
From Neural PCA to Deep Unsupervised Learning

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Abstract

A network supporting deep unsupervised learning is presented. The network is an autoencoder with lateral shortcut connections from the encoder to decoder at each level of the hierarchy. The lateral shortcut connections allow the higher levels of the hierarchy to focus on abstract invariant features. While standard autoencoders are analogous to latent variable models with a single layer of stochastic variables, the proposed network is analogous to hierarchical latent variables models.

Learning combines denoising autoencoder and denoising sources separation frameworks. Each layer of the network contributes to the cost function a term which measures the distance of the representations produced by the encoder and the decoder. Since training signals originate from all levels of the network, all layers can learn efficiently even in deep networks.

The speedup offered by cost terms from higher levels of the hierarchy and the ability to learn invariant features are demonstrated in experiments.

1 Introduction

Ever since Hubel and Wiesel (1962) published their findings about a hierarchy of increasingly abstract invariant visual features in the cat neocortex, researchers have been trying to mimic the same hierarchy of feature extraction stages with artificial neural networks. An early example is the neocognitron by Fukushima (1979).

Nowadays the art of estimating deep feature extraction hierarchies is called deep learning (for a thorough overview of the history of the field, including recent developments, see the review by Schmidhuber 2015). The topic has received a lot of attention after Hinton and Salakhutdinov (2006); Hinton et al. (2006) proposed an unsupervised¹ pre-training scheme which made subsequent supervised learning efficient for a deeper network than before.

What is somewhat embarrassing for the field, though, is that recently purely supervised learning has achieved as good or better results as unsupervised pre-training schemes (e.g., Ciresan et al. 2010; Krizhevsky et al. 2012). In most classification problems, finding and producing labels for the samples is hard. In many cases plenty of unlabeled data exist and it seems obvious that using them should improve the results. For instance, there are plenty of unlabeled images available and in most image classification tasks there are vastly more bits of information in the statistical structure of input images than in their labels².

¹Unsupervised learning aims at representing structure in the input data, often by means of features. The resulting features can be used as input for classification tasks or as initialization for further supervised learning.

²Consider, for example, the ImageNet classification problem that Krizhevsky et al. (2012) tackled. With 1000 target classes, each label carries less than 10 bits of information. Compare this with the amount of information contained in the 256×256 RGB images used as input. It is impossible to say exactly how many bits of information each image carries but certainly several orders of magnitude more than 10 bits.

It is argued here that the reason why **unsupervised learning has not been able to improve results is that most current versions are incompatible with supervised learning**. The problem is that **many unsupervised learning methods try to represent as much information about the original data as possible whereas supervised learning tries to filter out all the information which is irrelevant for the task at hand**.

This chapter presents an unsupervised learning network whose properties make it a good fit with supervised learning. First, learning is based on minimizing a cost function much the same way as in stochastic gradient descent of supervised feedforward networks. Learning can therefore continue alongside supervised learning rather than be restricted to a pre-training phase. Second, the network can discard information from the higher layers and leave the details for lower layers to represent. This means that the approach allows supervised learning to select the relevant features. The proposed unsupervised learning can filter out noise from the selected features and come up with new features that are related to the selected features.

The two main new ideas presented in this chapter are as follows:

1. Section 2 explains how adding lateral shortcut connections to an autoencoder gives it the same representational capacity as hierarchical latent variable models. This means that higher layers no longer need to represent all the details but can concentrate on abstract invariant representations. The model structure is called a ladder network because two vertical paths are connected by horizontal lateral connections at regular intervals.
2. Learning of deep autoencoders can be slow since training signals need to travel a long distance from the decoder output through both the decoder and encoder. Lateral shortcuts tend to slow it down even further since the shortcuts learn first, shunting the training signals along the longer paths. Section 3 explains how this can be remedied by adding training targets to each level of the hierarchy. The novel idea is to combine denoising source separation (DSS) framework (Särelä and Valpola 2005) with training denoising functions to remove injected noise (Vincent et al. 2008).

The experiments presented in Section 4 demonstrate that the higher levels of a ladder network can discard information and focus on invariant representations and that the training targets on higher layers speed up learning. The results presented here are promising but preliminary. Section 5 discusses potential extensions and related work.

2 Ladder network: an autoencoder which can discard information

As argued earlier, **unsupervised learning needs to tolerate discarding information in order to work well with supervised learning**. Many unsupervised learning methods are not good at this but one class of models stands out as an exception: **hierarchical latent variable models**. **Unfortunately their derivation can be quite complicated and often involves approximations which compromise their performance**. A simpler alternative is offered by autoencoders which also have the benefit of being compatible with standard supervised feedforward networks. They would be a promising candidate for combining supervised and unsupervised learning but **unfortunately autoencoders normally correspond to latent variable models with a single layer of stochastic variables, that is, they do not tolerate discarding information**.

This section summarizes the complementary roles of supervised and unsupervised learning, reviews latent variable models and their relation to standard autoencoder networks and proposes a new network structure, the ladder network, whose **lateral shortcut connections give it the same representational capacity as hierarchical latent variable models**.

2.1 Complementary roles of supervised and unsupervised learning

Consider the roles of supervised and unsupervised learning in a particular task, such as classification, prediction or regression. Further assume that 1) **there are input-output pairs which can be used for supervised learning but far more unlabeled samples that are lacking the output, with only the inputs available** and 2) **inputs have far more information than the outputs**.

In general, this typical setup means that **unsupervised learning should be used for anything it works for because the precious bits of information in the available output samples should be reserved for those tasks that unsupervised learning cannot handle.**

The main role for supervised learning is clear enough: **figure out which type of representations are relevant for the task at hand. Only supervised learning can do this because, by definition, unsupervised learning does not have detailed information about the task.**

One obvious role, the traditional one, for unsupervised learning is to act as a pre-processing or pre-training step for supervised learning. However, the key question is: **what can unsupervised learning do after supervised learning has kicked in.** This is important because in many problems there is so much information in the inputs that it cannot possibly be fully summarized by unsupervised learning first. **Rather, it would be useful for unsupervised learning to continue tuning the representations even after supervised learning has started to tune the relevant features and filter out the irrelevant ones.**

The combination of supervised and unsupervised learning is known as semi-supervised learning. As argued earlier, it can be a happy marriage only if **unsupervised learning is content with discarding information and concentrating on the features which supervised learning deems relevant.**

What unsupervised learning should be able to do efficiently is to **find new features which correlate with and predict the features selected by supervised learning.** This improves generalization to new samples. As an example, consider learning to recognize a face. **Suppose supervised learning has figured out that an eye is an important feature for classifying faces vs. non-faces from a few samples. What unsupervised learning can do with all the available unlabeled samples, is find other features which correlate with the selected one.** Such new features could be, for instance, a detector for nose, eye brow, ear, mouth, and so on. **These features improve the generalization of a face detector in cases where the eye feature is missing, for instance due to eyes being closed or occluded by sunglasses.**

Specifically, what unsupervised learning must *not* do is keep pushing new features to the representation intended for supervised learning simply because these features carry information about the inputs. While such behavior may be reasonable as a pre-training or pre-processing step, it is not compatible with semi-supervised learning. In other words, before unsupervised learning knows which features are relevant, it is reasonable to select features which carry as much information as possible about the inputs. **However, after supervised learning starts showing a preference of some features over some others, unsupervised learning should follow suite and present more of the kind that supervised learning seems to be interested in.**

2.2 Latent variable models

Many unsupervised learning methods can be framed as latent variable models (see, e.g., Bishop 1999) where unknown latent variables $\mathbf{s}(t)$ are assumed to generate the observed data $\mathbf{x}(t)$. A common special case with continuous variables is that the latent variables predict the mean of the observations:

$$\mathbf{x}(t) = g(\mathbf{s}(t); \xi) + \mathbf{n}(t), \quad (1)$$

where $\mathbf{n}(t)$ denotes the noise or modeling error and ξ the parameters of mapping g . Alternatively, the same can be expressed through a probability model

$$p_{\mathbf{x}}(\mathbf{x}(t)|\mathbf{s}(t), \xi) = p_{\mathbf{n}}(\mathbf{x}(t) - g(\mathbf{s}(t); \xi)), \quad (2)$$

where $p_{\mathbf{n}}$ denotes the probability density function of the noise term $\mathbf{n}(t)$. Inference of the unknown latent variables $\mathbf{s}(t)$ and parameters ξ can then be based simply on minimizing the mismatch between the observed $\mathbf{x}(t)$ and its reconstruction $g(\mathbf{s}(t); \xi)$, or more generally on probabilistic modeling.

The models defined by Eqs. (1) or (2) have just one layer of latent variables which tries to represent everything there is to represent about the data. Such models have trouble letting go of any piece of information since this would increase the reconstruction error. Also, in many cases an abstract invariant feature (such as “a face”) cannot reduce the reconstruction error alone without plenty of accompanying details such as position, orientation, size, and so on. All of those details need to be represented alongside the relevant feature to show any benefit in reducing the reconstruction error. This means that latent variable models with a single layer of latent variables have trouble discarding information and focusing on abstract invariant features.

It is possible to fix the situation by introducing a hierarchy of latent variables:

$$p(\mathbf{s}^{(l)}(t) | \mathbf{s}^{(l+1)}(t), \xi^{(l)}), \quad (3)$$

where the superscript (l) refers to variables on layer l . The observations can be taken into the equation by defining $\mathbf{s}^{(0)} := \mathbf{x}$. Now the latent variables on higher levels no longer need to represent everything. Lower levels can take care of representing details while higher levels can focus on selected features, abstract or not.

