# Analyzing $(1, \lambda)$ Evolution Strategy via Stochastic Approximation Methods

G. Yin<sup>\*</sup>, G. Rudolph<sup>†</sup> and H.-P. Schwefel<sup>‡</sup>

#### Abstract

The main objective of this paper is to analyze the  $(1, \lambda)$  evolution strategy by use of stochastic approximation methods. Both constant and decreasing step size algorithms are studied. Convergence and estimation error bounds for the  $(1, \lambda)$  evolution strategy are developed. First, the algorithm is converted to a recursively defined scheme of stochastic approximation type. Then the analysis is carried out by using the analytic tools from stochastic approximation. In lieu of examining the discrete iterates, suitably scaled sequences are defined. These interpolated sequences are then studied in detail. It is shown that the limits of the sequences have natural connection to certain continuous time dynamical systems.

**Key words:** evolutionary computation, evolution strategy, stochastic approximation, convergence, rate of convergence.

Abbreviated title. Analyzing  $(1, \lambda)$  Evolution Strategy

<sup>\*</sup>Department of Mathematics, Wayne State University, Detroit, MI 48202. The research of this author was supported in part by the National Science Foundation under grant DMS-9224372, and in part by the Deutscher Akademischer Austauschdienst (DAAD).

<sup>&</sup>lt;sup>†</sup>Informatik Centrum Dortmund, Joseph-von-Fraunhofer-Str. 20, D-44227, Dortmund, Germany. The research of this author was supported by BMFT under grant 01 IB 403A, project EVOALG.

<sup>&</sup>lt;sup>‡</sup>Universität Dortmund, FB Informatik 11, D-44221 Dortmund, Germany.

# 1 Introduction

In this work, we develop asymptotic properties, in particular, convergence and upper bounds of estimation errors for the  $(1, \lambda)$  evolution strategy. Both constant and decreasing step size algorithms are considered. We treat the problem as a recursive algorithm of stochastic approximation type, and use the methods in stochastic approximation to analyze the algorithms and obtain weak convergence and with probability one (w.p.1) convergence results.

Based upon collective processes with a population of individuals, which are search points for a given problem, the evolutionary algorithms carry out desired computing tasks by use of randomized selection, mutation and recombination. These algorithms have been applied to many problems in parameter optimization and related fields with great success. Significant progress has been made in the study of evolutionary algorithms for almost thirty years. Many interesting and useful results have been obtained. To mention just a few, we cite the work of Rechenberg (1973), Schwefel (1965), Fogel, Owens, and Walsh (1966), Fogel (1992), Holland (1962), De Jong (1975) among others. For an extensive review of the recent advances, the readers are referred to Bäck and Schwefel (1993), Bäck, Rudolph, and Schwefel (1993), Schwefel (1993) and the references therein.

The evolution strategies were first developed by Rechenberg and Schwefel in the mid-60's (Rechenberg 1965; Schwefel 1965). At that time, applications in hydrodynamics such as optimizing the shape of a bent pipe and a flashing nozzle were dealt with. Different versions of the strategy were simulated (Schwefel 1965). The research in this subject has become a rapidly growing one ever since. Nowadays, the  $(\mu, \lambda)$  evolution strategies, introduced by Schwefel (1977), Schwefel (1981) are the state-of-the-art in evolution strategy research.

By examining the evolutionary algorithms (EAs) closely, there appears to be a natural connection between EAs and stochastic approximation. However, such a connection has not been explored until very recently (Yin, Rudolph, and Schwefel 1995). In the aforementioned paper, the authors dealt with the 'connection' question in a general setting, whereas in the current paper, we are aiming at deriving asymptotic properties for a class of evolutionary algorithms.

Such a study is important. First it will enhance our understanding on the intrinsic properties of the  $(1, \lambda)$  strategy, which in turn will lead to further improvement of the computation procedures. In addition, by formulating the problem as a stochastic approximation

algorithm, many analytical tools can be employed to carry out the theoretical investigation.

In Yin, Rudolph, and Schwefel (1995), the hidden step size of the  $(1, \lambda)$  strategy was discussed in an example. Here we take a closer scrutiny, and try to understand the basic properties of the scale parameter in the randomized sequence. It should be pointed out that EAs and stochastic approximation do have distinct features. The objective function  $f(\cdot)$  under consideration in EAs is available through simulation, whereas the corresponding counter part in stochastic approximation is only observable and available in the form of noisy measurements. Nevertheless, such a difference should not prevent us with employing stochastic approximation methods to analyze EA procedures.

