_t \sim \mathcal{N}(\mathbf{0}, Q_t) \\ \Longrightarrow & \text{Transition probability:} & \rho(\mathbf{x}_t | \mathbf{x}_{t-1}) \sim \mathcal{N}(A\mathbf{x}_{t-1} + B, Q_t) \\ \\ y_t &= H\mathbf{x}_t + v_t & v_t \sim \mathcal{N}(\mathbf{0}, R_t) \\ \Longrightarrow & \text{Measurement probability:} & \rho(y_t | \mathbf{x}_t) \sim \mathcal{N}(H\mathbf{x}_t, R_t) \end{aligned}$$

- ▶ Kalman Filter can be used to in this Gaussian, Linear case.
- In general, there are many other Dyanmic models which are non-Gaussian, non-Linear. They can NOT be solved using Kalman Filter.

Kalman Filter - Prediction

- following marginal distribution of linear Gaussian model, given:
- \triangleright $p(X_1) \sim \mathcal{N}(x_1|\mu, \Sigma)$
- ▶ $p(X_2|x_1) \sim \mathcal{N}(x_2|Ax_1 + B, Q)$
- then, its marginal is:

$$p(X_2) = \int_{x_1} p(X_2|x_1)p(x_1) \sim \mathcal{N}\left(x_2|A\mu + B, A\Sigma A^T + Q\right) dx_1$$

$$\begin{aligned} \textbf{Prediction}: \quad & p(x_{t}|\textbf{y}_{1:t-1}) \sim \mathcal{N}(\bar{\mu}_{t}, \bar{\Sigma}_{t}) = \int_{x_{t-1}} p(x_{t}|x_{t-1}) p(dx_{t-1}|\textbf{y}_{1:t-1}) \\ & = \int_{x_{t-1}} \mathcal{N}(x_{t}|Ax_{t-1} + B, Q_{t}) \mathcal{N}(x_{t-1}|\hat{\mu}_{t-1}, \hat{\Sigma}_{t-1}) \\ & = \mathcal{N}\left(x_{t}|A\hat{\mu}_{t-1} + B, A\hat{\Sigma}_{t-1}A^{T} + Q_{t}\right) \end{aligned}$$

Question How about **update**? Let's be "cool", and try an alternative method using Moment representation.



Kalman Filter Update: $p(x_t|y_1,...y_t) = \mathcal{N}(\hat{\mu}_t,\hat{\Sigma}_t)$

$$p(x_t|y_1,\ldots y_t) \sim \mathcal{N}(\hat{\mu}_t,\hat{\Sigma}_t)$$

$$\propto p(y_t|x_t)p(x_t|\mathbf{y}_{1:t-1}) = \mathcal{N}(y_t|Hx_t,R_t)\mathcal{N}(x_t|\bar{\mu}_t,\bar{\Sigma}_t)$$

- given marginal $p(\mathbf{x}_t) = \mathcal{N}(\mu_x, \Sigma_{xx})$ $p(\mathbf{y}_t) = \mathcal{N}(\mu_y, \Sigma_{yy})$
- ▶ and **joint** density, $\mathbf{x}_t \equiv x_t | y_1, \dots, y_{t-1}$ and $\mathbf{y}_t \equiv y_t | y_1, \dots, y_{t-1}$

$$\begin{bmatrix} \mathbf{x}_t \\ \mathbf{y}_t \end{bmatrix} \sim \mathcal{N}\left(\begin{bmatrix} \mu_x \\ \mu_y \end{bmatrix}, \begin{bmatrix} \Sigma_{xx}, \Sigma_{xy} \\ \Sigma_{xy}, \Sigma_{yy} \end{bmatrix} \right)$$

then, conditional density is:

$$p(\mathbf{x}_t|\mathbf{y}_t) \sim \mathcal{N}\left(\mu_x + \Sigma_{xy}\Sigma_{yy}^{-1}(y - \mu_y), \Sigma_{xx} - \Sigma_{xy}\Sigma_{yy}^{-1}\Sigma_{yx}\right)$$

In filtering:

$$p(\mathbf{x}_t) \equiv p(x_t|y_1, \dots, y_{t-1}) = \mathcal{N}(x_t|\bar{\mu}_t, \bar{\Sigma}_t)$$

$$p(\mathbf{y}_t) \equiv p(y_t|y_1, \dots, y_{t-1})$$

We are after conditional $p(\mathbf{x}_t|\mathbf{y}_t, y_1, \dots, y_{t-1})$



Kalman Filter Update

in order to compute:

$$p(\mathbf{x}_t|\mathbf{y}_t) \sim \mathcal{N}\left(\mu_{x} + \Sigma_{xy}\Sigma_{yy}^{-1}(y - \mu_{y}), \Sigma_{xx} - \Sigma_{xy}\Sigma_{yy}^{-1}\Sigma_{yx}\right)$$

- we need to know: $(\mu_x, \mu_y, \Sigma_{xx}, \Sigma_{yy}, \Sigma_{xy}, \Sigma_{yx})$
- ▶ introduce a zero-mean variable: $\triangle x_{t-1}$:

$$\triangle x_{t-1} \equiv x_{t-1} - \mathbb{E}[x_{t-1}] \sim \mathcal{N}(0, \hat{\Sigma}_{t-1}) \implies x_{t-1} = \triangle x_{t-1} + \mathbb{E}[x_{t-1}]$$

▶ attempt to write both $\triangle x_t$ and $\triangle y_t$ in terms of $\triangle x_{t-1}$:

$$\mathbf{x}_{t} = A\mathbf{x}_{t-1} + w_{t} \qquad w_{t} \sim \mathcal{N}(0, Q_{t}) \implies \qquad x_{t} = A(\triangle x_{t-1} + \mathbb{E}[x_{t-1}]) + w_{t}$$

$$= A\mathbb{E}x_{t-1} + \underbrace{A\triangle x_{t-1} + w_{t}}_{\triangle x_{t}}$$

$$y_{t} = H\mathbf{x}_{t} + v_{t} \qquad v_{t} \sim \mathcal{N}(0, R_{t}) \implies \qquad y_{t} = Hx_{t} + v_{t}$$

$$= H(A\mathbb{E}x_{t-1} + A\triangle x_{t-1} + w_{t}) + v_{t}$$

$$= HA\mathbb{E}x_{t-1} + \underbrace{HA\triangle x_{t-1} + Hw_{t} + v_{t}}_{\triangle y_{t}}$$

The Independence assumptions:

$$\mathbb{COV}(x_{t-1}, w_t) = 0 \qquad \mathbb{COV}(x_{t-1}, v_t) = 0 \qquad \mathbb{COV}(w_t, v_t) = 0$$

Moment Representation

• we need to know: $(\mu_x, \mu_y, \Sigma_{xx}, \Sigma_{yy}, \Sigma_{xy}, \Sigma_{yx})$

$$\mu_{X} \equiv \qquad \mathbb{E}[x_{t}|y_{1:t-1}] = A\mathbb{E}[x_{t-1}] = A\hat{\mu}$$

$$\mu_{Y} \equiv \qquad \mathbb{E}[y_{t}|y_{1:t-1}] = HA\mathbb{E}[x_{t-1}] = HA\hat{\mu}$$

$$\Sigma_{XX} \equiv \qquad \mathbb{E}[\triangle x_{t}(\triangle x_{t})^{T}|y_{1:t-1}] = \mathbb{E}[(A\triangle x_{t-1} + w_{t})(A\triangle x_{t-1} + w_{t})^{T}]$$

$$= A\hat{\Sigma}_{t-1}A^{T} + Q_{t} = \bar{\Sigma}_{t}$$

$$\Sigma_{YX} \equiv \qquad \mathbb{E}[\triangle y_{t}(\triangle x_{t})^{T}|y_{1:t-1}] = \mathbb{E}\left[(HA\triangle x_{t-1} + Hw_{t} + v_{t})(A\triangle x_{t-1} + w_{t})^{T}\right]$$

$$= H\left(A\hat{\Sigma}_{t-1}A^{T} + Q_{t}\right) = H\bar{\Sigma}_{t}$$

$$\Longrightarrow \Sigma_{YX} \equiv \qquad \mathbb{E}[\triangle x_{t}(\triangle y_{t})^{T}|y_{1:t-1}]$$

$$= \left(A\hat{\Sigma}_{t-1}A^{T} + Q_{t}\right)^{T}H^{T} = \bar{\Sigma}_{t}H^{T}$$

$$\Sigma_{YY} \equiv \qquad \mathbb{E}[\triangle y_{t}(\triangle y_{t})^{T}|y_{1:t-1}] = \mathbb{E}\left[(HA\triangle x_{t-1} + Hw_{t} + v_{t})(HA\triangle x_{t-1} + Hw_{t} + v_{t})^{T}\right]$$

$$= H\left(A\hat{\Sigma}_{t-1}A^{T} + Q_{t}\right)H^{T} + R_{t}$$

$$= H(\bar{\Sigma}_{t})H^{T} + R_{t}$$

Kalman Filter Update: $p(x_t|y_1,...y_t) = \mathcal{N}(\hat{\mu}_t,\hat{\Sigma}_t)$