In such hierarchical models, higher-level latent variables can still represent just the mean of the lower-level variables,

$$\mathbf{s}^{(l)}(t) = g^{(l)}(\mathbf{s}^{(l+1)}(t); \xi^{(l)}) + \mathbf{n}^{(l)}(t), \quad (4)$$

but more generally, the higher-level variables can represent any properties of the distribution, such as the variance (Valpola et al. 2004). For binary variables, sigmoid units are often used for representing the dependency (for a recent example of such a model, see Gregor et al. 2014).

Exact inference, that is, computing the posterior probability of the unknown variables (latent variables and the parameters of the mappings) is typically mathematically intractable. Instead, approximate inference techniques such as variational Bayesian methods are employed (e.g., Valpola et al. 2004; Gregor et al. 2014). They amount to approximating the intractable exact posterior probability with a simpler tractable approximation. Learning then corresponds to iteratively minimizing the cost function with respect to the posterior approximation. For instance, in the case of continuous latent variables, the posterior could be approximated as Gaussian with a diagonal covariance. For each unknown variable, the mean and variance would then be estimated in the course learning. These posterior means and variances typically depend on the values on both lower and higher layers and inference would therefore proceed iteratively.

If hierarchical latent variable models were easy to define and learn, the problem would be solved. Unfortunately hierarchical models often require complex probabilistic methods to train them. They often involve approximations which compromise their performance (Ilin and Valpola 2003) or are limited to restricted model structures which are mathematically tractable. Also, many training schemes require the latent variable values to be updated layer-wise by combining bottom-up information with top-down priors. This slows down the propagation of information in the network.

2.3 Autoencoders and deterministic and stochastic latent variables

Autoencoder networks resemble in many ways single-layer latent variable models. The key idea is that the inference process of mapping observations $\mathbf{x}(t)$ to the corresponding latent variables, now called hidden unit activations $\mathbf{h}(t)$, is modeled by an encoder network f and the mapping back to observations is modeled by a decoder network g :

$$\mathbf{h}(t) = f(\mathbf{x}(t); \xi_f) \quad (5)$$

$$\hat{\mathbf{x}}(t) = g(\mathbf{h}(t); \xi_g). \quad (6)$$

The mappings f and g are called encoder and decoder mappings, respectively. In connection to latent variable models, analogous mappings are called the recognition and reconstruction mappings.

Learning of autoencoders is based on minimizing the difference between the observation vector $\mathbf{x}(t)$ and its reconstruction $\hat{\mathbf{x}}(t)$, that is, minimizing the cost $\|\mathbf{x}(t) - \hat{\mathbf{x}}(t)\|^2$ with respect to the parameters ξ_f and ξ_g . For the remainder of this chapter, all mappings f and g are assumed to have their own parameters but they are omitted for brevity.

Just like latent variable models, autoencoders can be stacked together:

$$\mathbf{h}^{(l)}(t) = f^{(l)}(\mathbf{h}^{(l-1)}(t)) \quad (7)$$

$$\hat{\mathbf{h}}^{(l-1)}(t) = g^{(l)}(\hat{\mathbf{h}}^{(l)}(t)). \quad (8)$$

As before, the observations are taken into the equation by defining $\mathbf{h}^{(0)} := \mathbf{x}$. Furthermore, now $\hat{\mathbf{h}}^{(L)} := \mathbf{h}^{(L)}$ for the last layer L , connecting the encoder and decoder paths.

Typically over the course of learning, new layers are added to the previously trained network. After adding and training the last layer, training can continue in a supervised manner using just the mappings $f^{(l)}$, which define a multi-layer feedforward network, and minimizing the squared distance between the actual outputs $\mathbf{h}^{(L)}$ and desired targets outputs.

It is tempting to assume that the hierarchical version of the autoencoder in Eqs. (7–8) corresponds somehow to the hierarchical latent variable model in Eq. (4). Unfortunately this is not the case because the intermediate hidden layers $0 < l < L$ act as so called deterministic variables while the hierarchical latent variable model requires so called stochastic variables. The difference is that stochastic variables have independent representational capacity. No matter what the priors tell, stochastic latent variables $s^{(l)}$ can overrule this and add their own bits of information to the reconstruction. By contrast, deterministic variables such as $\hat{h}^{(l)}$ add zero bits of information and, assuming the deterministic mappings $g^{(l)}$ are implemented as layered networks, correspond to the hidden layers of the mappings $g^{(l)}$ between the stochastic variables $s^{(l)}$ in Eq. (3).

In order to fix this, we are going to take cue from the inference structure of the hierarchical latent variable model in Eq. (3). The main difference to Eq. (8) is that inference of $\hat{s}^{(l)}(t)$ combines information from bottom-up likelihood and top-down prior but Eq. (8) only depends on top-down information. In other words, $\hat{h}^{(l)}(t)$ in Eq. (8) cannot add any new information to the representation because it does not receive that information from the bottom-up path. The fix is to add a shortcut connection from the bottom-up encoder path to the modified top-down decoder path:

$$\hat{\mathbf{h}}^{(l-1)}(t) = g^{(l)}(\hat{\mathbf{h}}^{(l)}(t), \mathbf{h}^{(l-1)}(t)). \quad (9)$$

Now $\hat{\mathbf{h}}^{(l)}$ can recover information which is missing in $\hat{\mathbf{h}}^{(>l)}$. In other words, the higher layers do not need to represent all the details. Also, the mapping $g^{(l)}$ can learn to combine abstract information from higher levels, such as “face”, with detailed information about position, orientation, size, and so on, from lower layers³. This means that the higher layers can focus on representing abstract invariant features if they seem more relevant to the task at hand than the more detailed information.

Figure 1 shows roughly the inference structure of a hierarchical latent variable model and compares it with a standard autoencoder and the ladder network. Note that while $\hat{\mathbf{h}}^{(l)}(t)$ combines information both from bottom-up and top-down paths in the ladder network, $\mathbf{h}^{(l)}(t)$ does not. This direct path from inputs to the highest layer means that training signals from the highest layers can propagate directly through the network in the same way as in supervised learning. Gradient propagation already combines information from bottom-up activations and top-down gradients so there is no need for extra mixing of information.

3 Parallel learning on every layer

A general problem with deep models which have an error function only at the input layer (autoencoder) or at the output layer (supervised feedforward models) is that many parts of the network are far away from the source of training signals. In fact, if the ladder model shown in Fig. 1 is trained in the same fashion as regular autoencoders, that is, by minimizing the difference between $\mathbf{x}(t)$ and $\hat{\mathbf{x}}(t)$, the problem only becomes worse. That is because each shortcut connection has a chance of contributing to the reconstruction $\hat{\mathbf{x}}(t)$, leaving a shrinking share of the error for the higher layers. Standard autoencoders force all training signals to pass through all levels in the hierarchy and even then learning through multiple layers of nonlinear functions is difficult and slow.

By contrast, hierarchical latent variable models have cost functions for all stochastic variables. Since the ladder network shares many properties with hierarchical latent variable models, it seems reasonable to try to find a way to introduce training signals at each level of the hierarchy of the ladder network. This section shows how this can be done by combining denoising source separation (DSS) framework (Särelä and Valpola 2005) with training denoising functions to remove injected noise (Vincent et al. 2008).

³The shortcut connections are only part of the solution since adding them actually short-circuits standard autoencoder learning: mapping $g^{(l)}$ can simply copy the shortcut input $\mathbf{h}^{(l-1)}(t)$ into its output so there is no incentive to use any of the higher layers. This is related to the inability of standard autoencoder training to learn over-complete representations. Fortunately, denoising autoencoders can overcome this problem simply by adding noise to inputs as will be explained in Section 3.2.

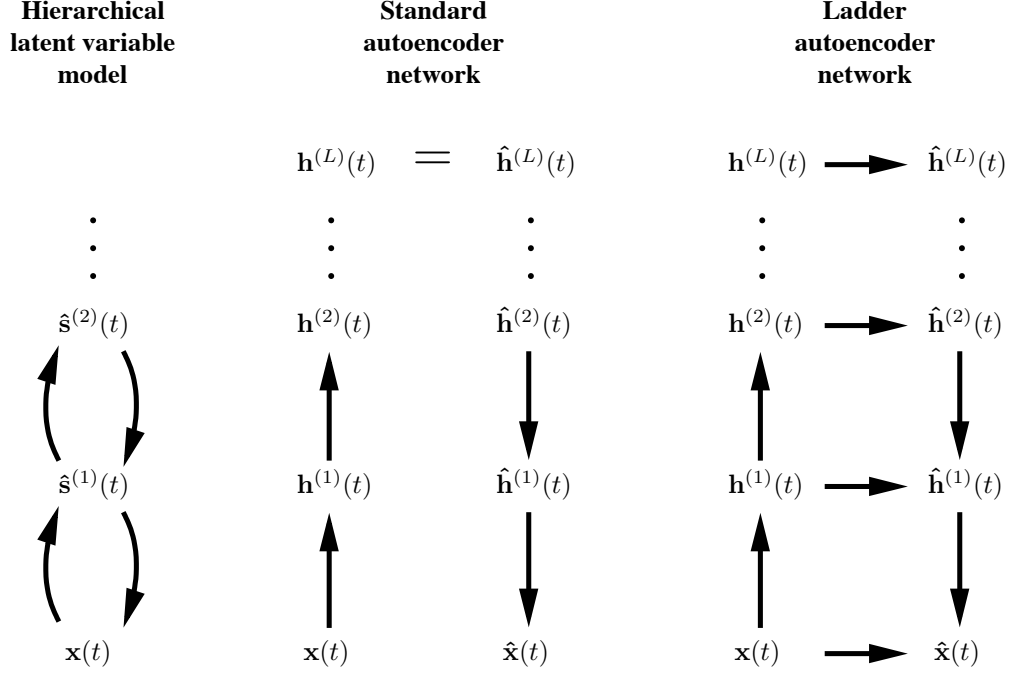


Figure 1: The inference structure of a hierarchical latent variable model is compared with the standard autoencoder and the proposed ladder network. The details of inference in latent variable models are often complex and the posterior distribution approximation is more complex than just the posterior mean $\hat{\mathbf{s}}^{(l)}(t)$, but overall the picture is approximately as shown in the left. Since all information in the standard autoencoder network has to go through the highest layer, it needs to represent all the details of the input $\mathbf{x}(t)$. Intermediate hidden layer activations $\hat{\mathbf{h}}^{(l)}(t)$ cannot independently represent information because they only receive information from the highest layer. In the ladder network, by contrast, lateral connections at each layer give a chance for each $\hat{\mathbf{h}}^{(l)}(t)$ to represent information independently from the higher layers. Also, abstract invariant representations at the higher levels can be interpreted in the context of detailed information without the higher levels having to represent all the details.