Our plan is as follows. We formulate the problem and then convert it into a stochastic approximation algorithm in the next section. Section 3 deals with weak convergence issues. We show how the discrete iteration is related to a continuous time dynamical system. By taking appropriate interpolation, it is shown that an interpolated process converges weakly to a solution of an ordinary differential equation. Then, we proceed with obtaining upper bounds for the estimation errors in Section 4. This step is carried out via the use of the Liapunov function approach and stability of the dynamic system. Utilizing the upper bound as a bridge, we seek further development on a suitably scaled sequence for small a and large n. In the process of getting the asymptotic properties for the constant step size algorithms, our main technique is the weak convergence methods developed by Kushner (see Kushner 1984 and the references therein). Section 5 contains the analysis of algorithms with decreasing step size. W.p.1 convergence is derived by means of the ordinary differential equation (ODE) method. Some concluding remarks are issued in Section 6. Finally, an appendix containing the proof of a lemma is provided.

# 2 Problem setup

#### **2.1** $(1, \lambda)$ Evolution strategy

We wish to minimize a function  $f : \mathbb{R}^d \to \mathbb{R}$ . The plan is to employ the  $(1, \lambda)$  evolution strategy, for  $\lambda \geq 2$ . Loosely, the strategy can be described as follows. In each generation, one parent produces  $\lambda$  offspring. Among the offspring, choose the best one (with respect to the evaluation of the objective function) to form the next estimate.

To be more specific, generate sequences of random vectors  $\{z_n(i)\}$ , for  $1 \leq i \leq \lambda$  that

are independent and identically distributed (i.i.d.) normal random variables with mean zero and covariance  $\sigma^2 I_d$ , where  $I_d$  denotes the  $d \times d$  identity matrix such that for each  $n, z_n(1), \ldots, z_n(\lambda)$  are independent. To carry out the minimization task, choose an initial estimate  $x_0 \in \mathbb{R}^d$ . At iteration n, add the random vector  $z_n(i)$  to the current content, i.e.,  $x_n + z_n(i)$ , for  $i = 1, \ldots, \lambda$ . We evaluate the corresponding values  $f(x_n + z_n(i))$ . Next, choose the smallest among the  $\lambda$  values of  $f(\cdot)$ . That is,

$$f(x_n + z_n(j)) = \min_{\mu \in \Lambda_n} f(x_n + \mu), \text{ where } \Lambda_n = \{z_n(i), i = 1, \dots, \lambda\}.$$
 (1)

Then assign  $x_n + z_n(j)$  to  $x_{n+1}$ . In short

$$x_{n+1} = \operatorname{argmin}\{f(x_n + z_n(1)), \dots, f(x_n + z_n(\lambda))\}.$$
(2)

This problem was studied in Rudolph (1994) by using martingale convergence theorem. Here we take a different approach. Our task now is to convert (2) to a recursive algorithm of stochastic approximation type so that the techniques in analyzing stochastic approximation type of algorithms can be applied.

Remark: In Eq. (1) above, without loss of generality, we have assumed that there is only one j satisfying (1). If there are more than one indices satisfying (1), choose j to be the smallest one among them, i.e.,

$$j = \min\{1 \le l \le \lambda; f(x_n + z_n(l)) = \min_{\mu \in \Lambda_n} f(x_n + \mu)\}.$$

### 2.2 A recursive algorithm

It is well known that the standard deviation  $\sigma$  is a scale factor in the problem. Since  $z_n(i)$  are i.i.d. random vectors and  $z_n(i) \sim N(0, \sigma I_d)$ , we can re-scale the sequence  $z_n(i)$  or equivalently, define another sequence  $\{\tilde{z}_n(i)\}$  by setting  $z_n(i) = \sigma \tilde{z}_n(i)$  such that  $\tilde{z}_n(i) \sim N(0, I_d)$ . That is,  $\tilde{z}_n(i)$  follows the standard normal distribution. Now (2) can be rewritten as

$$x_{n+1} = x_n + \sigma \sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x_n + z_n(i) = \min_{\mu \in \Lambda_n} f(x_n + \mu)\}},$$
(3)

where  $I_{\cdot}$  is an indicator function.