- finally we put all elements in:

$$\mu_{x} + \Sigma_{xy} \Sigma_{yy}^{-1} (y - \mu_{y})$$

$$\equiv \mathbb{E}[\mathbf{x}_{t}] + \mathbb{E}[\Delta \mathbf{x}_{t} (\Delta \mathbf{y}_{t})^{T}] \mathbb{E}[\Delta y_{t} (\Delta \mathbf{y}_{t})^{T}]^{-1} (\mathbf{y}_{t} - \mathbb{E}[y_{t}])$$

$$= A\hat{\mu}_{t-1} + \underbrace{\bar{\Sigma}_{t}^{T} H (H\bar{\Sigma}_{t} H^{T} + R_{t})^{-1}}_{K} (\mathbf{y}_{t} - HA\hat{\mu}_{t-1})$$

$$= \bar{\mu}_{t-1} + K(\mathbf{y}_{t} - HA\hat{\mu}_{t-1})$$

co-variance:
$$\hat{\Sigma}_t = \mathbb{COV}[x_t|y_{1:t}]$$
:

$$\begin{split} & \Sigma_{xx} - \Sigma_{xy} \Sigma_{yy}^{-1} \Sigma_{yx} \\ & \equiv \mathbb{E}[\triangle x_t (\triangle x_t)^T] - \mathbb{E}[\triangle x_t (\triangle y_t)^T] \mathbb{E}[\triangle y_t (\triangle y_t)^T]^{-1} \mathbb{E}[\triangle y_t (\triangle x_t)^T] \\ & = \bar{\Sigma}_t - \underbrace{\bar{\Sigma}_t H^T (H(\bar{\Sigma}_t) H^T + R_t)^{-1}}_{K} H \bar{\Sigma}_t \\ & = (I - KH) \bar{\Sigma}_t \end{split}$$

Kalman Filter and Neural Networks

- Kalman filters require a motion model and measurement model to be specified at priory
- it's hard!
- can be crude approximation of reality
- this is where LSTM can help out!
- "Coskun., Long Short-Term Memory Kalman Filters: Recurrent Neural Estimators for Pose Regularization"

- you see it everywhere, so I don't talk about it in detail:
- a compact form of representation:

$$\begin{bmatrix} i \\ f \\ o \\ \tilde{C} \end{bmatrix} = \begin{bmatrix} \sigma \\ \sigma \\ \sigma \\ \tanh \end{bmatrix} \left(\mathbf{W} \begin{bmatrix} h_{t-1} \\ x_t \end{bmatrix} + \mathbf{b} \right) \qquad C_t = f_t \odot C_{t-1} + i \odot \tilde{C}_t$$

$$h_t = o_t \odot \tanh(C_t)$$

ightharpoonup in LSTMs, cell state C_t . The derivative of consecutive states is of the form:

$$C_{t} = f_{t}(C_{t-1}) \odot C_{t-1} + i_{t}(C_{t-1}) \odot \tilde{C}_{t}(C_{t-1})$$

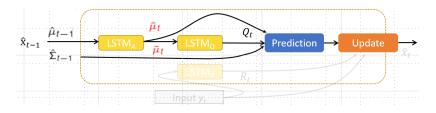
$$= f_{t} \odot C_{t-1} + i_{t} \odot \tanh(W_{C}[h_{t-1}, x_{t}] + b_{C})$$

$$= f_{t}(C_{t-1})C_{t-1} + i_{t}(h_{t-1}(C_{t-1}))\tanh(W_{C}[o_{t-1}(h_{t-1}(C_{t-1})) \odot \tanh(C_{t-1}), x_{t}] + b_{C})$$

$$\frac{\partial C_{t}}{\partial C_{t-1}} = \int_{\text{gradient highway}} + \frac{\partial f_{t}}{\partial C_{t-1}}C_{t-1} + \frac{\partial \xi(C_{t-1})}{\partial C_{t-1}}C_{t-1}$$
contains exponentially fast decay function

- \triangleright of course, f_t may still close to zero
- ▶ trick is to initialize bias to be a large positive number, e.g., $f_t = \sigma(W_t[h_{t-1}, x_t] + \text{large positive number})$ so to make f_t closer to 1 initially

LSTM Kalman Filters: Prediction



the state-space model is changed into:

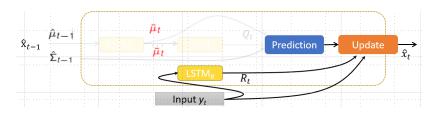
$$\begin{split} \mathbf{x}_t &= A\mathbf{x}_{t-1} + w_t & \quad w_t \sim \mathcal{N}(0, Q_t) & \rightarrow & \quad \mathbf{x}_t &= \operatorname{lstm}_A(\mathbf{x}_{t-1}) + w_t & \quad w_t \sim \mathcal{N}(0, Q_t) \\ \mathbf{y}_t &= H\mathbf{x}_t + v_t & \quad v_t \sim \mathcal{N}(0, R_t) & \rightarrow & \quad \operatorname{unchanged} \end{split}$$

prediction

$$\begin{split} \bar{\mu}_t &= A\hat{\mu}_{t-1} \quad \rightarrow \quad \bar{\underline{\mu}}_t = \text{Istm}_A(\hat{\mu}_{t-1}) \\ \bar{\Sigma}_t &= A\hat{\Sigma}_{t-1}A^T + Q_t \quad \rightarrow \quad \bar{\Sigma}_t = \mathcal{A}'\hat{\Sigma}_{t-1}\mathcal{A'}^\top + \text{Istm}_Q(\bar{\underline{\mu}}_t), \text{ where } \mathcal{A'} = \nabla_{\mathbf{x}_{t-1}} \text{Istm}_A(\hat{\mu}_{t-1}) \end{split}$$



LSTM Kalman Filters: Update



the state-space model is changed into:

$$\begin{split} \mathbf{x}_t &= A \mathbf{x}_{t-1} + w_t & \quad w_t \sim \mathcal{N}(0, Q_t) & \rightarrow & \quad \mathbf{x}_t &= \mathrm{lstm}_A(\mathbf{x}_{t-1}) + w_t & \quad w_t \sim \mathcal{N}(0, Q_t) \\ \mathbf{y}_t &= H \mathbf{x}_t + v_t & \quad v_t \sim \mathcal{N}(0, R_t) & \rightarrow & \quad \text{unchanged} \end{split}$$

update:

$$\begin{split} & K_t = \bar{\Sigma}_t H^T (H(\bar{\Sigma}_t) H^T + R_t)^{-1} & \to & K_t = \bar{\Sigma}_t H^T (H(\bar{\Sigma}_t) H^T + \text{lstm}_R(\mathbf{y}_t))^{-1} \\ & \hat{\mu}_t = \bar{\mu}_t + K(\mathbf{y}_t - HA\hat{\mu}_{t-1}) & \to & \text{unchanged} \\ & \hat{\Sigma}_t = (I - KH)\bar{\Sigma}_t & \to & \text{unchanged} \end{split}$$

Module Four

Binary Classifier to help probability model estimation

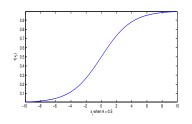
probability and classification

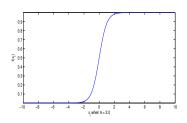
firstly, probability models and classification are closely related:

$$\operatorname*{arg\;max}_{\theta} \left(p_{\theta}(\mathbf{Y}) \right) \implies \operatorname*{arg\;min}_{\theta} \left(- \log p_{\theta}(\mathbf{Y}) \right)$$

in following example, let's show classification models incorporating our favorite sigmoid function:

$$\sigma(\mathbf{x}_i^{\top}\theta) = \frac{1}{1 + \exp(-\mathbf{x}_i^{T}\theta)}$$





Example: Bernoulli & Logistic regression

Bernoulli distribution using Sigmoid function

$$p_{\boldsymbol{\theta}}(\mathbf{Y}|\mathbf{X}) = \prod_{i=1}^{n} \left[\frac{1}{1 + \exp(-\mathbf{x}_{i}^{T}\boldsymbol{\theta})} \right]^{y_{i}} \left[1 - \frac{1}{1 + \exp(-\mathbf{x}_{i}^{T}\boldsymbol{\theta})} \right]^{1-y_{i}}$$

Logistic regression

$$\begin{aligned} \mathcal{C}(\boldsymbol{\theta}) &= -\log[p_{\boldsymbol{\theta}}(\mathbf{Y}|\mathbf{X})] \\ &= -\left(\sum_{i=1}^{n} y_{i} \log\left[\frac{1}{1 + \exp(-\mathbf{x}_{i}^{T}\boldsymbol{\theta})}\right] + (1 - y_{i}) \log\left[1 - \frac{1}{1 + \exp(-\mathbf{x}_{i}^{T}\boldsymbol{\theta})}\right]\right) \end{aligned}$$

Example: Multinomial Distribution & Cross Entropy Loss

Multinomial Distribution with softmax

$$p_{\theta}(\mathbf{Y}|\mathbf{X}) = \prod_{i=1}^{n} \prod_{k=1}^{K} \left[\left(\frac{\exp(\mathbf{X}_{i}^{T} \boldsymbol{\theta}_{k})}{\sum_{l=1}^{K} \exp(\mathbf{X}_{i}^{T} \boldsymbol{\theta}_{l})} \right) \right]^{y_{i,k}}$$

cross entropy loss with Softmax

$$C(\boldsymbol{\theta}) = -\log[p_{\boldsymbol{\theta}}(\mathbf{Y}|\mathbf{X})] = -\sum_{i=1}^{N} \sum_{k=1}^{K} y_{i,k} \left[\log\left(\frac{\exp(\mathbf{x}_{i}^{T}\boldsymbol{\theta}_{k})}{\sum_{l=1}^{K} \exp(\mathbf{x}_{i}^{T}\boldsymbol{\theta}_{l})}\right) \right]$$

Example: Gaussian Distribution & Sum of Square Loss

- ▶ this time, let's go from $C(\theta) \rightarrow p_{\theta}(\mathbf{Y})$
- Sum of Square Loss

$$C(\boldsymbol{\theta}) = \sum_{k=1}^{K} (\hat{y}_k(\boldsymbol{\theta}) - y_k)^2$$

Gaussian distribution

$$p_{\boldsymbol{\theta}}(\mathbf{Y}|\mathbf{X}) \propto \exp\left[-\mathcal{C}(\boldsymbol{\theta})\right] = \exp\left[-\sum_{k=1}^{K} \left(\hat{y}_k(\boldsymbol{\theta}) - y_k\right)^2\right]$$

question: what if we use *Square* loss instead of *Cross Entropy* loss in Softmax, where:

$$\hat{y}_k(\theta) = \frac{\exp(\mathbf{x}_i^T \theta_k)}{\sum_{l=1}^K \exp(\mathbf{x}_i^T \theta_l)}$$



Think about Classification's best friend, "Softmax" again!