3.1 From neural PCA to denoising source separation

In order to develop a system where learning is distributed rather than guided by gradients propagating from a single error term, we shall turn our attention to competitive unsupervised learning.

The starting point for the algorithms we will study is the neural principal component analysis (PCA) learning rule by Oja (1982) which can utilize second order statistics of the input data to find principal component projections. When slight nonlinear modifications are made to the learning rule and input data are whitened, the method becomes sensitive to higher-order statistics and performs independent component analysis (ICA) (Oja 1997).

The nonlinearity used in the algorithm can be interpreted as a contrast function which measures the non-Gaussianity of the source distribution. This is the interpretation originally given to the popular FastICA algorithm (Hyvärinen and Oja 1997). However, there is an alternative view: the nonlinearity can be interpreted as a denoising function. Hyvärinen (1998) derived this as a maximum likelihood estimate but we are going to follow the derivation by Valpola and Pajunen (2000) who showed that the nonlinearity can be interpreted as the expectation step of the expectation maximization algorithm. Overall, nonlinear PCA learning rule, combined with input whitening and orthogonalization of the projections, can be interpreted as an efficient approximation to the expectation maximization (EM) algorithm applied to a linear latent variable model tuned for ICA (Valpola and Pajunen 2000). This interpretation led to the development of denoising source separation framework (Särelä and Valpola 2005).

The EM algorithm (Dempster et al. 1977) is a method for optimizing the parametric mappings of latent variable models. It operates by alternating the E step (expectation of latent variables) and M step (maximization of likelihood of the parameters). The E step assumes the mapping fixed and updates the posterior distribution of $\mathbf{s}(t)$ for all t while the M step does the reverse, updating the mapping while assuming the posterior distribution of $\mathbf{s}(t)$ fixed.

The derivation by Valpola and Pajunen (2000) assumed a linear reconstruction model

$$\hat{\mathbf{x}}(t) = g^{(0)}(\mathbf{s}(t)) = \mathbf{A}\mathbf{s}(t), \quad (10)$$

which means that the E step boils down to

$$\hat{\mathbf{s}}(t) = g^{(1)}(\mathbf{A}^{-1}\mathbf{x}(t)) = g^{(1)}(\mathbf{s}_0(t)), \quad (11)$$

where we have denoted $\mathbf{s}_0(t) = \mathbf{A}^{-1}\mathbf{x}(t)$. The mapping $g^{(1)}$ depends on the prior distribution of $\mathbf{s}(t)$ and the noise distribution p_n . When the noise has low variance σ_n^2 , it can be approximated as

$$\hat{\mathbf{s}}(t) = g^{(1)}(\mathbf{s}_0(t)) \approx \mathbf{s}_0(t) + \sigma_n^2 \frac{\partial \log p_s(\mathbf{s}(t))}{\partial \mathbf{s}(t)} \Big|_{\mathbf{s}(t)=\mathbf{s}_0(t)}. \quad (12)$$

The M step amounts to solving the regression problem in Eq. (10) with $\hat{\mathbf{s}}(t)$ substituting $\mathbf{s}(t)$, that is, minimizing the cost

$$C = \frac{1}{T} \sum_{t=1}^T \|\mathbf{x}(t) - g^{(0)}(\hat{\mathbf{s}}(t))\|^2. \quad (13)$$

The problem with the above EM algorithm for estimating the model is that its convergence scales with σ_n^2 . Without noise, the algorithm stalls completely.

Valpola and Pajunen (2000) showed that this can be remedied by considering the fixed point of the algorithm. At the same time, it is useful to parametrize the inverse mapping

$$\mathbf{s}_0(t) = f^{(1)}(\mathbf{x}(t)) = \mathbf{W}\mathbf{x}(t) \quad (14)$$

and for M step minimize

$$C = \frac{1}{T} \sum_{t=1}^T \|\mathbf{s}_0(t) - \hat{\mathbf{s}}(t)\|^2 = \frac{1}{T} \sum_{t=1}^T \|f^{(1)}(\mathbf{x}(t)) - \hat{\mathbf{s}}(t)\|^2 \quad (15)$$

instead of Eq. (13). Equation (15) needs to be accompanied by constraints on the covariance of $\mathbf{s}_0(t)$ because the trivial solution $\mathbf{W} = \mathbf{0}$ and $\hat{\mathbf{s}}(t) = \mathbf{0}$ yields $C = 0$.

The algorithm resulting from these assumptions is essentially the same as the nonlinear PCA learning rule. When the input data are whitened, the M step becomes simple matrix multiplication just as in the nonlinear PCA learning rule, amounting to essentially Hebbian learning. The nonlinearity $g^{(1)}(s_0(t))$ has the interpretation that it is the expected value of the latent variables given noisy observations, that is denoising.

What will be crucial for our topic, learning deep models, is that the cost function (15) does not directly refer to the input $\mathbf{x}(t)$ but only to the latent variable $\mathbf{s}(t)$ and its denoised version. In a hierarchical model this will mean that each layer contributes terms to the cost function, bringing the source of training signals close to the parameters on each layer.

Just as with the nonlinear PCA learning rule, there needs to be an additional constraint which implements competition between the latent variables because they could otherwise all converge to the same values. In case of a linear model, the easiest approach is to require \mathbf{W} to be orthogonal. This does not apply to nonlinear models. Instead, it is possible to require that the covariance matrix of the latent variables $\mathbf{s}(t)$ is a unit matrix (Almeida et al. 2006).

3.2 Denoising autoencoders and generative stochastic networks

The denoising function used in Eq. (11) can be derived from the prior distribution p_s of the latent variables. There are many techniques for learning such distributions but a particularly useful technique that directly learns the denoising function was proposed by Vincent et al. (2008) in connection with autoencoders. The idea is to corrupt the inputs fed into the autoencoder with noise and ask the network to reconstruct the original uncorrupted inputs. This forces the autoencoder to learn how to denoise the corrupted inputs.

Bengio et al. (2013) further showed that it is possible to sample from these so called denoising autoencoders simply by iterating corruption and denoising. The distribution of the denoised samples converges to the original data distribution because during training, the denoising function learns to cancel the diffusion resulting from the corruption of input data. The diffusive forces are proportional to $-\frac{\partial \log p_{\mathbf{x}}(\mathbf{x})}{\partial \mathbf{x}}$ and, on average, carry samples from areas of high density towards lower densities. The denoising function learns to oppose this, with the same force but opposite sign⁴. When sampling starts with any given distribution, the combined steps of corruption and denoising produce an average flow of samples which only disappears when the diffusion flow caused by corruption exactly cancels the flow caused by denoising, that is, when the sample distribution follows the original training distribution. Bengio et al. (2013) suggested that sampling is more efficient in hierarchical models if corruption takes place not only on inputs but on all levels of the encoder path and called such networks generative stochastic networks (GSN).

What is surprising is that from denoising functions it is even possible to derive probability estimates for the data. Note that the denoising function loses information about absolute probability and only conserves information about relative probabilities because the logarithm first turns multiplication into summation and the constant normalization term then disappears in differentiation. Such a representation bears similarity to energy-based probability models where only relative probabilities can be readily accessed. It turns out, however, that any model which can reconstruct missing data can be turned into a probability density estimator (Uribe et al. 2014). By using input erasure as corruption, the autoencoder can thus be used for deriving normalized probability estimates even if denoising function loses information about the normalization factor of the probability.

For our purposes, a particularly important feature of denoising autoencoders is that they can handle over-complete representations, including the shortcut connections in Eq. (9). The reason for this is that it is not enough for the network to simply copy its inputs to outputs since the inputs are corrupted by noise. Rather, the network has to find a representation which makes removing noise as easy as possible.