Again, in the equation above, we have assumed that there is only one i satisfying (1). For the case of multiple indices leading to the minimal, choose i to be smallest of them and as a result

$$x_{n+1} = x_n + \sigma \sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x_n + z_n(i) = \min_{\mu \in \Lambda_n} f(x_n + \mu); i = \min_{1 \le l \le \lambda} f(x_n + z_n(l)) = \min_{\mu \in \Lambda_n} f(x_n + \mu)\}}.$$

For notational simplicity reason, we shall use (3) throughout. As mentioned before, there is no loss in generality to do so.

In evolution strategy, one often chooses  $\sigma$  so that it is proportional to  $(1/d)H(f_x(x_n))$ , where  $f_x(\cdot)$  denotes the gradient of  $f(\cdot)$ , d is the dimension of the problem and  $H(\cdot) : \mathbb{R}^d \mapsto [0, \infty)$  is an appropriate real-valued function such that H(0) = 0 and the only root of  $H(\cdot)$  is 0. With a denoting the proportional constant multiplied by 1/d, the recursive formula can be written as

$$x_{n+1} = x_n + aH(f_x(x_n)) \sum_{i=1}^{\lambda} z_n(i) I_{\{f(x_n + z_n(i)) = \min_{\mu \in \Lambda_n} f(x_n + \mu)\}}.$$
(4)

Eq. (4) in fact, is a constant step size stochastic approximation algorithm with step size a. Since normally the problems we treat are large dimensional ones, a is relatively small. Our interest lies in obtaining convergence and rate of convergence results for the limit case  $a \rightarrow 0$ . We wish to emphasize the following point. In the actual computation, we neither change the evolution algorithm nor modify it in any way. The equivalent expression (4) is simply a convenient form that allows us to analyze the algorithm by using methods of stochastic approximation.

#### 2.3 An assumption

In the sequel, K denotes a generic positive constant. By convention, K + K = K and KK = K. We make the following assumption throughout of the paper.

(A) The function  $f(\cdot)$  is convex and is twice continuously differentiable such that  $f_{xx}(\cdot)$ is bounded, i.e., for all  $x \in \mathbb{R}^d$ ,  $|f_{xx}(x)| \leq K$ , where  $f_{xx}(\cdot)$  denotes the second derivative (or Hessian) of  $f(\cdot)$ .  $\{\tilde{z}_n(1)\}, \ldots, \{\tilde{z}_n(\lambda)\}$  are sequences of i.i.d. normal random vectors such that  $\tilde{z}_n(i) \sim N(0, I_d)$  for  $i = 1, \ldots, \lambda$ .

Remark: Since we can generate  $\{\tilde{z}_n(i)\}\$ , they are at our disposal so the i.i.d. condition is not a restriction. From the basic assumption, we know that the components of  $\tilde{z}_n(i)$  denoted by  $\tilde{z}_{n,j}(i)$  for  $j = 1, 2, \ldots, d$ , are N(0, 1) and  $\tilde{z}_{n,1}(i), \ldots, \tilde{z}_{n,d}(\lambda)$  are also independent. Since  $f_{xx}(\cdot)$  is continuous, it is bounded on bounded sets. Here we require a slightly stronger condition. When we carry out the EA computation, we are normally interested in the solution on bounded sets only. Furthermore, it is possible to design algorithms with projections and/or truncations so that the boundedness of the iterates are fulfilled.

### **3** Convergence

Under very natural conditions, we derive the convergence theorem and relate the discrete iteration to a continuous time ordinary differential equation. Note that the iterates  $x_n$  in (4) should really have been written as  $x_n^a$ . We have suppressed the *a*-dependence to keep the notation simple. In the sequel, if it is necessary we may retain it as needed.

We recall the definition of weak convergence first. A sequence of random variables  $\{w_n\}$  is said to converge to w weakly, if for any bounded and continuous function  $g(\cdot)$ ,  $Eg(w_n) \to Eg(w)$  as  $n \to \infty$ . Weak convergence is a substantial generalization of the notion of convergence in distribution. It implies much more than the simple convergence of multidimensional distributions since the function  $g(\cdot)$  can be chosen in various ways. Note that  $g(\cdot)$  is an arbitrary bounded and continuous function, and is not related to the objective function  $f(\cdot)$  in any ways. The concept of weak convergence can be employed not only to random variables living in an Euclidean space, but also random processes taking values in function spaces as well.