- for example, in word embedding, we want to align a target word u_w with center word v_c:
- ightharpoonup for simplicity, for the rest of the article, we let $\mathbf{w} \equiv \mathbf{u}_w$ and $\mathbf{c} \equiv \mathbf{v}_c$

$$\Pr_{\theta}(\mathbf{w}|\mathbf{c}) = \frac{u_{\theta}(\mathbf{w}|\mathbf{c})}{\sum_{\mathbf{w}' \in \mathcal{V}} u_{\theta}(\mathbf{w}'|\mathbf{c})} = \frac{u_{\theta}(\mathbf{w}|\mathbf{c})}{Z_{c}} \equiv \frac{\exp(\mathbf{w}^{\top}\mathbf{c})}{\sum_{\mathbf{w}' \in \mathcal{V}} \exp(\mathbf{w}'^{\top}\mathbf{c})}$$

ightharpoonup the denominator, i.e., the $\sum_{\mathbf{w}' \in \mathcal{V}} u(\mathbf{w}' | \mathbf{c})$ can be too computational

Turn the problem around!

- ▶ data distribution: we sample $\mathbf{w} \sim \bar{p}(\mathbf{w}|\mathbf{c})$ from its empirical (data) distribution, and give a label $\mathcal{Y} = 1$
- Noise distribution: we can sample k w̄ ~ q(w), and give them labels y = 0 importantly, condition for q(.) is: it does not assign zero probability to any data.
- Can we build a binary classifier to classify its label, i.e., which distribution has generated it?



Noise Contrastive Estimation (NCE)

- training data generation: (w, c, y)
 - 1. sample (\mathbf{w}, \mathbf{c}) : using $\mathbf{c} \sim \tilde{p}(\mathbf{c}), \mathbf{w} \sim \tilde{p}(\mathbf{w}|\mathbf{c})$ and label them as $\mathcal{Y} = 1$
 - 2. k "noise" samples from q(.), and label them as $\mathcal{Y}=0$
- can we instead, try to maximize the joint posterior Bernoulli distribution:

$$\Pr_{\theta}(\mathcal{Y}|\mathbf{W},\mathbf{c}) = \prod_{i=1}^{k+1} \left(\Pr(\mathcal{Y}_i|\mathbf{w}_i,\mathbf{c})\right)^{y_i} \left(1 - \Pr(\mathcal{Y}_i|\mathbf{w}_i,\mathbf{c})\right)^{1-y_i}$$

or minimize the corresponding Logistic regression:

$$\begin{split} \mathcal{C} &= -\log[\Pr_{\theta}(\mathcal{Y}|\mathbf{W}, \mathbf{c})] \\ &= -\sum_{i=1}^{k+1} y_i \log\left[\Pr_{\theta}(\mathcal{Y}_i|\mathbf{w}_i, \mathbf{c})\right] + (1 - y_i) \log\left[1 - \Pr_{\theta}(\mathcal{Y}_i|\mathbf{w}_i, \mathbf{c})\right] \end{split}$$

Noise Contrastive Estimation (NCE)

we assume there are *k* negative samples per positive sample, so the prior density is:

$$P(\mathcal{Y} = y) = \begin{cases} \frac{1}{k+1} & y = 1\\ \frac{k}{k+1} & y = 0 \end{cases}$$

▶ then the posterior of $P(\mathcal{Y}|\mathbf{c},\mathbf{w})$:

$$\begin{split} P(\mathcal{Y} = 1 | \mathbf{c}, \mathbf{w}) &= \frac{\Pr(\mathcal{Y} = 1, \mathbf{w} | \mathbf{c})}{\Pr(\mathbf{w} | \mathbf{c})} = \frac{\Pr(\mathbf{w} | \mathcal{Y} = 1, \mathbf{c}) P(\mathcal{Y} = 1)}{\sum_{y \in \{0,1\}} P(\mathbf{w} | \mathcal{Y} = y, \mathbf{c}) P(\mathcal{Y} = y)} \\ &= \frac{\tilde{P}(\mathbf{w}) \times \frac{1}{1+k}}{\tilde{P}(\mathbf{w} | \mathbf{c}) \times \frac{1}{k+1} + q(\mathbf{w}) \times \frac{k}{1+k}} \\ &= \frac{\tilde{P}(\mathbf{w} | \mathbf{c})}{\tilde{P}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} \\ \Pr(\mathcal{Y} = 0 | \mathbf{c}, \mathbf{w}) &= 1 - \Pr(\mathcal{Y} = 1 | \mathbf{c}, \mathbf{w}) \\ &= 1 - \frac{\tilde{P}(\mathbf{w} | \mathbf{c})}{\tilde{P}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} \\ &= \frac{kq(\mathbf{w})}{\tilde{P}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} \end{split}$$

Apply NCE to NLP problem

in summary:

$$\Pr(\mathcal{Y} = y | \mathbf{c}, \mathbf{w}) = \begin{cases} \frac{\frac{\bar{P}(\mathbf{w} | \mathbf{c})}{\bar{P}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})}}{\frac{\bar{P}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})}{\bar{P}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})}} & y = 1 \end{cases}$$

it can be replaced by un-normalized function:

$$\Pr(\mathcal{Y} = y | \mathbf{c}, \mathbf{w}) = \begin{cases} \frac{u_{\theta}(\mathbf{w} | \mathbf{c})}{u_{\theta}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} & y = 1\\ \frac{kq(\mathbf{w})}{u_{\theta}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} & y = 0 \end{cases}$$

- formal proof can be found "Gutmann, 2012, Noise-Contrastive Estimation of Unnormalized Statistical Models, with Applications to Natural Image Statistics"
- let's see an intuition through softmax

Intuition through Softmax

think about Softmax in word embedding:

$$\Pr_{\theta}(\mathbf{w}|\mathbf{c}) = \frac{u_{\theta}(\mathbf{w}|\mathbf{c})}{\sum_{\mathbf{w}' \in \mathcal{V}} u_{\theta}(\mathbf{w}'|\mathbf{c})} = \frac{u_{\theta}(\mathbf{w}|\mathbf{c})}{Z_{c}} \equiv \frac{\exp(\mathbf{w}^{\top}\mathbf{c})}{\sum_{\mathbf{w}' \in \mathcal{V}} \exp(\mathbf{w}^{\top}\mathbf{c})}$$

- ▶ say $\{\mathbf{w}_1, \mathbf{w}_2, \dots, \mathbf{w}_k\}$ are target words having high frequencies given **c**
- $ightharpoonup \{r_1, r_2, \dots r_n\}$ are words having low frequency given **c**
- ▶ say we pick $\mathbf{w}_i \in \{\mathbf{w}_1, \dots \mathbf{w}_k\}$ to optimize: at each round, we aim to increase $\mathbf{w}_i^{\top}\mathbf{c}$; at the same time, sum of rest of softmax weights: $\left\{\{\mathbf{w}_j^{\top}\mathbf{c}\}_{j\neq i} \cup \{\mathbf{r}_j^{\top}\mathbf{c}\}\right\}$ decrease
- in softmax, such decrease is guaranteed by the sum in denominator
- ightharpoonup each \mathbf{w}_i has a chance to increase $\mathbf{w}_i^{\top} \mathbf{c}$, but each $\mathbf{r}_i^{\top} \mathbf{c}$ will (hopefully) stay low
- ▶ **intuition**: in NCE, instead of using sum in the denominator, we "designed" a probability q(.), such that, while letting \mathbf{w}_i be a positive training sample, we also have chance to let $\mathbf{w}_{j\neq i}$ to be part of negative training sample, i.e., to reduce the value of $\mathbf{w}_j^{\top}\mathbf{c}$; it somewhat has a similar effect as **softmax**



NCE in a nutshell

NCE transforms:

- a problem of model estimation (computationally expensive) to:
- a problem of estimating parameters of probabilistic binary posterior classifier (computationally acceptable):
- main advantage: it allows us to fit models that are not explicitly normalized, making training time effectively independent of the vocabulary size

NCE objective function

let $u_{\theta}(\mathbf{w}|\mathbf{c}) = \exp[s_{\theta}(\mathbf{w}|\mathbf{c})]$:

$$\begin{split} \Pr(\mathcal{Y} = 1 | \mathbf{c}, \mathbf{w}) &= \frac{u_{\theta}(\mathbf{w} | \mathbf{c})}{u_{\theta}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} = \sigma \big(\triangle s_{\theta}(\mathbf{w} | \mathbf{c}) \big) \\ \Pr(\mathcal{Y} = 0 | \mathbf{c}, \mathbf{w}) &= \frac{kq(\mathbf{w})}{u_{\theta}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} = 1 - \sigma \big(\triangle s_{\theta}(\mathbf{w} | \mathbf{c}) \big) \\ &\qquad \qquad \text{where } \triangle s_{\theta}(\mathbf{w} | \mathbf{c}) \equiv s_{\theta}(\mathbf{w} | \mathbf{c}) - \log(kq(\mathbf{w})) \qquad \text{let's see why} \end{split}$$

$$\begin{split} \sigma\big(\triangle s_{\theta}(\mathbf{w}|\mathbf{c})\big) &= \frac{1}{1 + \exp\big[-s_{\theta}(\mathbf{w}|\mathbf{c}) + \log(kq(\mathbf{w}))\big]} \\ &= \frac{1}{1 + \exp\big(-s_{\theta}(\mathbf{w}|\mathbf{c})\big) \times kq(\mathbf{w})} \\ &= \frac{\exp\big[s_{\theta}(\mathbf{w}|\mathbf{c})\big]}{\exp\big[s_{\theta}(\mathbf{w}|\mathbf{c})\big] + kq(\mathbf{w})} = \frac{u_{\theta}(\mathbf{w}|\mathbf{c})}{u_{\theta}(\mathbf{w}|\mathbf{c}) + kq(\mathbf{w})} \end{split}$$

therefore the objective function is:

$$\boldsymbol{\theta}^* = \arg\max_{\boldsymbol{\theta}} \sum_{(\boldsymbol{w}, \boldsymbol{c}) \in \mathcal{D}} \sigma(\triangle \boldsymbol{s}_{\boldsymbol{\theta}}(\boldsymbol{w}|\boldsymbol{c})) + \sum_{(\bar{\boldsymbol{w}}, \boldsymbol{c}) \in \tilde{\mathcal{D}}} \sigma(-\triangle \boldsymbol{s}_{\boldsymbol{\theta}}(\bar{\boldsymbol{w}}|\boldsymbol{c}))$$