3.3 Recursive derivation of the learning rule

We are now ready to derive a learning rule with a distributed cost function for the ladder network. The basic idea is to apply a denoising autoencoder recursively. The starting point is the standard

⁴Notice the similarity to the denoising in Eq. (12).

denoising autoencoder which minimizes the following cost C :

$$\tilde{\mathbf{x}}(t) = \text{corrupt}(\mathbf{x}(t)) \quad (16)$$

$$\hat{\mathbf{x}}(t) = g(\tilde{\mathbf{x}}(t)) \quad (17)$$

$$C = \frac{1}{T} \sum_{t=1}^T \|\mathbf{x}(t) - \hat{\mathbf{x}}(t)\|^2. \quad (18)$$

During learning the denoising function g learns to remove the noise which we are injecting to corrupt $\mathbf{x}(t)$. Now assume that the denoising function uses some internal variables $\mathbf{h}^{(1)}(t)$ for implementing a multi-layer mapping:

$$\tilde{\mathbf{h}}^{(1)}(t) = f^{(1)}(\tilde{\mathbf{x}}(t)) \quad (19)$$

$$\hat{\mathbf{x}}(t) = g(\tilde{\mathbf{h}}^{(1)}(t)). \quad (20)$$

Rather than giving all the responsibility of denoising to g , it is possible to learn first how to denoise $\mathbf{h}^{(1)}$ and then use that result for denoising \mathbf{x} :

$$\mathbf{h}^{(1)}(t) = f^{(1)}(\mathbf{x}(t)) \quad (21)$$

$$\tilde{\mathbf{h}}^{(1)}(t) = f^{(1)}(\tilde{\mathbf{x}}(t)) \quad (22)$$

$$\hat{\mathbf{h}}^{(1)}(t) = g^{(1)}(\tilde{\mathbf{h}}^{(1)}(t)) \quad (23)$$

$$\hat{\mathbf{x}}(t) = g^{(0)}(\hat{\mathbf{h}}^{(1)}(t)) \quad (24)$$

$$C^{(1)} = \frac{1}{T} \sum_{t=1}^T \|\mathbf{h}^{(1)}(t) - \hat{\mathbf{h}}^{(1)}(t)\|^2 \quad (25)$$

$$C^{(0)} = \frac{1}{T} \sum_{t=1}^T \|\mathbf{x}(t) - \hat{\mathbf{x}}(t)\|^2. \quad (26)$$

Training could alternate between training the mapping $g^{(1)}$ by minimizing $C^{(1)}$ and training all the mappings by minimizing $C^{(0)}$.

We can continue adding layers and also add the lateral connections of the ladder network:

$$\mathbf{h}^{(l)}(t) = f^{(l)}(\mathbf{h}^{(l-1)}(t)) \quad (27)$$

$$\tilde{\mathbf{h}}^{(l)}(t) = f^{(l)}(\tilde{\mathbf{h}}^{(l-1)}(t)) \quad (28)$$

$$\hat{\mathbf{h}}^{(l)}(t) = g^{(l)}(\tilde{\mathbf{h}}^{(l)}(t), \hat{\mathbf{h}}^{(l+1)}(t)) \quad (29)$$

$$C^{(l)} = \frac{1}{T} \sum_{t=1}^T \|\mathbf{h}^{(l)}(t) - \hat{\mathbf{h}}^{(l)}(t)\|^2. \quad (30)$$

As before, we assume that $\mathbf{h}^{(0)}$ refers to the observations \mathbf{x} .

This derivation suggests that the cost functions $C^{(l)}$ should only be used for learning the encoding mappings $f^{(>l)}$ of the layers above but not $f^{(\leq l)}$, those of the layers below, because $C^{(l)}$ is derived assuming $\mathbf{h}^{(l)}(t)$ is fixed. This is problematic because it means that there cannot be a single consistent cost function for the whole learning process and learning has to continuously alternate between learning different layers.

However, recall that in the DSS framework it is precisely the forward mapping $f^{(l)}$ which is updated using Eq. (15), practically the same equation as Eq. (30). This means that we can in fact minimize a single cost function

$$C = C^{(0)} + \sum_{l=1}^L \alpha_l C^{(l)}, \quad (31)$$

where the coefficients α_l determine the relative weights of the cost terms originating in different layers.

The DSS framework assumes that $\hat{\mathbf{h}}^{(l)}$ is constant and optimizes using gradients stemming from $\mathbf{h}^{(l)}$ while in the denoising autoencoder framework the roles are reversed. This means that by combining

the cost functions and learning everything by minimizing the cost with respect to all the parameters of the model, we are essentially making use of both types of learning.

Just like in hierarchical latent variable models, higher level priors offer guidance to lower-level forward mappings (cf. EM algorithm). Since the gradients propagate backward along the encoding path, this model is fully compatible with supervised learning: the standard supervised cost function can simply be added to the top-most layer L , measuring the distance between $\mathbf{h}^{(L)}(t)$ and the target output.

3.4 Decorrelation term for the cost function

There is one final thing we must take care of: the decorrelation term needed by DSS algorithms. Recall that Eq. (30) is minimized if $\mathbf{h}^{(l)}(t) = \hat{\mathbf{h}}^{(l)}(t) = \text{constant}$. Minimization of Eq. (30) with respect to $\hat{\mathbf{h}}^{(l)}(t)$ actually typically promotes decorrelation because it amounts to regression and any extra information can be used to reduce the reconstruction error. Minimization of Eq. (30) with respect to $\mathbf{h}^{(l)}(t)$ promotes finding projections that can be predicted as well as possible and, since mutual information is symmetric, therefore also help predicting other features as long as the entropy of the hidden unit activations is kept from collapsing by avoiding the trivial solution where $\mathbf{h}(t) = \text{constant}$.

We are going to assume that the mappings $f^{(l)}$ and $g^{(l)}$ are sufficiently general that we can, without loss of generality, assume that the covariance matrix $\Sigma^{(l)}$ of the hidden unit activations on layer l equals the unit matrix: $\Sigma^{(l)} = \mathbf{I}$, where

$$\Sigma^{(l)} = \frac{1}{T} \sum_{t=1}^T \mathbf{h}^{(l)}(t) [\mathbf{h}^{(l)}(t)]^T. \quad (32)$$

Here we assumed that the average activation is zero, a constraint which we can also enforce without any loss of generality as long as the first stages of mappings $f^{(l)}$ and $g^{(l)}$ are affine transformations.

A very simple cost function to promote $\Sigma^{(l)} \approx \mathbf{I}$ would be $\sum_{i,j} [\Sigma_{ij}^{(l)} - \delta_{ij}]^2$, where δ_{ij} is the Kronecker delta. In other words, this measures the sum of squares of the difference between $\Sigma^{(l)}$ and \mathbf{I} . However, this cost function does not distinguish between too small and too large eigenvalues of $\Sigma^{(l)}$ but from the viewpoint of keeping the DSS-style learning from collapsing the representation of $\mathbf{h}^{(l)}(t)$, only too small eigenvalues pose a problem. To analyse the situation, note that

$$\sum_{i,j} [\Sigma_{ij}^{(l)} - \delta(i,j)]^2 = \text{tr} \left([\Sigma^{(l)} - \mathbf{I}]^2 \right) = \sum_i (\lambda_i^{(l)} - 1)^2, \quad (33)$$

where $\lambda_i^{(l)}$ are the eigenvalues of $\Sigma^{(l)}$. The first equality follows from the definition of the trace of a matrix and the second from the fact that the trace equals the sum of eigenvalues.

Since Eq. (33) is symmetric about $\lambda = 1$, it penalizes $\lambda = 0$ just as much as $\lambda = 2$ while the former is infinitely worse from the viewpoint of keeping \mathbf{h} from collapsing.

A sound measure for the information content of a variable is the determinant of the covariance matrix because it measures the square of the (hyper)volume of the (hyper)cuboid whose sides have the length determined by the standard deviations of the distribution along its eigenvectors. Since the determinant of a matrix equals the product of its eigenvalues, the logarithm of the determinant equals the sum of the logarithms of the eigenvalues:

$$\log \det \Sigma^{(l)} = \sum_i \log \lambda_i^{(l)} = \text{tr} \left(\log \Sigma^{(l)} \right). \quad (34)$$

The latter equality follows from the fact that any analytical function can be defined for square matrices so that it applies to the eigenvalues of the matrix. This is because $(\mathbf{E}\mathbf{\Lambda}\mathbf{E}^{-1})^k = \mathbf{E}\mathbf{\Lambda}^k\mathbf{E}^{-1}$ and therefore any power series expansion of a matrix turns into the same power series expansion of the eigenvalues. Note that $\log \Sigma$ is the matrix logarithm, *not* the logarithm of the elements of the matrix.

Equation (34) is a measure which grows smaller when the information content diminishes but it can be turned into a sensible cost function which reaches its minimum value of zero when $\lambda = 1$:

$$C_{\Sigma}^{(l)} = \sum_i (\lambda_i^{(l)} - \log \lambda_i^{(l)} - 1) = \text{tr} \left(\Sigma^{(l)} - \log \Sigma^{(l)} - \mathbf{I} \right) \quad (35)$$

This cost penalizes $\lambda = 0$ infinitely and grows relatively modestly for $\lambda_i > 1$.

It is relatively simple to differentiate this cost with respect to $\Sigma^{(l)}$ since, for any analytical function ϕ , it holds

$$\frac{\partial \text{tr}(\phi(\Sigma))}{\partial \Sigma} = \phi'(\Sigma). \quad (36)$$

In our case $\phi(a) = a - \log a - 1$ and thus $\phi'(a) = 1 - a^{-1}$. We therefore have

$$\frac{\partial C_{\Sigma}^{(l)}}{\partial \Sigma^{(l)}} = \mathbf{I} - [\Sigma^{(l)}]^{-1}. \quad (37)$$

The rest of the formulas required for computing the gradients with the chain rule are straight-forward since Eq. (32) has a simple quadratic form.

Note that all twice differentiable cost functions that are minimized when $\lambda_i = 1$ have the same second-order behaviour (up to scaling) close to the minimum so the simpler Eq. (33) works just as well if all λ_i are sufficiently close to 1. However, to avoid any potential problems, Eq. (35) was used in the experiments presented in this chapter.

Finally, as suggested earlier, we will add a simple term to the cost function to make sure that the hidden unit activations really have a zero mean:

$$\mu^{(l)} = \frac{1}{T} \sum_{t=1}^T \mathbf{h}^{(l)}(t) \quad (38)$$

$$C_{\mu}^{(l)} = \|\mu^{(l)}\|^2. \quad (39)$$

While DSS algorithms require decorrelation of the output representation, it is now also important to decorrelate (i.e., whiten) the inputs. Normally denoising autoencoders have a fixed input but now the cost functions on the higher layers can influence their input mappings and this creates a bias towards PCA-type solutions. This is because the amount of noise injected to \mathbf{x} is relatively smaller for projections for which the variance of \mathbf{x} is larger. The terms $C^{(\geq 1)}$ are therefore smaller if the network extracts mainly projections with larger variance, that is, PCA-type solutions. While PCA may be a desirable in some cases, often it is not.