In the process of getting weak convergence result, one often needs to verify that the sequence involved is tight. A sequence  $\{w_n\}$  of  $\mathbb{R}^d$ -valued process is tight, if for any  $\varepsilon > 0$ , there is a compact set  $S_{\varepsilon}$  in  $\mathbb{R}^d$ , such that  $P(w_n \notin S_{\varepsilon}) \leq \varepsilon$  for all n. The definition of tightness carries over to the more general metric space valued sequences. A well-known theorem due to Prohorov (see Ethier and Kurtz 1986) states that, in a complete separable metric space, tightness is equivalent to sequential compactness. In other words, once the tightness is verified, one may proceed to extract convergent subsequences.

There are reasons that weak convergence analysis is more preferable in many applications. First, it requires much weaker conditions than its with probability one convergence counter part. Second, dealing with the problem of rates of convergence, we often need to obtain weak limit results. Therefore, one is forced to treat convergence in distribution or convergence in the weak sense any way. Third, to analyze a constant step size algorithm, we need to use weak convergence tools since if a constant step size is used, almost sure (w.p.1) convergence results cannot generally be expected. In addition, the constant step size algorithms are known to have the ability of tracking small parameter variations and are rather robust with respect to the noise processes.

For technical purposes, it is easier to deal with paths than measures. A device known as Skorokhod representation allows one to 'change' the weak convergence to w.p.1 convergence on a larger space. For the detailed account on the concept of weak convergence as well as many related materials, we refer the readers to the book Ethier and Kurtz (1986) and the references therein.

In our weak convergence analysis to follow, we often work with  $D^d[0,\infty)$ , which is the space of functions, that are right continuous, have left-hand limit endowed with certain weak topology (Skorokhod topology). Our analysis requires that first the tightness of the underlying processes be verified and then the limit process be characterized.

To proceed, with step size parameter a, we define a process  $x^{a}(\cdot)$  by a piecewise constant interpolation as follows:

$$x^{a}(t) = x_{n}$$
 for  $t \in [na, na + a)$ .

Thus, in lieu of examining the discrete iterates, we treat the process in continuous time, which gives us a better description on the dynamic behavior of the system involved.

In what follows, we apply the direct averaging methods (see Kushner 1984, Chapter 5) to study the process  $x^{a}(\cdot)$ . Notice that due to the distinct features of the evolutionary algorithms, the argument in Kushner (1984) needs to be modified for our needs.

In the weak convergence approach, normally, one needs to have an averaging condition of law of large numbers type. Such a condition now holds for our case based on the basic assumption (A). The essence of the direct averaging approach is to treat the variable x as fixed, and only average out the noise processes. Keeping this in mind, we first derive a preparatory result below.

**Theorem 3.1.** Let  $E_m$  denote the conditional expectation with respect to the  $\sigma$ -algebra  $\mathcal{F}_m$ , generated by  $\{x_0, \tilde{z}_j(i), j < m, i = 1, 2, ..., \lambda\}$ . Under Condition (A),

1. for each x, denote

$$E\sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x+\tilde{z}_n(i))=\min_{\mu\in\Lambda_n}f(x+\mu)\}} = \zeta(x).$$

Then for any  $n \geq m$ ,

$$E_m \sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x+\tilde{z}_n(i))=\min_{\mu \in \Lambda_n} f(x+\mu)\}} = \zeta(x);$$
(5)

 For each x ≠ x<sub>min</sub>, (that is, f(x<sub>min</sub>) = min<sub>x</sub> f(x)), ζ(x) is in the descent or 'downhill' direction, i.e., f'<sub>x</sub>(x)ζ(x) < 0. Moreover, there exists a function Ĥ(·) such that for all r with 0 < r < Ĥ(f<sub>x</sub>(x)), we have f(x + rζ(x)) < f(x).</li>

Remark: We notice that  $\zeta(\cdot)$  is a function of x. It also implicitly depends on the function  $f(\cdot)$ . The second assertion indicates that the algorithm is stable (we will be more precise about this in the sequel). In view of the first assertion in the above theorem, for each m, each n, and each x,

$$\frac{1}{n} \sum_{k=m}^{n+m-1} E_m \sum_{i=1}^{\lambda} \tilde{z}_k(i) I_{\{f(x+\tilde{z}_j(i))=\min_{\mu \in \Lambda_k} f(x+\mu)\}} = \zeta(x),$$

which will be needed in the sequel. Thus a law of large numbers type of condition holds for the underlying sequence.