NCE and Negative Sampling

- negative sampling is a special case of NCE
- we let $k = |\mathcal{V}|$ and q(.) is uniform:

$$\begin{split} P(\mathcal{Y} = 1|\mathbf{c}, \mathbf{w}) &= \frac{u_{\theta}(\mathbf{w}|\mathbf{c})}{u_{\theta}(\mathbf{w}|\mathbf{c}) + |\mathcal{V}| \frac{1}{|\mathcal{V}|}} = \frac{u_{\theta}(\mathbf{w}|\mathbf{c})}{u_{\theta}(\mathbf{w}|\mathbf{c}) + 1} \\ P(\mathcal{Y} = 0|\mathbf{c}, \mathbf{w}) &= \frac{|\mathcal{V}| \frac{1}{|\mathcal{V}|}}{u_{\theta}(\mathbf{w}|\mathbf{c}) + |\mathcal{V}| \frac{1}{|\mathcal{V}|}} = \frac{1}{u_{\theta}(\mathbf{w}|\mathbf{c}) + 1} \end{split}$$

correspondingly, we have:

$$\triangle s_{\theta}(\mathbf{w}|\mathbf{c}) \equiv s_{\theta}(\mathbf{w}|\mathbf{c}) - \log\left(|\mathcal{V}|\frac{1}{|\mathcal{V}|}\right) = s_{\theta}(\mathbf{w}|\mathbf{c}) = \mathbf{w}^{\top}\mathbf{c}$$

in Skip-gram:

$$\begin{split} \boldsymbol{\theta}^* &= \arg\max_{\boldsymbol{\theta}} \sum_{(\mathbf{w}, \mathbf{c}) \in D} \sigma(\mathbf{w}^\top \mathbf{c}) + \sum_{(\tilde{\mathbf{w}}, c) \in \tilde{D}} \sigma(-\bar{\mathbf{w}}^\top \mathbf{c}) \\ &= \arg\min_{\boldsymbol{\theta}} \sum_{(\mathbf{w}, c) \in D} \sigma(-\mathbf{u}_{\mathbf{w}}^\top \mathbf{v}_c) + \sum_{(\tilde{\mathbf{w}}, c) \in \tilde{D}} \frac{1}{1 + \exp\left(-\bar{\mathbf{w}}^\top \mathbf{c}\right)} \end{split}$$

why un-normalised $u_{\theta}(\mathbf{w}, \mathbf{c})$ still works?

▶ talk a look at this again, let $u_{\theta}(\mathbf{w}|\mathbf{c}) = \exp[s_{\theta}(\mathbf{w}|\mathbf{c})]$:

$$\Pr(\mathcal{Y} = 1 | \mathbf{c}, \mathbf{w}) = \frac{u_{\theta}(\mathbf{w} | \mathbf{c})}{u_{\theta}(\mathbf{w} | \mathbf{c}) + kq(\mathbf{w})} = \sigma(\triangle s_{\theta}(\mathbf{w} | \mathbf{c}))$$

$$\text{where } \triangle s_{\theta}(\mathbf{w} | \mathbf{c}) \equiv s_{\theta}(\mathbf{w} | \mathbf{c}) - \log(kq(\mathbf{w}))$$

we already know:

$$= \sigma(\triangle s_{\theta}(\mathbf{w}|\mathbf{c})) = \frac{1}{1 + \underbrace{\exp(-s_{\theta}(\mathbf{w}|\mathbf{c})) \times kq(\mathbf{w})}_{G(\mathbf{w},\theta)}}$$

in this case,

$$\begin{split} G(\mathbf{w}, \theta) &= \exp \left(- s_{\theta}(\mathbf{w}|\mathbf{c}) \right) \times kq(\mathbf{w}) \\ &= \frac{kq(\mathbf{w})}{\exp(s_{\theta}(\mathbf{w}|\mathbf{c}))} = \frac{kq(\mathbf{w})}{u_{\theta}(\mathbf{w}|\mathbf{c})} \end{split}$$

or more generically:

$$G(\mathbf{w}, \theta) = \frac{m}{n} \frac{q(\mathbf{w})}{u_{\theta}(\mathbf{w}|\mathbf{c})}$$



what do we need to prove?

- look at $G(\mathbf{w}, \theta) = \frac{m}{n} \frac{q(\mathbf{w})}{u_{\theta}(\mathbf{w}|\mathbf{c})}$:
- \triangleright $G(\mathbf{w}, \theta)$ is a function of θ , so this ratio changes; However, the **real trick** is if let:

$$\boldsymbol{\theta^*} = \arg\max_{\boldsymbol{\theta}} \frac{1}{n} \left(\sum_{i=1}^n \log \Pr(\mathcal{Y}_i = 1 | \mathbf{w}_i, \boldsymbol{\theta}) + \sum_{i=1}^m \log[\Pr(\mathcal{Y}_i = 0 | \mathbf{w}_i, \boldsymbol{\theta})] \right)$$

and we prove the following: (under large sample size *n* and *m*):

$$G(\mathbf{w}, \frac{\theta^*}{\theta}) \to \frac{m}{n} \frac{q(\mathbf{w})}{p(\mathbf{w})} \implies u_{\theta^*}(\mathbf{w}|\mathbf{c}) \to p(\mathbf{w})$$
 as $\theta \to \theta^*$

▶ let,

$$\begin{split} \mathcal{C}_n(\theta) &= \frac{1}{n} \left(\sum_{i=1}^n \log \Pr(\mathcal{Y}_i = 1 | \mathbf{w}_i, \theta) + \sum_{i=1}^m \log [\Pr(\mathcal{Y}_i = 0 | \mathbf{w}_i, \theta)] \right) \\ &= \frac{1}{n} \sum_{i=1}^n \log \Pr(\mathcal{Y}_i = 1 | \mathbf{w}_i, \theta) + \underbrace{\frac{m}{n} \frac{1}{m} \sum_{i=1}^m \log [\Pr(\mathcal{Y}_i = 0 | \mathbf{w}_i, \theta)]}_{l_{\mathcal{Y}_i}} \end{split}$$

▶ let $n \to \infty$ and $m \to \infty$: $C_n \to C$:

$$\begin{split} \mathcal{C} &= \mathbb{E}_{\mathbf{w} \sim p(\mathbf{w})}[\log \Pr(\mathcal{Y}_i = 1 | \mathbf{w}_i, \theta)] + \nu \mathbb{E}_{\mathbf{w} \sim q(\mathbf{w})}[\log [\Pr(\mathcal{Y}_i = 0 | \mathbf{w}_i, \theta)] \\ &= \mathbb{E}_{\mathbf{w} \sim p(\mathbf{w})} \bigg[\log \frac{1}{1 + G(\mathbf{w}, \theta)} \bigg] + \nu \mathbb{E}_{\mathbf{w} \sim q(\mathbf{w})} \bigg[\log \frac{G(\mathbf{w}, \theta)}{1 + G(\mathbf{w}, \theta)} \bigg] \\ &= - \mathbb{E}_{\mathbf{w} \sim p(\mathbf{w})} \bigg[\log (1 + G(\mathbf{w}, \theta)) \bigg] + \nu \mathbb{E}_{\mathbf{w} \sim q(\mathbf{w})} \bigg[\log G(\mathbf{w}, \theta) - \log (1 + G(\mathbf{w}, \theta)) \bigg] \\ &= - \int \log \big(1 + G(\mathbf{w}, \theta) \big) p(\mathbf{w}) d\mathbf{w} + \nu \int \big(\log G(\mathbf{w}, \theta) - \log (1 + G(\mathbf{w}, \theta)) \big) q(\mathbf{w}) d\mathbf{w} \bigg] \end{split}$$

using functional derivative

$$\mathcal{C} = -\int \log \left(1 + \textit{G}(\mathbf{w}, \theta)\right) p(\mathbf{w}) \mathrm{d}\mathbf{w} + \nu \int \left(\log \textit{G}(\mathbf{w}, \theta) - \log(1 + \textit{G}(\mathbf{w}, \theta))\right) q(\mathbf{w}) \mathrm{d}\mathbf{w}$$

take functional derivative:

$$\begin{split} \frac{\delta \mathcal{C}(G)}{\delta G} &= -\frac{p(\mathbf{w})}{1 + G(\mathbf{w}, \theta)} + \nu q(\mathbf{w}) \left(\frac{1}{G(\mathbf{w})} - \frac{1}{1 + G(\mathbf{w})}\right) \\ &= -\frac{p(\mathbf{w})}{1 + G(\mathbf{w}, \theta)} + \frac{\nu q(\mathbf{w})}{G(\mathbf{w})(1 + G(\mathbf{w}))} = 0 \\ \implies \frac{\nu q(\mathbf{w})}{G(\mathbf{w})(1 + G(\mathbf{w}))} &= \frac{p(\mathbf{w})}{1 + G(\mathbf{w}, \theta)} \\ \implies \frac{\nu q(\mathbf{w})}{G(\mathbf{w})} &= p(\mathbf{w}) \\ \implies G(\mathbf{w}) &= \nu \frac{q(\mathbf{w})}{p(\mathbf{w})} \end{split}$$

Module Five

Probability density re-parameterization

Score Function Estimator

we love to have integral in a form:

$$\mathcal{I} = \int_{z} f(z) p(z) dz \equiv \mathbb{E}_{z \sim p(z)} [f(z)]$$

as we can approximate the expectation with:

$$\mathcal{I} \approx \frac{1}{N} \sum_{i=1}^{N} f(z^{(i)})$$
 $z^{(i)} \sim p(z)$