3.5 Learning rule for the ladder network

We are now ready to collect together the recipe for learning the ladder network. Given (typically pre-whitened) observations $\mathbf{h}^{(0)}(t) := \mathbf{x}(t)$, the cost function C is computed using the following

formulas:

$$\mathbf{h}^{(l)}(t) = f^{(l)}(\mathbf{h}^{(l-1)}(t)) \quad \text{for } 1 \leq l \leq L \quad (40)$$

$$\tilde{\mathbf{h}}^{(0)}(t) = \text{corrupt}(\mathbf{h}^{(0)}(t)) \quad (41)$$

$$\tilde{\mathbf{h}}^{(l)}(t) = f^{(l)}(\tilde{\mathbf{h}}^{(l-1)}(t)) \quad \text{for } 1 \leq l \leq L \quad (42)$$

$$\hat{\mathbf{h}}^{(L)}(t) = g^{(L)}(\tilde{\mathbf{h}}^{(L)}(t)) \quad (43)$$

$$\hat{\mathbf{h}}^{(l)}(t) = g^{(l)}(\tilde{\mathbf{h}}^{(l)}(t), \hat{\mathbf{h}}^{(l+1)}(t)) \quad \text{for } 0 \leq l \leq L-1 \quad (44)$$

$$C^{(l)} = \frac{1}{T} \sum_{t=1}^T \|\mathbf{h}^{(l)}(t) - \hat{\mathbf{h}}^{(l)}(t)\|^2 \quad (45)$$

$$\Sigma^{(l)} = \frac{1}{T} \sum_{t=1}^T \mathbf{h}^{(l)}(t) [\mathbf{h}^{(l)}(t)]^T \quad (46)$$

$$C_{\Sigma}^{(l)} = \text{tr}(\Sigma^{(l)} - \log \Sigma^{(l)} - \mathbf{I}) \quad (47)$$

$$\mu^{(l)} = \frac{1}{T} \sum_{t=1}^T \mathbf{h}^{(l)}(t) \quad (48)$$

$$C_{\mu}^{(l)} = \|\mu^{(l)}\|^2 \quad (49)$$

$$C = C^{(0)} + \sum_{l=1}^L \alpha_l C^{(l)} + \beta_l C_{\Sigma}^{(l)} + \gamma_l C_{\mu}^{(l)}. \quad (50)$$

Learning the parameters of the mappings $f^{(l)}$ and $g^{(l)}$ is based on minimizing C . The simple solution is to apply gradient descent (stochastic or batch version) but basically any optimization method can be used, for example nonlinear conjugate gradient or quasi-Newton methods. Whichever method is chosen, the existence of a single cost function to be minimized guarantees that learning converges as long as the minimization method performs properly.

Equation (42) could include corruption in the same manner as in GSN. However, to keep things simple, the experiments reported in this chapter only applied corruption to the input layer.

Figure 2 depicts the computational diagram of the cost function of the ladder network. The two f paths going upward share their mappings $f^{(l)}$ and the only difference is that the inputs on the corrupted path are corrupted by noise. Each layer of the network adds its own term to the cost function, measuring how well the clean activations $\mathbf{h}^{(l)}(t)$ are reconstructed from the corrupted activations. During forward computations, information will flow from the observations towards the cost function terms along the arrows. During learning, gradients will flow from the cost function terms in the opposite direction. Training signals arriving along the clean f path correspond to DSS-style learning while the training signals in the denoising g path and corrupted f path correspond to the type of learning taking place in denoising autoencoders. The terms $C_{\Sigma}^{(l)}$ and $C_{\mu}^{(l)}$ which promote unit covariance and zero mean of the clean activations $\mathbf{h}^{(l)}(t)$, respectively, are not shown. They are required for keeping DSS-style learning from collapsing the representations and are functions of the clean f path only.

From the perspective of learning deep networks, it is important that any mapping, $f^{(l)}$ or $g^{(l)}$, is close to one of the cost function terms $C^{(l)}$. This means that learning is efficient even if propagating gradients through the mappings would not be efficient.

4 Experiments

This section presents a few simple experiments which demonstrate the key aspects of the ladder network:

- How denoising functions can represent probability distributions.
- How lateral connections relieve the pressure to represent every detail at the higher layers of the network and allow them to focus on abstract invariant features.

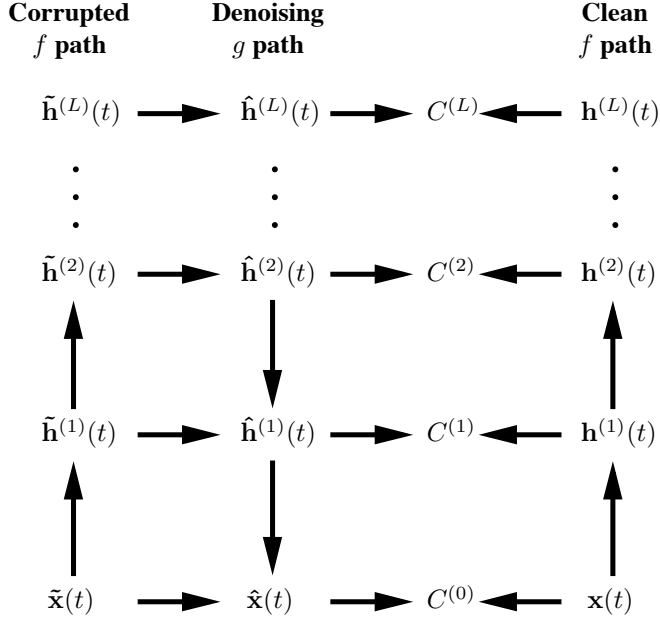


Figure 2: Ladder network’s cost computations are illustrated. The clean f path shares exactly the same mappings $f^{(l)}$ as the corrupted f path. The only difference is that corruption noise is added in the corrupted path. The resulting corrupted activations are denoted by $\tilde{\mathbf{h}}^{(l)}(t)$. On each layer, the cost function has a term $C^{(l)}$ which measures the distance between the clean activations $\mathbf{h}^{(l)}(t)$ and their reconstructions $\hat{\mathbf{h}}^{(l)}(t)$. The terms $C_{\Sigma}^{(l)}$ and $C_{\mu}^{(l)}$ which measure how well the activations $\hat{\mathbf{h}}^{(l)}(t)$ are normalized are not shown.

- How the cost function terms on higher layers speed up learning.

The three following sections will gradually develop a two-layered ladder network which can learn abstract invariant features. First, simple distributions are modeled. Then this is put into use in a linear ICA model with one hidden layer. Finally, a second layer is added which models the correlations between the variances of the first layer activations. All the experiments used the set of learning rules described in Eqs. (40–50). The hyperparameters β_l were automatically adjusted to keep the smallest eigenvalue of Σ_l above 0.7. The hyperparameter γ_l was set to the the same value as β_l .

4.1 Representing distributions with denoising functions

We will start by a simple experiment which elucidates the relation between the prior distribution of activations and their denoising functions, illustrated in Fig. 3. Three different distributions were tested, super-Gaussian, sub-Gaussian and Gaussian distribution. Each of them had a unit variance and zero mean. Each plot shows five different results overlaid on top of each other.

The super-Gaussian distribution was a Laplace distribution whose probability density function (pdf) is

$$p(x) = \frac{1}{\sqrt{2}} e^{-\sqrt{2}|x|}.$$

On a logarithmic scale, it behaves as $-\sqrt{2}|x|$ plus constant.

The sub-Gaussian distribution was generated by scaling a sinusoidal signal by $\sqrt{2}$ to obtain a variable with unit variance. The pdf of this distribution is

$$p(x) = \frac{1}{\pi\sqrt{2-x^2}}$$

for $|x| < \sqrt{2}$.