Proof of Theorem 3.1. The verification of the first assertion above is almost obvious. We only note that although  $\sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x_n+z_n(i))=\min_{\mu\in\Lambda_n}f(x_n+\mu)\}}$  is used, at any given instance, only one of the random vectors is non-zero. For any  $n \ge m$ , since  $\tilde{z}_n(i)$  is independent of  $\mathcal{F}_m$ ,

$$E_m \sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x+\tilde{z}_j(i))=\min_{\mu \in \Lambda_n} f(x+\mu)\}} = E \sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x+\tilde{z}_j(i))=\min_{\mu \in \Lambda} f(x+\mu)\}} = \zeta(x).$$

As for the second part of the theorem, for simplicity, we suppress the subscript in  $z_n$  and write it as z instead in what follows. We note that the  $(1, \lambda)$ -ES generates  $\lambda$  i.i.d. Gaussian random variables, which can be decomposed as  $z \stackrel{D}{=} ru$ , where u denotes a random vector distributed uniformly on the surface of a unit hypersphere and r a nonnegative random variable stochastically independent from u.

Let  $x_n$  be the current position with  $x_n \neq x_{\min}$ . Then the algorithm compares the values

$$f(x_n + z(1)), f(x_n + z(2)), \dots, f(x_n + z(\lambda))$$

The new position  $x_{n+1}$  is the argument  $x_n + z(b)$ ,  $b \in \{1, \ldots, \lambda\}$ , that offers the lowest objective function value. Owing to Condition (A), for L > 0

$$f(x+h) \le f(x) + f'_x(x)h + Lh'h,$$
 (6)

the algorithm compares the values f(x) - f(x + z(i)), which can be bounded by

$$f(x) - f(x+z) \ge -f'_x(x) \, z - Lz'z = -rf'_x(x) \, u - r^2 L \tag{7}$$

with the usage of inequality (6). If we can show that the expectation of the maximum of the right-hand side of (7) within  $\lambda$  trials is larger than zero, then it is clear that the expectation of the selected step z(b) is a step of descent.

Assume that r is a constant for the moment, inequality (7) may be used to formulate the condition

$$\max_{i} \{ -f'_x(x) u_i \} \ge rL ,$$

where  $f'_x(x)$  denotes the transpose of  $f_x(x)$ . This leads to

$$\max_{i}\{|f_x(x)|\cos\omega_i\} \ge rL ,$$

where we used the relation

$$\frac{-f'_x(x)\,u_i}{|f_x(x)|} = \cos\,\omega_i\;.$$

Here,  $\omega_i$  denotes the random angle between the negative gradient and the random direction  $u_i$ . Finally, taking expectations we obtain

$$M_{\lambda} := E[\max_{i} \{\cos \omega_{i}\}] \ge rac{rL}{|f_{x}(x)|}$$

so that

$$0 \le H(f_x(x)) := r \le \frac{M_\lambda}{L} |f_x(x)| := \hat{H}(f_x(x)) .$$
(8)

To proceed, we state a lemma. Its proof is contained in the appendix.

**Lemma 3.2.** The random variable  $c = \cos \omega$  possesses a Beta distribution with parameters p = q = d/2 - 1, and distribution function

$$F_c(x) = \frac{1}{B((d-1)/2, (d-1)/2)} \int_0^{(x+1)/2} y^{(d-3)/2} (1-y)^{(d-3)/2} \, dy \tag{9}$$

where d denotes the dimensionality of random vector u.

Now come back to the proof of Theorem 3.1. Using Lemma 3.2, it remains to show that  $M_{\lambda} > 0$ . Note that  $M_{\lambda} = 2 \cdot E[\max\{B_i : i = 1, ..., \lambda\}] - 1$ , where the random variables  $B_i$  possess Beta distribution with parameters as in (9) but with support (0, 1). Let  $m_{\lambda} = E[\max\{B_i : i = 1, ..., \lambda\}]$ . Since the probability density function of  $B_i$  is unimodal and symmetric with respect to 1/2, we may use the inequality (see David 1970, p. 63)

$$F(m_{\lambda}) \ge \frac{\lambda}{\lambda+1} > \frac{1}{2} \ (\lambda \ge 2)$$

where  $F(\cdot)$  denotes the distribution function of  $B_i$ . Since  $F(\cdot)$  is convex-concave, the inverse  $F^{-1}(\cdot)$  is concave-convex, so that  $m_{\lambda} > F^{-1}(1/2)$ . Moreover, F(1/2) = 1/2 due to symmetry. Thus,  $m_{\lambda} > 1/2$  for  $\lambda \ge 2$  and  $M_{\lambda} = 2 m_{\lambda} - 1 > 0$ , and the proof of Theorem 3.1 is completed.  $\Box$ 