- we do **not** love $\int_{Y} f(z) \nabla_{\theta} p(z|\theta) dz$,
- ▶ in general, $\nabla_{\theta} p(z|\theta)$ is **not** a probability, e.g., look at derivative of a Gaussian distribution:

$$\frac{\partial}{\partial \mu} \left(\frac{\exp^{-(z-\mu)^2/\sigma^2}}{\sqrt{2\pi}\sigma} \right) = \frac{2(z-\mu)}{\sigma^2} \frac{\exp^{-(z-\mu)^2/\sigma^2}}{\sqrt{2\pi}\sigma}$$



Score Function Estimator

however, in machine learning, we have to deal with:

$$\nabla_{\theta} \left[\int_{z} f(z) p(z|\theta) dz \right] = \int_{z} \nabla_{\theta} \left[f(z) p(z|\theta) \right] dz = \int_{z} f(z) \left[\nabla_{\theta} p(z|\theta) \right] dz$$

- \blacktriangleright i.e, θ is the parameter of the distribution
- e.g., in **Reinforcement Learning**: let $\Pi \equiv \{s_1, a_1, \dots, s_T, a_T\}$

$$p_{\theta}(\Pi) \equiv p_{\theta}(s_1, a_1, \dots s_T, a_T) = p(s_1) \prod_{t=1}^{T} \pi_{\theta}(a_t | s_t) p(s_{t+1} | s_t, a_t)$$

$$\implies \theta^* = \arg \max_{\theta} \left\{ \mathbb{E}_{\Pi \sim p_{\theta}(\Pi)} \left[\underbrace{\sum_{t=1}^{T} R(s_t, a_t)}_{f(z)} \right] \right\}$$

Score Function Estimator

we use REINFORCE trick, with the follow property:

$$p(z|\theta)f(z)\nabla_{\theta}[\log p(z|\theta)] = p(z|\theta)f(z)\frac{\nabla_{\theta}p(z|\theta)}{p(z|\theta)} = f(z)\nabla_{\theta}p(z|\theta)$$

looking at the original integral:

$$\int_{z} f(z) \nabla_{\theta} \rho(z|\theta) dz = \int_{z} \rho(z|\theta) f(z) \nabla_{\theta} [\log \rho(z|\theta)] dz$$
$$= \mathbb{E}_{z \sim \rho(z|\theta)} \left[f(z) \nabla_{\theta} [\log \rho(z|\theta)] \right]$$

can approximated by:

$$\frac{1}{N} \sum_{i=1}^{N} f(z^{(i)}) \nabla_{\theta} [\log p(z^{(i)}|\theta)] \qquad z^{(i)} \sim p(z|\theta)$$

suffers from high variance and is slow to converge



Re-parameterization trick

• we let z = g(x):

$$\begin{split} \mathbb{E}_{x \sim p(x)}[g(x)] &= \mathbb{E}_{z \sim p(z)}[z] \\ \mathbb{E}_{x \sim p(x)}[g(x,\theta)] &= \mathbb{E}_{z \sim p_{\theta}(z)}[z] \quad \text{paramterize the distribution with } \theta \\ \mathbb{E}_{x \sim p(x)}[f(g(x,\theta))] &= \mathbb{E}_{z \sim p_{\theta}(z)}[f(z)] \quad \text{introduce function } f(.) \\ \int_{x \in \Omega_x} f(g(x,\theta))p(x) \mathrm{d}x &= \int_{z \in \Omega_z} f(z)p_{\theta}(z) \mathrm{d}z \end{split}$$

- only need to know deterministic function $z = g(x, \theta)$ and distribution p(x)
- does not need to explicitly know distribution of z
- e.g., Gaussian variable: $z \sim \mathcal{N}(z; \mu(\theta), \sigma(\theta))$ can be rewritten as a function of a standard Gaussian variable:

$$z = g(x, \theta) = \underbrace{\mu(\theta) + x\sigma(\theta)}_{g(x, \theta)} \qquad \text{can be re-parameterised into} \qquad x \sim \underbrace{\mathcal{N}(0, 1)}_{p(x)}$$



revision on change of variable

▶ Let $y = T(x) \implies x = T^{-1}(y)$:

$$F_Y(y) = \Pr(T(X) \le y) = \Pr(X \le T^{-1}(y)) = F_X(T^{-1}(y)) = F_X(x)$$

$$f_Y(y) = \frac{dF_Y(y)}{dy} = \frac{dF_X(x)}{dy} = \frac{dF_X(x)}{dx} \frac{dx}{dy} = f_X(x) \frac{dx}{dy}$$

without change of limits

$$f_Y(y)|dy| = f_X(x)|dx|$$

with change of limits

$$f_Y(y)dy = f_X(x)dx$$



re-parameterization trick (2)

main motivation p(x) is **no longer** parameterized by θ :

$$\begin{split} \mathbb{E}_{x \sim p(x)}[f(g(x,\theta))] &= \int_{x} f(g(x,\theta))p(x)\mathrm{d}x \\ \implies \frac{\partial}{\partial \theta} \mathbb{E}_{x \sim p(x)}[f(g(x,\theta))] &= \frac{\partial}{\partial \theta} \int_{x} f(g(x,\theta))p(x)\mathrm{d}x \\ &= \int_{x} \left[\frac{\partial}{\partial \theta} f(g(x,\theta)) \right] p(x)\mathrm{d}x \\ &\approx \frac{1}{N} \sum_{i=1}^{N} \frac{\partial}{\partial \theta} f(g(x^{(i)},\theta)) \qquad x \sim p(x) \\ &= \frac{1}{N} \sum_{i=1}^{N} \nabla_{\theta} f(g(x^{(i)},\theta)) \qquad \text{use shorthand notation: } \nabla_{\theta}[\cdot] \equiv \frac{\partial}{\partial \theta}[\cdot] \end{split}$$

 \blacktriangleright during gradient decent, x are sampled independent of θ

example¹

let $\mu(\theta) = a\theta + b$, and $\sigma(\theta) = 1$, and we would like to compute:

$$\begin{split} \theta^* &= \arg\max_{\theta}[F(\theta)] \\ &= \arg\min_{\theta} \mathbb{E}_{z \sim \mathcal{N}(\mu(\theta), \sigma(\theta))}[z^2] \\ &= \arg\min_{\theta} \left[\int_{z} \underbrace{z^2}_{f(z)} \mathcal{N} \bigg(\underbrace{a\theta + b}_{\mu(\theta)}, \underbrace{1}_{\sigma(\theta))} \bigg) \right] \end{split}$$

- we can solve it by imagine its diagram . . .
- in words, it says: find mean of Gaussian, so that the "expected square of samples" from this Gaussian are minimized:
- \blacktriangleright it's obvious that you want to move μ to close to **zero** as possible
- which implies $\theta = -\frac{b}{a} \implies \mu(\theta) = 0$
- without using any tricks, the gradient is computed by:

$$\nabla_{\theta} F(\theta) = \int_{z} \underbrace{z^{2}}_{f(z)} \times \underbrace{\frac{2(z-\mu)}{\sigma^{2}} \frac{\exp^{-(z-\mu)^{2}/\sigma^{2}}}{\sqrt{2\pi}\sigma}}_{\underbrace{\frac{\partial \mathcal{N}(\mu, \sigma^{2})}{\partial u}}} \times \underbrace{\underbrace{\frac{\partial \mu}{\partial \theta}}}_{\underbrace{\frac{\partial \mu}{\partial \theta}}} dz$$

very hard!



solve it using REINFORCE trick

- let's solve it by gradient descend by **REINFORCE**:
- let $\mu(\theta) = a\theta + b$, and $\sigma(\theta) = 1$:

$$\begin{split} \int_{z} f(z) \nabla_{\theta} \rho(z|\theta) \mathrm{d}z &= \mathbb{E}_{z \sim \rho(z|\theta)} \big[f(z) \nabla_{\theta} [\log \rho(z|\theta)] \big] \\ &= \mathbb{E}_{z \sim \rho(z|\theta)} \bigg[z^{2} \nabla_{\theta} \log \left(\frac{1}{\sigma \sqrt{2\pi}} \exp^{-\frac{(z-\mu)^{2}}{2\sigma^{2}}} \right) \bigg] \\ &= \mathbb{E}_{z \sim \rho(z|\theta)} \bigg[z^{2} \nabla \mu \bigg[-\log(\sqrt{2\pi}\sigma) - \frac{(z-\mu)^{2}}{2\sigma^{2}} \bigg] \times \frac{\partial \mu(\theta)}{\theta} \bigg] \\ &= \mathbb{E}_{z \sim \mathcal{N} \big(z; a\theta + b, 1 \big)} \big[z^{2} (z - \mu(\theta)) \times a \big] \qquad \text{let } \sigma = 1 \\ &= \mathbb{E}_{z \sim \mathcal{N} \big(z; a\theta + b, 1 \big)} \big[z^{2} a(z - a\theta - b) \big] \end{split}$$

solve it using re-parameterization trick:

- $ightharpoonup z \sim \mathcal{N}(z; \mu(\theta), \sigma(\theta))$ can be **re-parameterised** into:
- if we need to compute: $f(z) = z^2$

$$x \sim \mathcal{N}(0, 1)$$

 $z \equiv g(x, \theta) = \mu(\theta) + x\sigma(\theta)$

the re-parameterised version is:

$$\nabla_{\theta} \mathbb{E}_{x \sim p(x)} [f(g(x, \theta))] \equiv \mathbb{E}_{x \sim \mathcal{N}(x; 0, 1)} [\nabla_{\theta} (z^{2})]$$

$$= \mathbb{E}_{x \sim \mathcal{N}(x; 0, 1)} [\nabla_{\theta} (\mu(\theta) + x\sigma(\theta))^{2}]$$

$$= \mathbb{E}_{x \sim \mathcal{N}(x; 0, 1)} [\nabla_{\theta} (a\theta + b + x)^{2}]$$