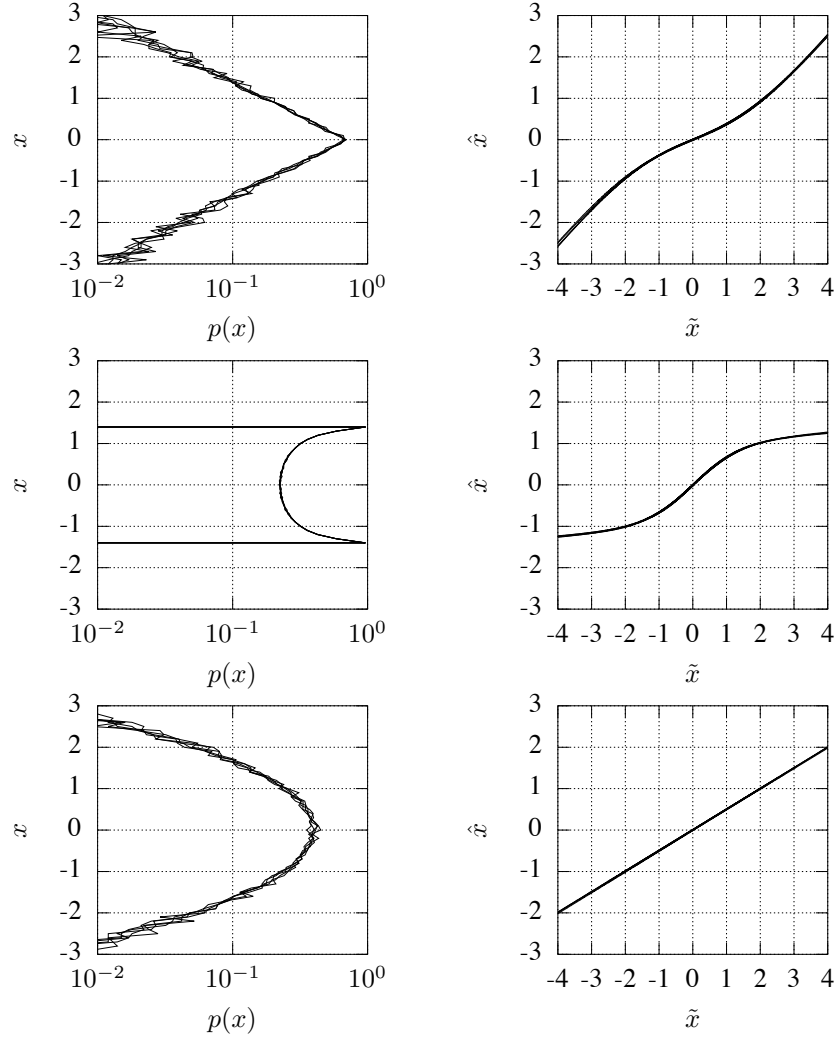


Figure 3: Illustration of the connection between the marginal distribution and denoising function. On the left, three different probability distributions are shown on a logarithmic scale. From top: super-Gaussian, sub-Gaussian and Gaussian distribution. Each distribution has a unit variance and zero mean. On the right, denoising functions have been trained to remove corruptive noise with unit variance. In the Gaussian case, theoretically the optimal solution is $\hat{x} = \tilde{x}/2$. Each plot shows five different random samples plotted on top of each other. Note that in the plots showing the pdf or the data, x is plotted on the vertical axis to help side-by-side comparison with the denoising function.

For this experiment, the model did not have any hidden layers ($L = 0$) and therefore there are no forward functions f , only one denoising function which was implemented as a single hidden neuron with tanh activation and a bypass connection:

$$\hat{x} = g(\tilde{x}) = \xi_1 \tilde{x} + \xi_2 \tanh(\xi_3 \tilde{x} + \xi_4) + \xi_5,$$

where ξ_i are scalar parameters. All parameters were optimized by minimizing $C^{(0)} = \sum_t \|\hat{x}(t) - x(t)\|^2$.

With small enough noise, the denoising functions should theoretically approach

$$\hat{x} \approx \tilde{x} + \sigma_n^2 \frac{\partial \log p(x)}{x} \Big|_{x=\tilde{x}},$$

where σ_n^2 is the variance of the noise used for corrupting \tilde{x} . With small corruption noise, this would then mean that the denoising function of the Laplacian input would be a sum of x and a scaled step function. This is not the case now since σ_n^2 is as large as the variance of the input. This tends to smoothen the denoising function.

As can be readily seen from Fig. 3, the denoising function for a Gaussian observation is linear. Theoretically, the function should be

$$\hat{x} = \frac{\sigma_x^2}{\sigma_x^2 + \sigma_n^2} \tilde{x},$$

where σ_x^2 is the variance of the observations. Since both variances equal one in these experiments, the theoretical optimum is $\hat{x} = \tilde{x}/2$. The estimated denoising function follows this very closely.

4.2 ICA model

We will now move on to a simple linear ICA model which serves as an example of how statistical modeling will be translated into function approximation in this framework. It also gives some intuition on how the lateral connections help higher levels to focus on relevant features.

In linear independent component analysis, the data is assumed to be a linear mixture of independent identically distributed sources. Unlike in principal component analysis, the mixing is not restricted to be orthogonal because sources with non-Gaussian marginal distributions can be recovered. If more than one source has a Gaussian distribution, these sources cannot be expected to be recovered and will remain mixed. However, the subspace spanned by the Gaussian sources should be recoverable and separate from all the non-Gaussian sources.

One limitation is that unless there is some extra information available, the sources can only be recovered up to scaling and permutation. Scaling is usually fixed by assuming that source distributions have a unit variance. This will still leave permutation and sign of the sources ambiguous.

The dataset was generated by linearly mixing 10,000 samples from 15 sources into 15 observations. The elements of the mixing matrix were sampled from a zero-mean Gaussian distribution. The sources had the same distributions as in the previous example: five super-Gaussian, five sub-Gaussian and five Gaussian sources. The data were *not* whitened because one purpose of the experiments was to demonstrate that normal autoencoders have a bias towards PCA solution even if the cost function terms $C^{(\geq 1)}$ are not used. In these experiments, α_l in Eq. (50) were set to zero.

4.2.1 Model structure

The model had one hidden layer ($L = 1$) and the only nonlinearity was on the hidden layer denoising which was a simplified version of the model used in the previous experiment. With essentially zero mean observations, there is no need for bias terms in the model. The mappings of the model are as follows:

$$f(\mathbf{x}) = \mathbf{W}\mathbf{x} \tag{51}$$

$$g_i^{(1)}(\mathbf{h}) = a_i h_i + b_i \tanh(h_i) \tag{52}$$

$$g^{(0)}(\mathbf{x}, \mathbf{h}) = \mathbf{A}\mathbf{h} + \mathbf{B}\mathbf{x} \tag{53}$$

The parameters to be estimated are the three matrices \mathbf{W} , \mathbf{A} and \mathbf{B} and the two vectors \mathbf{a} and \mathbf{b} which were used for the denoising functions of individual hidden units.

Note that the underlying assumption of ICA models is that the sources are independent. This is incorporated in the model by making the denoising on the hidden layer unit-wise. Also, the lateral linear mapping with matrix \mathbf{B} can model any covariance structure in the observation data. This means that the hidden layer should be able to focus on representing the non-Gaussian sources.

4.2.2 Results

Experiments verify that the model structure defined by Eqs. (51–53) is indeed able to recover the original sources that were used for generating the observed mixtures (up to permutation and scaling). This is apparent when studying the normalized loading matrix which measures how much the original sources contribute to the values of the hidden units. The loading matrix is obtained from the product $\mathbf{W}\mathbf{A}_{\text{orig}}$ where \mathbf{W} is the unmixing matrix learned by the model and \mathbf{A}_{orig} is the original mixing matrix. Rows of this matrix measure how much contribution from the original source there is in each of the recovered hidden neuron activation. In the normalized loading matrix, the rows are scaled so that the squares sum up to one, that is, the row vectors have unit lengths. A successful unmixing is characterized by a single dominant loading in each row and can be measured by the average angle of each vector to the dominant source. In the experiments, a typical value was around 10 which corresponds to the contribution $\cos(10) \approx 0.985$ from the dominant source.

The denoising mappings $g_i^{(1)}(\mathbf{h})$ for each source depend on the distribution of the source just as expected. In particular, the sign of the parameter b_i is determined by the super- or sub-Gaussianity of the sources. When there are more hidden units than non-Gaussian sources, the model will also represent Gaussian sources but the preference is for non-Gaussian sources. This is to be expected because the lateral mapping \mathbf{B} at the lowest level of the network can already represent any Gaussian structure. In other words, the model performs just as expected.

What is more interesting is what happens if the lateral mapping \mathbf{B} is missing. Since the reconstruction $\hat{\mathbf{x}}(t)$ can then only contain information which is present in the hidden units, the network has a strong pressure to conserve as much information as possible. Essentially the dominant mode of operation is then PCA: the network primarily extracts the subspace spanned by the eigenvectors of the data covariance matrix corresponding to the largest eigenvalues. The network can only secondarily align the representation along independent components. If the independent components do not happen to align with the principal subspace, PCA wins over ICA.

In one experiment, for instance, a network with \mathbf{B} and 11 hidden units was able to retrieve the ten non-Gaussian sources with loadings between 0.958 and 0.994, averaging 0.981. By contrast, the ten best loadings in exactly the same setting but without \mathbf{B} were between 0.499 and 0.824 and averaged 0.619 after the same number of iterations. This is not significantly different from random mappings which yield average loadings around 0.613 ± 0.026 (average \pm std).

It turned out that the network was able to do better but it converged tremendously slowly, requiring about 100 times more iterations than with \mathbf{B} . Still, even after the network had seemingly converged, the ten best loadings were between 0.645 and 0.956 and averaged 0.870. While this is clearly better than random, even the best loading was worse than the worst loading when the lateral connections \mathbf{B} were used.

That this is due to the network’s tendency to extract a principal subspace can be seen by analysing how large portion of the subspace spanned by \mathbf{W} falls outside the subspace spanned by the 11 largest eigenvectors of the data covariance matrix. In the network with \mathbf{B} this was 28 % whereas in the network lacking \mathbf{B} this was just 0.03 %.

To be fair, it should be noted that pre-whitening the inputs $\mathbf{x}(t)$ restores the autoencoder’s ability to recover independent components just as it allows nonlinear PCA learning rule to perform independent component analysis rather than principal subspace analysis. However, in more complex cases it may not be as easy to normalize away the information that is not wanted.

4.3 Hierarchical variance model

We shall now move on to expanding the ICA model by adding a new layer to capture the nonlinear dependencies remaining in the hidden unit activations in the ICA model. This makes sense because it is usually impossible to produce truly statistically independent components simply by computing different linear projections of the observations. Even if the resulting feature activations lack linear correlations, there are normally higher-order dependencies between the features (for more discussion of such models, see Valpola et al. 2004; Hyvärinen and Hoyer 2000).