Remark: The second part of Theorem 3.1 was shown for the case that r is a deterministic choice, whereas  $z \stackrel{D}{=} ru$  was assumed to be a normal random vector. We argue that this result reflects the situation in large parameter spaces (d >> 1). Since  $r \sim \chi_n(\sigma)$ , we have  $E(r) \approx \sigma \sqrt{d-1/2}$ , and  $Var(r) \approx \sigma^2/2$ . Therefore, the realized step size r has only small variation so that we can regard  $r = \sigma \sqrt{d-1/2} \approx \sigma \sqrt{d}$  as a good approximation. Then relation (8) becomes

$$\sigma < \frac{M_{\lambda}}{L\sqrt{d}} |f_x(x)|.$$

Now we are in a position to present the weak convergence theorem which links the discrete iteration with a continuous time dynamic system.

**Theorem 3.3.** Suppose that Condition (A) holds and  $H(\cdot)$  and  $\zeta(\cdot)$  are continuous. Then  $\{x^a(\cdot)\}$  is tight in  $D^d[0,\infty)$ , such that as  $a \to 0$ , every weakly convergent subsequence has a limit  $x(\cdot)$  satisfying the following differential equation

$$\dot{x} = \frac{dx}{dt} = H(f_x(x))\zeta(x), \quad x(0) = x_0,$$
(10)

provided that the equation has a unique solution for each initial condition  $x_0$ .

Remark. Before going through the proof, let us make the following remarks. The conditions indicates that  $H : \mathbb{R}^d \to \mathbb{R}$  is a continuous function. Therefore, the composite function  $H(f_x(\cdot))$  is continuous. We also require  $\zeta : \mathbb{R}^d \to \mathbb{R}^d$  be continuous.

In the theorem above, we assumed that the initial condition  $x_0$  is the same as that of the discrete iteration and does not depend on a. More complex situations, i.e.,  $x_0 = x_0^a$  can be treated. The analysis is about the same. The only difference is that we have to add another condition  $x_0^a$  converges weakly to  $x_0$  as  $a \to 0$ . The continuity of  $\zeta(x)$  is not a restriction. Taking expectation is a smoothing process. Even indicator functions after such a process become continuous.

Proof of Theorem 3.3. For clarity, we divide the proof into several steps.

Step 1: (The use of N-truncation). Since it is not known *a priori*, whether the iterates  $\{x_n\}$  are bounded, we utilize the N-truncation technique (see Kushner 1984, p. 43). For

each  $N < \infty$ , define  $S_N = \{x; |x| \le N\}$ , that is the ball with radius N. Recall that  $x^{a,N}(t)$  is an N-truncation of  $x^a(t)$  if  $x^{a,N}(t) = x^a(t)$  up until first exit from  $S_N$ , and

$$\lim_{A \to \infty} \limsup_{a \to 0} P\left\{ \sup_{t \le T} |x^{a,N}(t)| \ge A \right\} = 0 \text{ for each } T < \infty.$$
(11)

For the discrete algorithm, we use

$$x_{n+1}^{N} = x_{n}^{N} + aH(f_{x}(x_{n}^{N}))\sum_{i=1}^{\lambda} \tilde{z}_{n}^{(i)}I_{\{f(x_{n}^{N}+z_{n}^{(i)})=\min_{\mu\in\Lambda_{n}}f(x^{N}+\mu)\}}q_{N}(x_{n}^{N}),$$
(12)

where  $q_N(\cdot)$  is a truncation function taking the form

$$q_N(x) = \begin{cases} 1, & x \in S_N; \\ 0, & x \in \mathbb{R}^d - S_{N+1}; \\ \text{smooth, otherwise.} \end{cases}$$