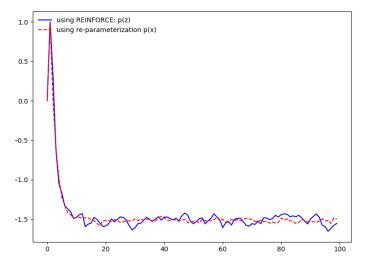
$$= \mathbb{E}_{x \sim \mathcal{N}(x; 0, 1)} [2a(a\theta + b + x)]$$

- both REINFORCE and re-parameterization must achieve the same result!
- knowing p(X) and $g(x, \theta)$ is sufficient, we do **not** need to know explicitly p(Z)



results

ightharpoonup compare both methods using a = 2, b = 3:



apply re-parameterization trick to softmax

when we have the following

$$\begin{split} \mathbb{E}_{K \sim \mathsf{softmax}(\mu_1(\theta), \dots, \mu_L(\theta))}[f(\mathbf{v}(K))] &= \sum_{k=1}^L f(\mathbf{v}(k)) \, \mathsf{Pr}(k|\theta) \\ &\equiv \sum_{k=1}^L f(\mathbf{v}(k)) \big(\mathsf{softmax}(\mu_1(\theta), \dots, \mu_L(\theta)) \big)_k \end{split}$$

can we find their corresponding:

$$\mathcal{K} = g(\mathcal{G}, \theta)$$
 $\mathcal{G} \sim p(\mathcal{G})$

Re-parameterization using Gumbel-max trick

Gumbel-max trick also means:

$$\begin{split} U &\sim \underbrace{\mathcal{U}(0,1)}_{p(\mathcal{G})} \quad \mathcal{G} = -\log(-\log(U)) \\ k &= \underset{i \in \{1,\dots,K\}}{\arg\max} \left\{ \mu_1(\theta) + \mathcal{G}, \dots, \mu_K(\theta) + \mathcal{G} \right\} \\ &= \underbrace{\mathsf{v} = \mathsf{one-hot}(k)}_{q(\mathcal{G},\theta)} \end{split}$$

- ▶ this is a form of re-paramterization: instead of sample $\mathcal{K} \sim \operatorname{softmax}(\mu_1(\theta), \dots, \mu_K(\theta))$, we i.i.d. sample \mathcal{G} instead
- well, there is two problems, firstly why is such true?

Gumbel-max trick and Softmax (1)

pdf of Gumbel with **unit scale** and location parameter μ :

gumbel(
$$Z = z; \mu$$
) = exp $\left[-(z - \mu) - \exp\{-(z - \mu)\} \right]$

CDF of Gumbel:

Gumbel(
$$Z \le Z$$
; μ) = exp $\left[-\exp\{-(Z - \mu)\} \right]$

Gumbel-max trick and Softmax (1)

• given a set of Gumbel random variables $\{Z_i\}$, each having own location parameters $\{\mu_i\}$, probability of all other $Z_{i\neq k}$ are less than a particular value of z_k :

$$p\left(\max\{Z_{i\neq k}\} = \mathbf{Z}_{\mathbf{k}}\right) = \prod_{i\neq k} \exp\left[-\exp\{-(\mathbf{Z}_{\mathbf{k}} - \mu_i)\}\right]$$

▶ obviously, $Z_k \sim \text{gumbel}(Z_k = Z_k; \mu_k)$:

$$\begin{aligned} &\Pr(k \text{ is largest} \mid \{\mu_i\}) \\ &= \int \exp\left\{-(Z_k - \mu_k) - \exp\{-(Z_k - \mu_k)\}\right\} \prod_{i \neq k} \exp\left\{-\exp\{-(Z_k - \mu_i)\}\right\} \, \mathrm{d}Z_k \\ &= \int \exp\left[-Z_k + \mu_k - \exp\{-(Z_k - \mu_k)\}\right] \exp\left[-\sum_{i \neq k} \exp\{-(Z_k - \mu_i)\}\right] \, \mathrm{d}Z_k \\ &= \int \exp\left[-Z_k + \mu_k - \exp\{-(Z_k - \mu_k)\} - \sum_{i \neq k} \exp\{-(Z_k - \mu_i)\}\right] \, \mathrm{d}Z_k \\ &= \int \exp\left[-Z_k + \mu_k - \sum_i \exp\{-(Z_k - \mu_i)\}\right] \, \mathrm{d}Z_k \\ &= \int \exp\left[-Z_k + \mu_k - \sum_i \exp\{-Z_k + \mu_i)\}\right] \, \mathrm{d}Z_k \\ &= \int \exp\left[-Z_k + \mu_k - \exp\{-Z_k\} \sum_i \exp\{\mu_i)\}\right] \, \mathrm{d}Z_k \end{aligned}$$

Gumbel-max trick and Softmax (2)

keep on going:

$$\begin{aligned} \Pr(k \text{ is largest} \mid \{\mu_i\}) &= \int \exp\left[-Z_k + \mu_k - \exp\{-Z_k\} \sum_i \exp\{\mu_i\}\right] dZ_k \\ &= \exp^{\mu_k} \int \exp\left[-Z_k - \exp\{-Z_k\} C\right] dZ_k \\ &= \exp^{\mu_k} \left[\frac{\exp(-C \exp(-Z_k))}{C}\Big|_{Z_k = -\infty}^{\infty}\right] \\ &= \exp^{\mu_k} \left[\frac{1}{C} - 0\right] = \frac{\exp^{\mu_k}}{\sum_i \exp\{\mu_i\}} \end{aligned}$$

Gumbel-max trick and Softmax (2)

moral of the story is, if one is to sample the largest element from softmax:

$$\begin{split} \mathcal{K} \sim & \left\{ \frac{\exp(\mu_1)}{\sum_i \exp(\mu_i)}, \dots, \frac{\exp(\mu_L)}{\sum_i \exp(\mu_i)} \right\} \\ \implies & \mathcal{K} = \underset{i \in \{1, \dots, L\}}{\arg \max} \left\{ G_1, \dots, G_L \right\} \\ & \text{where } G_i \sim \text{gumbel}(z \, ; \, \mu_i) \equiv \exp\left[- (z - \mu_i) - \exp\{-(z - \mu_i)\} \right] \\ \implies & \mathcal{K} = \underset{i \in \{1, \dots, L\}}{\arg \max} \left\{ \mu_1 + \mathcal{G}, \dots, \mu_L + \mathcal{G} \right\} \\ & \text{where } \mathcal{G} \overset{\text{iid}}{\sim} \text{gumbel}(z \, ; \, 0) \equiv \exp\left[- (z) - \exp\{-(z)\} \right] \end{split}$$

- what is μ_i? for example.

 - $\mu_i \equiv \mathbf{x}^{\top} \theta_i$ in classification $\mu_i \equiv \mathbf{u}_i^{\top} \mathbf{v}_c$ for word vectors
- some literature writes it as :

$$\equiv \underset{i \in \{1, \dots, L\}}{\operatorname{arg max}} \left\{ \log(\mu_1) + \mathcal{G}, \dots, \log(\mu_L) + \mathcal{G} \right\}$$

meaning, they let $\mu_i \equiv \exp(\mathbf{x}^{\top} \theta_i)$



how to sample a Gumbel?

CDF of a Gumbel:

$$u = \exp^{-\exp^{-(x-\mu)/\beta}}$$

$$\Rightarrow \log(u) = -\exp^{-(x-\mu)/\beta}$$

$$\Rightarrow \log(-\log(u)) = -(x-\mu)/\beta$$

$$\Rightarrow -\beta\log(-\log(u)) = x - \mu$$

$$\Rightarrow x = \text{CDF}^{-1}(u) \equiv \mu - \beta\log(-\log(u))$$

▶ for standard Gumbel, i.e., $\mu = 0, \beta = 1$:

$$x = \mathsf{CDF}^{-1}(u) \equiv -\log(-\log(u))$$

therefore, sampling strategy:

$$\begin{split} & \mathcal{U} \sim \mathcal{U}(0,1) \\ & \mathcal{G} = -\log(-\log(\mathcal{U})) \\ & \mathcal{K} = \underset{i \in \{1,\dots,K\}}{\text{arg max}} \left\{ \mu_1 + \mathcal{G}, \dots, \mu_L + \mathcal{G} \right\} \\ & \mathbf{v} = \text{one-hot}(\mathcal{K}) \end{split}$$

Second problem with Softmax re-parameterisation

- the other remaining problem: sample v also has an arg max operation, it's a discrete distribution!
- one can relax the softmax distribution, for example softmax map
- several solutions proposed, for example: "Maddison, Mnih, and Teh (2017), The Concrete Distribution: a Continuous Relaxation of Discrete Random Variables"

Relax the Softmax

softmax map

$$\begin{split} f_{\tau}(x)_k &= \frac{\exp(\mu_k/\tau)}{\sum_{k=1}^K \exp(\mu_k/\tau)} \qquad \mu_k \equiv \mu_k(X_k) \\ \text{as } \tau &\to 0 \implies f_{\tau}(x) = \max\left(\left\{\frac{\exp(\mu_k)}{\sum_{k=1}^K \exp(\mu_k)}\right\}_{k=1}^K\right) \end{split}$$

- questions can you also think about the relationship between Gaussian Mixture Model and K-means?
- one can say $\tau = 1$ is softmax, and $\tau = 0$ is hard-max!
- then we can apply the same softmax map with added Gumbel variables:

$$(X_k^{\tau})_k = f_{\tau}(\mu + G)_k = \left(\frac{\exp(\mu_k + G_k)/\tau}{\sum_{i=1}^K \exp(\mu_i + G_i)/\tau}\right)_k$$



Module Six

Stochastic matrices and Monte Carlo Inference

Stochastic matrices

Right stochastic matrix (or row stochastic matrix) is a real square matrix, with each row summing to 1.

$$\begin{bmatrix} K_{1\rightarrow 1} & \dots & K_{1\rightarrow n} \\ \dots & \dots & \dots \\ K_{d\rightarrow 1} & \dots & K_{d\rightarrow n} \\ \dots & \dots & \dots \\ K_{n\rightarrow 1} & \dots & K_{n\rightarrow n} \end{bmatrix}$$

Left stochastic matrix (or column stochastic matrix) is a real square matrix, with each column summing to 1

$$\begin{bmatrix} K_{1\to 1} & \dots & K_{n\to 1} \\ \vdots & \ddots & \ddots \\ K_{1\to d} & \dots & K_{n\to d} \\ \vdots & \ddots & \ddots \\ K_{1\to n} & \dots & K_{n\to n} \end{bmatrix}$$

doubly stochastic matrices: is a real square matrix, where both each column and each row summing to 1.