One typical example is that the variances of the activations are correlated because the underlying cause of a feature activation is likely to generate other activations, too. In order to find a network structure that could represent such correlated variances, let us recall that the optimal denoising of a Gaussian variable h with prior variance σ_h^2 and Gaussian corruption noise σ_n^2 is

$$\hat{h} = \frac{\sigma_h^2}{\sigma_h^2 + \sigma_n^2} \tilde{h}.$$

This can be written as

$$\hat{h} = \frac{1}{1 + \sigma_n^2/\sigma_h^2} \tilde{h} = \text{sigmoid}(\log \sigma_h^2 - \log \sigma_n^2) \tilde{h},$$

where the sigmoidal activation function is defined as

$$\text{sigmoid}(x) = \frac{1}{1 + e^{-x}}.$$

What this means is that information about the variance of h translates to a modulation of the connection strength in the denoising function mapping \tilde{h} to \hat{h} .

The experiments in this section demonstrate that all we need to do to represent correlations in the variances in the hidden unit activations is add modulatory connections from the layer above and give enough flexibility for the forward mapping to the highest layer. The network will then find a way to use the representations of the highest layer. Moreover, lateral connections on the middle layer now play a crucial role because they allow the higher layers to focus on representing higher-order correlations. Unlike in the case with the ICA model in the previous section, this could not have been replaced by whitening the inputs. The experiments also demonstrate how the cost function terms on the higher layers speed up learning.

4.3.1 Data

The dataset of 10,000 samples was generated as a random linear mixture of sources s_i just as in the ICA experiment. This time, however, the sources were Gaussian with a changing variance. The variances of the sources were determined by higher-order variance sources, each of which was used by a group of four sources. There were four such groups, in other words, four higher-order variance sources which determined the variances of 16 sources. Such groups of dependent sources mean that the data follows the model used in independent subspace analysis (Hyvärinen and Hoyer 2000).

The variance sources v_j were sampled from a Gaussian distribution and the variance σ_i^2 of the lower-level sources $i \in \mathbb{G}_j$ was obtained by computing e^{v_j} . The set \mathbb{G}_j contain all the indices i for which the lower-level sources s_i are modulated by the variance source v_j . Since there were four non-overlapping groups of four sources, $\mathbb{G}_1 = \{1, 2, 3, 4\}$, $\mathbb{G}_2 = \{5, 6, 7, 8\}$, and so on.

Note that although the sources are sampled from a Gaussian distribution, their marginal distribution is super-Gaussian since the variance is changing. In these experiments, the dataset was pre-whitened.

4.3.2 Model structure

The linear mappings between the observations and the first layer were the same as with the ICA model but now a second nonlinear layer was added and denoising $g^{(1)}$ of the first layer was modified

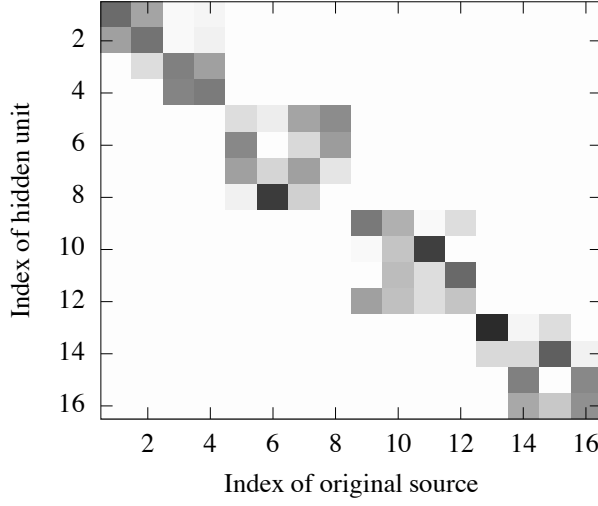


Figure 4: The squares of the normalized loading matrix from a hierarchical ladder network. Black corresponds to one and white to zero. As is evident from the plot, the four subspace \mathbb{G}_j have been cleanly separated from each other but remain internally mixed. This is because the distribution of the sources within each subspace is spherically symmetric which makes it impossible to determine the rotation within the subspace. This is reflected in the blocky structure of the matrix of loadings. The hidden units were ordered appropriately to reveal the block structure.

to make use of it:

$$f^{(1)}(\mathbf{x}) = \mathbf{W}^{(1)}\mathbf{x} \quad (54)$$

$$f^{(2)}(\mathbf{h}^{(1)}) = \text{MLP}(\mathbf{h}^{(1)}) \quad (55)$$

$$g_i^{(2)}(\mathbf{h}^{(2)}) = a_i h_i^{(2)} \quad (56)$$

$$g_i^{(1)}(\mathbf{h}^{(1)}, \mathbf{h}^{(2)}) = \text{sigmoid}(\mathbf{A}_i^{(2)}\mathbf{h}^{(2)} + b_i^{(1)})h_i^{(1)} \quad (57)$$

$$g^{(0)}(\mathbf{x}, \mathbf{h}^{(1)}) = \mathbf{A}^{(0)}\mathbf{h}^{(1)} + \mathbf{B}\mathbf{x}. \quad (58)$$

The multi-layer perceptron (MLP) network used in the second-layer encoder mapping $f^{(2)}$ was

$$\text{MLP}(\mathbf{h}) = \mathbf{W}^{(2b)}\psi(\mathbf{W}^{(2a)}\mathbf{h} + \mathbf{b}^{(2a)}) + \mathbf{b}^{(2b)}, \quad (59)$$

where the activation function $\psi(x) = \log(1 + e^x)$ operates on the elements of the vector separately. Note that we have included the bias term $\mathbf{b}^{(2b)}$ to make sure that the network can satisfy the constraint of having zero mean activations.

4.3.3 Results

Experiments verified that the network managed to separate individual source subspaces (Fig. 4) and learned to model the correlations between the variances of different sources (Fig. 5). The figures correspond to an experiment where the dimension of the first layer was 16 and second layer 10. The MLP network used for modeling $f^{(2)}$ had 50 hidden units. The figures show results after 1000 training for iterations. The results were relatively good already after 300 iterations and after 1000 iterations the network had practically converged.

Figure 4 shows squares of the loadings, that is, elementwise squares of the matrix $\mathbf{W}^{(1)}\mathbf{A}_{\text{orig}}$, scaled such that the squares sum up to one for each hidden unit. Since the data was generated by modulating the variance of the Gaussian sources of each subspace, the data distribution was spherically symmetric within each subspace. This is reflected in the blocky appearance of the the matrix of loadings as the network has settled on a random rotation within the subspace. Note that such rotation indeterminacy is a property of the input data. In other experiments the model was fed with the ICA data from the previous section and readily separated the sources (results not shown here).

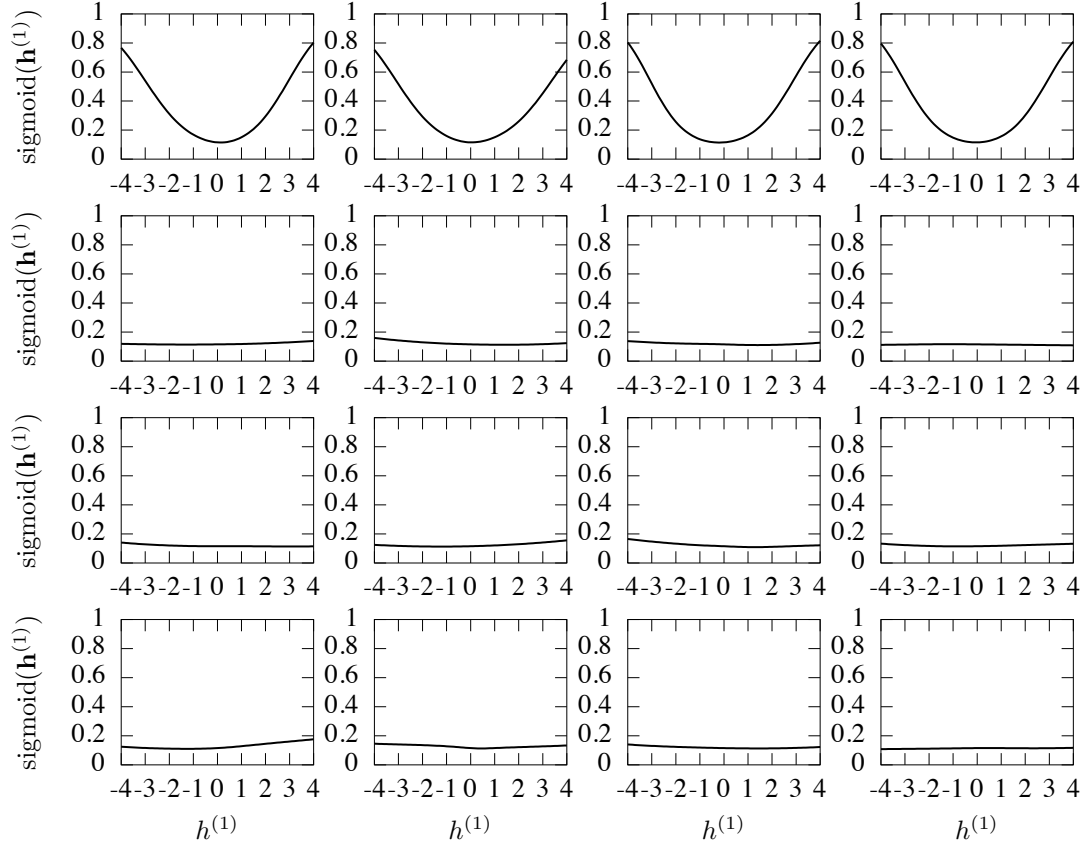


Figure 5: Illustration of one of the denoising functions learned by a hierarchical ladder network. Each plot shows how the term $\text{sigmoid}(\mathbf{A}_1^{(2)}\mathbf{h}^{(2)} + b_1^{(1)})$ modulating $\tilde{h}_1^{(1)}$ in Eq. (57) behaves as a function of one of the hidden neurons $h_i^{(1)}$. In each plot, $h_i^{(1)}$ for just one i takes nonzero values. The top row shows how sigmoid changes as a function of the hidden units which belong to the same subspace \mathbb{G}_1 as $h_1^{(1)}$ whose sigmoid is shown. The rest of the plots correspond to hidden units from other groups $\mathbb{G}_{\neq 1}$.

Figure 5 depicts the learned denoising functions. Each plot shows how the term $\text{sigmoid}(\mathbf{A}_1^{(2)}\mathbf{h}^{(2)} + b_1^{(1)})$ belonging to the first hidden unit in Eq. (57) behaves as a function of one of the hidden neurons $h_i^{(1)}$. The sigmoid term is a function of all the hidden neuron activations through the second layer hidden unit activations. Theoretically, the sigmoid term modulating the mapping from $\tilde{h}_1^{(1)}$ to $\hat{h}_1^{(1)}$ should be a function of the norm of the activation vector of subspace \mathbb{G}_1 , that is, a function of $\sum_{i \in \mathbb{G}_1} [h_i^{(1)}]^2$. As is readily seen from the plots, the second layer activations have apparently learned to represent such quadratic terms of first layer activations because the sigmoid indeed appears to be a function of the norm of the activation vector of the first subspace. When ICA data was used (results not shown here), the sigmoid term learned to neglect all the other hidden units and developed to be a function of $h_1^{(1)}$ only (assuming we are looking at the sigmoid of hidden unit 1).

The variance features developing on the second layer of the model can only improve the reconstruction of $\hat{\mathbf{x}}$ if they are combined with $\mathbf{h}^{(1)}$ because variance alone cannot say anything about the direction where the reconstruction should be changed. Without the shortcut connections in the ladder model, the highest layer with only ten hidden units⁵ could not have learned to represent the variance sources because there would not have been enough space to also represent the activations $\mathbf{h}^{(1)}$ which are also needed to make use of the variance sources. The higher layers can only let go of the details because they are recovered again when denoising proceeds from the highest layers towards the lowest.

Another important question was whether the terms of the cost function originating in the higher layers of the network really are useful for learning. To investigate this, 100 simulations were run with different datasets and random initializations of the network. It turned out that in this particular model, the higher layer cost function term $C^{(2)}$ was not important but $C^{(1)}$ could speed up learning considerably particularly during the early stages of learning. As expected, it was crucial to combine it with a proper decorrelation term $C_\Sigma^{(1)}$. The success of the model was measured by the value $C^{(0)}$ that was reached. Note that it is not *a priori* clear that adding other cost function terms could help reduce $C^{(0)}$. Nevertheless, this turned out to be the case. By iteration 100, the network consistently reached a lower value of $C^{(0)}$ than by iteration 200 when minimizing $C^{(0)}$ alone (as in standard denoising autoencoders). Subsequent learning continued approximately at the same pace which seems reasonable as denoising autoencoders should be able to optimize the model after it has been initialized close to a sensible solution.

Another interesting finding was that about one third of the improvement seems to be attributable to the decorrelation term $C_\Sigma^{(1)}$. It was able to speed up learning alone without $C^{(1)}$ despite initializing the network with mappings which ensure that all the representations start out as decorrelated. Whereas the speedup of $C^{(1)}$ was most pronounced in the beginning of learning, the speedup offered by $C_\Sigma^{(1)}$ was more important during the middle phases of learning. Presumably this is because the representations start diverging from the uncorrelated initialization gradually.

At the optimum of the cost function $C^{(0)}$, the addition of any extra term can only make the situation worse from the viewpoint of minimizing $C^{(0)}$. It is therefore likely that optimally the weights α_l and β_l in Eq. (50) should be gradually decreased throughout training and could be set to zero for a final finetuning phase. For simplicity, all α_l were kept fixed in the simulations presented here.

5 Discussion

The experiments verified that the ladder model with lateral shortcut connections and cost function terms at every level of the hierarchy is indeed able to learn abstract invariant features efficiently. Although the networks studied here only had a few layers (no more than six between $\tilde{\mathbf{x}}$ and $\hat{\mathbf{x}}$ no matter how they are counted), what is important is that the representations were abstract and invariant already on the second layer. In fact, the model with two layers, linear features on the first and variance features on the second, corresponds roughly to the architecture with simple and complex cells found by Hubel and Wiesel (1962) (for a more detailed discussion, see Hyvärinen and

⁵In principle the highest layer only needs four hidden units to represent the four variance sources. The network was indeed able to learn such a compact representation but learning tended to be slower than with ten hidden units.

Hoyer 2000). Promising as the results are, it is clearly necessary to conduct far larger experiments to verify that the ladder network really does support learning in much deeper hierarchies.

Similarly, it will be important to verify that the ladder network is indeed compatible with supervised learning and can therefore support useful semi-supervised learning. All the experiments reported in this chapter were unsupervised. However, all the results supported the notion that the shortcut connections of the ladder network allow it to discard information, making it a good fit with supervised learning.

The approach proposed here is not the only option for semi-supervised learning with autoencoders. Another alternative is to split the representation on the higher levels to features which are also used by supervised learning and to features which are only used by the decoder (Rifai et al. 2012). Recently Kingma et al. (2015) obtained good results with a model which combined this idea with hierarchical latent variable models. Their model included a trainable recognition mapping in addition to the generative model, so the overall structure resembles an autoencoder. Only additive interactions were included between the encoder and decoder and it remains to be seen whether their approach can be extended to include more complex interactions, such as the modulatory interaction in Eq. (57).

A somewhat similar approach, motivated by variational autoencoders and target propagation, was proposed by Bengio (2014). It also includes exchange of information between the decoder and encoder mappings on every level of the hierarchy, though only during learning.

One of the most appealing features of the approach taken here is that it replaces all probabilistic modeling with function approximation using a very simple cost function. In these experiments, an MLP network learned to extract higher-level sources that captured the dependencies in the first-level sources. This model corresponded to independent subspace analysis not because the model was tailored for it but because the input data had that structure. The forward mapping $f^{(2)}$ was very general, an MLP network. In these experiments the denoising mapping $g^{(1)}$ had a somewhat more limited structure mainly to simplify analysis of the results. It will be interesting to study whether $g^{(l)}$ can also be replaced by a more general mapping.

Another important avenue for research will be to take advantage of all the machinery developed for GSN and related methods, such as sampling, calculating probability densities and making use of multiple rounds of corruption and denoising during learning (Bengio et al. 2013; Uria et al. 2014; Raiko et al. 2014). Particularly the ability to sample from the model should be very useful. In order to make full use of these possibilities, the corruption procedure should be extended. Now simple Gaussian noise was added to the inputs. Noise could be added at every layer to better support sampling (Bengio et al. 2013) and could also involve masking out some elements of the input vector completely (Uria et al. 2014). If different types of corruption are needed at different times, it might be possible to extend the denoising functions to handle different types of corruption (information about the corruption strategy could be provided as side information to the denoising functions) or it might be possible to relearn just the denoising functions $g^{(l)}$ while keeping the previously learned forward mappings $f^{(l)}$ fixed.

A crucial aspect of the ladder network is that it captures the essential features of the inference structure of hierarchical latent variable models. Along the same line, it should be possible to extend the model to support even more complex inferences, such as those that take place in Kalman filters.

The model studied by Yli-Krekola (2007) takes even one step further: it implements a dynamical biasing process which gives rise to an emergent attention-like selection of information in a similar fashion as in the model suggested by Deco and Rolls (2004). The model studied by Yli-Krekola (2007) is derived from the DSS framework and already has a structure which is reminiscent of the ladder architecture presented here. The model is otherwise very elegant but is prone to overfit its lateral connections and exaggerate the feedback loops between units. By using the same tricks as here, injecting noise for the benefit of learning lateral and top-down denoising, it might be possible to learn the lateral connections reliably.

6 Conclusions

In this chapter, a ladder network structure was proposed for autoencoder networks. The network's lateral shortcut connections give each layer the same representational capacity as stochastic latent variables have in hierarchical latent variable models. This allows the higher levels of the network to discard information and focus on representing more abstract invariant features.

In order to support efficient unsupervised learning in deep ladder networks, a new type of cost function was proposed. The key aspect is that each layer of the network contributes its own terms to the cost function. This means that every mapping in the network receives training signals directly from some term which measures local reconstruction errors. In addition to the immediate training information, the network also propagates gradient information throughout the network. This means that it is also possible to add terms which correspond to supervised learning.

The price to pay is that each higher-level cost function needs to be matched with a decorrelation term which prevents the representation from collapsing. This is analogous to the competition used in unsupervised competitive learning. Additionally, it is often useful to decorrelate the inputs because otherwise the network is biased towards finding a PCA solution.

Preliminary experiments verified that the network was able to learn abstract invariant features and that the extra terms in the cost function speed up learning. The experiments support the notion that the model scales to very deep models and works well together with supervised learning but much larger experiments are still required to verify these claims.

Acknowledgments

I would like to thank Tapani Raiko, Antti Rasmus and Yoshua Bengio for useful discussions. Antti Rasmus has been running experiments in parallel to this work and his input on how different versions performed has been invaluable. Jürgen Schmidhuber, Kyunghyun Cho and Miquel Perelló Nieto have made available collections of citations which has saved plenty of my time when preparing the manuscript.

Last but certainly not least, I would like to thank Erkki Oja for creating the environment where the ideas which underly the work presented here have been able to develop. Erkki has always supported my research. His example has shown how it is possible to follow intuition in designing unsupervised learning algorithms but he has also always emphasized the importance of rigorous analysis of the convergence and other properties of the resulting algorithms. Without this combination and his pioneering work in neural PCA, nonlinear PCA learning rule and ICA, none of the research reported here would have gotten very far.

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