Product of two stochastic matrix is still stochastic

Each entry in the product AB is a dot product of a row from A and a column from B.

$$(AB)_{ij} = \sum_{k=1}^{n} A_{ik} B_{kj}$$

We need to prove, for a single row of product (AB)_{i,:}

$$\sum_{j=1}^{n} (AB)_{ij} = \sum_{j=1}^{n} \sum_{k=1}^{n} A_{ik} B_{kj} = \sum_{k=1}^{n} (A_{ik} \sum_{j=1}^{n} B_{kj})$$

- ▶ Because *B* is stochastic, $\sum_{i=1}^{n} B_{kj} = 1$
- ▶ Because A is stochastic, $\sum_{k=1}^{n} A_{ik} = 1$

Perron-Frobenius Theorem:

If *K* is a **positive**, left stochastic matrix, then:

- ▶ 1 is an eigenvalue of multiplicity one.
- ▶ 1 is the largest eigenvalue: all the other eigenvalues have absolute value smaller than 1.
- the eigenvectors corresponding to the eigenvalue 1 have either only positive entries or only negative entries.
- ▶ Note that K is a **positive** means, $K_{ij} \ge 0 \ \forall i, j$. It's NOT **positive definite** matrix

Power Method Convergence Theorem

- Let K be a positive, left (i.e., column) stochastic $n \times n$ matrix.
- \blacktriangleright π^* its **probabilistic eigenvector** corresponding to the eigenvalue 1.

$$\begin{bmatrix} K_{1\to 1} & \dots & K_{n\to 1} \\ \dots & \dots & \dots \\ K_{1\to n} & \dots & K_{n\to n} \end{bmatrix} \begin{bmatrix} \pi_1^* \\ \dots \\ \pi_n^* \end{bmatrix} = \begin{bmatrix} \pi_1^* \\ \dots \\ \pi_n^* \end{bmatrix}$$

- Let π⁽¹⁾ be the column vector with all entries equal to some arbitrary stochastic vector.
- ▶ Then sequence $\{\pi^{(1)}, K\pi^{(1)}, K^2\pi^{(1)}, \dots, K^t\pi^{(1)}, \dots, K^\infty\pi^{(1)}\}$ converges to the vector π^*

$$\lim_{t\to\infty} \mathbf{K}^t = \mathbf{K}^{\infty} \implies \lim_{t\to\infty} \mathbf{K}^t \pi^{(1)} = \pi^*$$

- ► Exercise Generate some random matrix in MATLAB and to show an example of the above.
- **Exercise** observe what K^{∞} looks like



Extend to continous case

in the discrete case:

$$\begin{bmatrix} K_{1 \to 1} & K_{2 \to 1} & \dots & K_{n \to 1} \\ \dots & \dots & \dots & \dots \\ K_{1 \to d} & K_{2 \to d} & \dots & K_{n \to d} \\ \dots & \dots & \dots & \dots \\ K_{1 \to n} & K_{2 \to n} & \dots & K_{n \to n} \end{bmatrix} \begin{bmatrix} \pi_1^* \\ \dots \\ \pi_d^* \\ \dots \\ \pi_n^* \end{bmatrix} = \begin{bmatrix} \pi_1^* \\ \dots \\ \pi_d^* \\ \dots \\ \pi_n^* \end{bmatrix} \implies \pi_d^* = \sum_{i=1}^n \pi_i^* K_{i \to d}$$

▶ in the **continous** case, let $\pi(x)$ be the target distribution:

$$\pi(x^{(n+1)}) = \int_{x_n} \pi(x^{(n)}) K(x^{(n)} \to x^{(n+1)})$$

- A transition kernel K contains element-wise entries: $\{K(x^{(n)} \to x^{(n+1)})\}\ \forall x^{(n)}, x^{(n+1)}$
- Sometimes we prefer to write $(x^{(n)}$ as x) and $(x^{(n+1)}$ as $x^*)$.
- $ightharpoonup K(x o x^*)$ is the probability a process at state x moves to state x^* in a **one step**
- $ightharpoonup K^n(x o x^*)$ is the probability a process at state x moves to state x^* in **n steps**



Power Method Convergence in continuous case

▶ One may have first sample $x^{(1)}$ distributed from an arbitrary distribution:

$$x^{(1)} \sim \pi^{(1)}$$

▶ by applying K function, to obtain $x^{(2)}$ given $x^{(1)}$ with probability:

$$\pi^{(2)}(x^{(2)}) = \int_{x^{(1)}} \pi^{(1)}(x^{(1)}) K(x^{(1)} \to x^{(2)}) dx^{(1)}$$

by applying K function again, to obtain $x^{(3)}$ with probability:

$$\begin{split} \pi^{(3)}\big(x^{(3)}\big) &= \int_{x^{(1)}} \int_{x^{(2)}} \pi^{(1)}\big(x^{(1)}\big) \mathcal{K}\big(x^{(1)} \to x^{(2)}\big) \mathcal{K}\big(x^{(2)} \to x^{(3)}\big) \mathrm{d}x^{(1)} \mathrm{d}x^{(2)} \\ &= \int_{x^{(1)}} \pi^{(1)}\big(x^{(1)}\big) \underbrace{\int_{x^{(2)}} \mathcal{K}\big(x^{(1)} \to x^{(2)}\big) \mathcal{K}\big(x^{(2)} \to x^{(3)}\big) \mathrm{d}x^{(2)}}_{} \mathrm{d}x^{(1)} \\ &= \int_{x^{(1)}} \pi^{(1)}\big(x^{(1)}\big) \underbrace{\mathcal{K}^2\big(x^{(1)} \to x^{(3)}\big)}_{} \mathrm{d}x^{(1)} \quad \to \text{converge closer to } \pi\big(x^{(3)}\big) \end{split}$$

This says,

$$\lim_{t\to\infty}\pi^{(t)}\big(x^{(t)}\big)\to\pi(x^{(t)})$$



Burn in samples

We know,

$$\lim_{t\to\infty}\pi^{(t)}\big(\mathbf{X}^{(t)}\big)\to\pi(\mathbf{X}^{(t)})$$

▶ But, in practice,

$$\lim_{t\to B} \pi^{(t)}(\mathbf{x}^{(t)}) \to \pi(\mathbf{x}^{(t)})$$

 $ightharpoonup \{x^{(1)}, \dots, x^{(B)}\}$ are the **burn-in** samples, which we discard.

What is MCMC research is all about

equilibrium equation:

$$\pi(x^*) = \int_x \pi(x) K(x \to x^*) dx$$

- In machine learning, we always know the expression of stationary distribution $\pi(x)$,
- ▶ Our task is therefore, **find an appropriate** $K(x \to x^*)$ to generate samples in a Markov fashion

Detailed Balance

At equilibrium, that stationary distribution satisfies:

$$\pi(x^*) = \int_x \pi(x) K(x \to x^*) dx$$
 equilibrium equation

- Proving equilibrium equation may be difficult in some cases, therefore, we instead prove detail balance:
- detailed balance condition holds when:

$$\pi(x)K(x\to x^*)=\pi(x^*)K(x^*\to x)$$

detailed balance implies equilibrium equation:

$$\begin{split} \int_x \pi(x) \mathcal{K} \big(x \to x^* \big) \mathrm{d} x &= \int_x \pi(x^*) \mathcal{K} \big(x^* \to x \big) \mathrm{d} x \\ &= \pi(x^*) \int_x \mathcal{K} (x^* \to x) \mathrm{d} x \\ &= \pi(x^*) \qquad \text{equilibrium equation} \end{split}$$

the reverse is not always true.



Extend target distribution with auxiliary variables

► At equilibrium, that stationary distribution satisfies:

$$\pi(x^*) = \int_x \pi(x) K(x \to x^*) dx$$

 \triangleright under many scenarios, we may have an extended joint density (x, u):

$$\pi(x|u)\pi(u)K(u,x\to u^*,x^*)=\pi(x^*|u^*)\pi(u^*)K(x^*,u^*\to x,u)$$

- u is auxiliary variables help samping
- one needs to ensure that:

$$\int_{u} \pi(x, u) du = \pi(x)$$

Alternative Use of Stochastic Matrix

- Before dive deep into MCMC algorithms, let's have a look at alternative use of stochastic matrix
- PageRank algorithm is different to MCMC, in PageRank algorithm: K is known
- PageRank algorithm then computes π which is the invariant distribution, tells the importance of each web page.

PageRank algorithm

- Imagine we have the following four web pages and their links
- we can then compute the probability of navigating from ith page (discrete state) to jth page (discrete state)
- ▶ Page 1 links to pages {2,3}

$$\implies K_{1\to 1} = 0, K_{1\to 2} = \frac{1}{2}, K_{1\to 3} = \frac{1}{2}, K_{1\to 4} = 0$$

▶ Page 2 has links to pages {1,3,4}

$$\Longrightarrow K_{2\to 1} = \frac{1}{3}, K_{2\to 2} = 0, K_{2\to 3} = \frac{1}{3}, K_{2\to 4} = \frac{1}{3}$$

Page 3 has links to pages {1,3}

$$\implies K_{3\to 1} = \frac{1}{2}, K_{3\to 2} = 0, K_{3\to 3} = \frac{1}{2}, K_{3\to 4} = 0$$

▶ Page 4 has links to pages {2,3}

$$\implies K_{4 \to 1} = 0, K_{4 \to 2} = \frac{1}{2}, K_{4 \to 3} = \frac{1}{2}, K_{4 \to 4} = 0$$



Stochastic matrix K

From the preceding example, Left stochastic matrix is:

$$\begin{bmatrix} K_{1 \to 1} & K_{2 \to 1} & K_{3 \to 1} & K_{4 \to 1} \\ K_{1 \to 2} & K_{2 \to 2} & K_{3 \to 2} & K_{4 \to 2} \\ K_{1 \to 3} & K_{2 \to 3} & K_{3 \to 3} & K_{4 \to 3} \\ K_{1 \to 4} & K_{2 \to 4} & K_{3 \to 4} & K_{4 \to 4} \end{bmatrix} = \begin{bmatrix} 0 & \frac{1}{3} & \frac{1}{2} & 0 \\ \frac{1}{2} & 0 & 0 & \frac{1}{2} \\ \frac{1}{2} & 1 & 1 & \frac{1}{2} & \frac{1}{2} \\ 0 & \frac{1}{3} & 0 & 0 \end{bmatrix}$$

- From Power Method Convergence Theorem, we know:
 - lacktriangle sequence $\{\pi^{(1)}, K\pi^{(1)}, K^2\pi^{(1)}, \dots, K^t\pi^{(1)}, \dots, K^\infty\pi^{(1)}\}$ converges to the vector π^*

$$\lim_{t\to\infty} K^t \pi^{(1)} = \pi^*$$

where π^* is a **probabilistic eigenvector** of K corresponding to the eigenvalue 1.

Exercise What is the usefulness of π^* in the setting of web pages?



Usefulness of π^* in the setting of web pages

The **answer** to usefulness of π^* in the setting of web pages is:

- ► Shows how **important** each webpage is
- \blacktriangleright i.e., regardless of the probabilities of the initial webpage visit: $\pi^{(1)}$,
- $\pi^{(1)} \to \pi^*$, where $\pi^*(i)$ is the target distribution i.e, the probability that the visit will end up at a web page i.
- ▶ Note that this is a reverse problem of MCMC

Dangling nodes

What happens when you have the following K:

$$\begin{bmatrix} K_{1 \to 1} & K_{2 \to 1} & K_{3 \to 1} & K_{4 \to 1} \\ K_{1 \to 2} & K_{2 \to 2} & K_{3 \to 2} & K_{4 \to 2} \\ K_{1 \to 3} & K_{2 \to 3} & K_{3 \to 3} & K_{4 \to 3} \\ K_{1 \to 4} & K_{2 \to 4} & K_{3 \to 4} & K_{4 \to 4} \end{bmatrix} = \begin{bmatrix} 0 & 0 & 0 & 0 \\ 0 & \frac{1}{2} & 0 & 0 & 0 \\ \frac{1}{2} & 0 & 0 & 0 & 0 \\ \frac{1}{2} & \frac{1}{2} & 1 & 0 \end{bmatrix}$$

- Note that 4th has no out-going node
- ► Exercise check eigenvector correspond to eigenvalue of 1
- What is the eigenvector correspond to eigenvalue of 1, if we change K into:

$$\begin{bmatrix} K_{1 \to 1} & K_{2 \to 1} & K_{3 \to 1} & K_{4 \to 1} \\ K_{1 \to 2} & K_{2 \to 2} & K_{3 \to 2} & K_{4 \to 2} \\ K_{1 \to 3} & K_{2 \to 3} & K_{3 \to 3} & K_{4 \to 3} \\ K_{1 \to 4} & K_{2 \to 4} & K_{3 \to 4} & K_{4 \to 4} \end{bmatrix} = \begin{bmatrix} 0 & 0 & 0 & 0 \\ 0 & \frac{1}{2} & 0 & 0 & 0 \\ \frac{1}{2} & 0 & 0 & 0 & 0 \\ \frac{1}{2} & \frac{1}{2} & 1 & 1 \end{bmatrix}$$

Exercise give reason to why this is so?

Exercise How can we solve this?



Dangling nodes: what may be the solution?

$$\begin{bmatrix} K_{1 \to 1} & K_{2 \to 1} & K_{3 \to 1} & K_{4 \to 1} \\ K_{1 \to 2} & K_{2 \to 2} & K_{3 \to 2} & K_{4 \to 2} \\ K_{1 \to 3} & K_{2 \to 3} & K_{3 \to 3} & K_{4 \to 3} \\ K_{1 \to 4} & K_{2 \to 4} & K_{3 \to 4} & K_{4 \to 4} \end{bmatrix} = \begin{bmatrix} 0 & 0 & 0 & 0 \\ 0 & \frac{1}{2} & 0 & 0 & 0 \\ \frac{1}{2} & 0 & 0 & 0 & 0 \\ \frac{1}{2} & \frac{1}{2} & 1 & 0 \end{bmatrix}$$

One simply solution is:

$$\begin{bmatrix} K_{1 \to 1} & K_{2 \to 1} & K_{3 \to 1} & K_{4 \to 1} \\ K_{1 \to 2} & K_{2 \to 2} & K_{3 \to 2} & K_{4 \to 2} \\ K_{1 \to 3} & K_{2 \to 3} & K_{3 \to 3} & K_{4 \to 3} \\ K_{1 \to 4} & K_{2 \to 4} & K_{3 \to 4} & K_{4 \to 4} \end{bmatrix} = \begin{bmatrix} 0 & 0 & 0 & \frac{1}{3} \\ 0 & \frac{1}{2} & 0 & \frac{1}{3} \\ \frac{1}{2} & 0 & 0 & \frac{1}{3} \\ \frac{1}{2} & \frac{1}{2} & 1 & 0 \end{bmatrix}$$

- in words, it means any page doesn't have out-link, we assume it has equal probability of visiting entire web.
- Of course, data mining researchers may argue certain web page (having certain properties) may attract higher weights etc.



Disconnected sub-graphs

▶ What happens when you have the following K:

$$\begin{bmatrix} K_{1\rightarrow1} & K_{2\rightarrow1} & K_{3\rightarrow1} & K_{4\rightarrow1} \\ K_{1\rightarrow2} & K_{2\rightarrow2} & K_{3\rightarrow2} & K_{4\rightarrow2} \\ K_{1\rightarrow3} & K_{2\rightarrow3} & K_{3\rightarrow3} & K_{4\rightarrow3} \\ K_{1\rightarrow4} & K_{2\rightarrow4} & K_{3\rightarrow4} & K_{4\rightarrow4} \end{bmatrix} = \begin{bmatrix} \frac{1}{2} & 0 & 0 & 0 \\ \frac{1}{2} & 1 & 0 & 0 \\ 0 & 0 & 0 & \frac{1}{2} \\ 0 & 0 & 1 & \frac{1}{2} \end{bmatrix}$$

- ▶ node {1,2} and {3,4} each form a sub-graph.
- Exercise check eigenvector correspond to eigenvalue of 1, also multiplicity of eigenvalue 1
- **Exercise** How can we solve this?



Disconnected sub-graphs: what may be the solution?

$$\begin{bmatrix} K_{1 \to 1} & K_{2 \to 1} & K_{3 \to 1} & K_{4 \to 1} \\ K_{1 \to 2} & K_{2 \to 2} & K_{3 \to 2} & K_{4 \to 2} \\ K_{1 \to 3} & K_{2 \to 3} & K_{3 \to 3} & K_{4 \to 3} \\ K_{1 \to 4} & K_{2 \to 4} & K_{3 \to 4} & K_{4 \to 4} \end{bmatrix} = \begin{bmatrix} \frac{1}{2} & 0 & 0 & 0 \\ \frac{1}{2} & 1 & 0 & 0 \\ 0 & 0 & 0 & \frac{1}{2} \\ 0 & 0 & 1 & \frac{1}{2} \end{bmatrix}$$

▶ One solution is to use a convex combination between K and a square matrix having identical elements $\frac{1}{n}$:

- ▶ in words, it means most of the time 1 p, a surfer will follow links to navigate a page
- but with probability p, it will arbitrarily close the current page and go to the new one
- **Exercise** Prove *K* remains a left stochastic matrix



How to compute the **one hundred billion** dimension eigenvector?

starting from the vector (not probabilistic eigenvector), x:

$$x = [1 \ 1 \ \dots \ 1]^{\top}$$

- ▶ generate the sequence: $\{x, Kx, K^2x \dots K^tx\}$ until convergence then its is the eigenvectors of K correspond to eigenvalue of 1, up to a normalisation constant c
- ► This is solved using power method

Power method

- power method is used to finding an eigenvector of a square matrix corresponding to the largest eigenvalue (in terms of absolute value)
- for stochastic matrix K has eigenvalues:

$$1 = \lambda_1 > |\lambda_2| \ge |\lambda_3| \ge \ldots \ge |\lambda_n|$$

▶ the initial vector: x as a linear combination of eigenvectors of K:

$$X = C_1 V_1 + C_2 V_2 + \dots C_n V_n$$

Then,

$$\begin{aligned} \mathcal{K} x &= \mathcal{K} (c_1 v_1 + c_2 v_2 + \ldots c_n v_n) \\ &= c_1 \underbrace{\lambda_1}_{=1} v_1 + c_2 \lambda_2 v_2 + \ldots c_n \lambda_n v_n \quad \text{definition of eigven value/vector} \\ &= c_1 v_1 + c_2 \lambda_2 v_2 + \ldots c_n \lambda_n v_n \\ &\Longrightarrow \mathcal{K}^2 x = c_1 v_1 + c_2 \lambda_2^2 v_2 + \ldots c_n \lambda_n^2 v_n \\ &\Longrightarrow \mathcal{K}^t x = c_1 v_1 + c_1 \lambda_2^t v_2 + \ldots c_n \lambda_n^t v_n \end{aligned}$$

 $\lambda_i^k \to 0$ when $i > 2 \implies K^t x \to c_1 v_1$

Thank you!

► Questions?