Step 2: (Tightness of the truncated process  $\{x^{a,N}(\cdot)\}$ ). By the definition (12),  $\{x_n^N\}$  is bounded. Since  $H(\cdot)$  is a continuous function and  $f(\cdot)$  is  $C^2$ ,  $H(f_x(x_n^N))$  is bounded. As a result,

$$E \left| H(f_{x}(x_{n}^{N})) \sum_{i=1}^{\lambda} \tilde{z}_{n}(i) I_{\{f(x_{n}^{N}+z_{n}(i))=\min_{\mu \in \Lambda_{n}} f(x_{n}^{N}+\mu)\}} \right|^{2}$$

$$\leq KE \sum_{i=1}^{\lambda} |\tilde{z}_{n}(i)|^{2} I_{\{f(x_{n}^{N}+z_{n}(i))=\min_{\mu \in \Lambda_{n}} f(x_{n}^{N}+\mu)\}}$$

$$\leq K \sum_{i=1}^{\lambda} E \left| \tilde{z}_{n}^{(i)} \right|^{2} < \infty.$$

It follows that

$$\left\{H(f_x(x_n^N))\sum_{i=1}^{\lambda}\tilde{z}_n(i)I_{\{f(x_n^N+z_n(i))=\min_{\mu\in\Lambda_n}f(x+\mu)\}}\right\}$$

is uniformly integrable. Lemma 7 in Chapter 3 of Kushner (1984) then yields that  $\{x^{a,N}(\cdot)\}$ is tight and the limit of any convergent subsequence has continuous paths w.p.1. Now extract a convergent subsequence. For notational simplicity, still use a as the index of the subsequence and denote the limit by  $x^{N}(\cdot)$ . By using the Skorokhod representation, without loss of generality, we may assume that  $x^{a,N}(\cdot)$  converges to  $x^{N}(\cdot)$  w.p.1, and the convergence is uniform on any bounded interval. Our next task is to characterize the limit process.

Step 3: (Characterization of the limit process  $x^{N}(\cdot)$ ). Our objective here is to show that the limit process  $x^{N}(\cdot)$  satisfies a truncated version of the equation (10). Introduce the notion

$$M^{N}(t) = x^{N}(t) - x^{N}(0) - \int_{0}^{t} H(f_{x}(x^{N}(\tau))\zeta(x^{N}(\tau))q_{N}(x^{N}(\tau))d\tau.$$

It is easily seen that  $M^N(\cdot)$  is Lipschitz continuous. If it is a martingale, then it must be a constant (see Kushner 1984). However,  $M^N(0) = 0$ . As a result,  $M^N(t)$  must be 0 identically, or equivalently,  $x^N(\cdot)$  is a solution of the ordinary differential equation (10).

Thus the problem reduces to verify the martingale property of  $M^N(\cdot)$ . To prove this, we need only show that for any bounded and continuous function  $h(\cdot)$ , any integer  $\nu$  and  $j \leq \nu$ , with  $t_j \leq t < t + s$ ,

$$Eh(x^{N}(t_{j}), j \leq \nu)[M^{N}(t+s) - M^{N}(t)] = 0.$$

We begin with the process  $x^{a,N}(\cdot)$ . In what follows, (t+s)/a, t/a etc. are all meant to be integers for notational convenience (if they are not integers, we can always take the integral parts anyway).

By using the interpolation,

$$\lim_{a} Eh(x^{a,N}(t_{j}), \ j \leq \nu) [x^{a,N}(t+s) - x^{a}(t)] \\
= \lim_{a} Eh(x^{a,N}(t_{j}), \ j \leq \nu) [x^{N}_{(t+s)/a} - x^{N}_{t/a}] \\
= \lim_{a} Eh(x^{a,N}(t_{j}), \ j \leq \nu) \left( a \sum_{k=t/a}^{(t+s)/a} H(f_{x}(x^{N}_{k})) + \sum_{k=1}^{\lambda} \tilde{z}_{k}(i) I_{\{f(x^{N}_{k}+z_{k}(i))=\min_{\mu \in \Lambda_{k}} f((x^{N}_{k}+\mu))\}} q_{N}(x^{N}_{k}) \right).$$
(13)

Now choose a sequence of integers  $\{n_a\}$  satisfying  $n_a \to \infty$  as  $a \to 0$  but  $\delta_a = an_a \to 0$ as  $a \to 0$ . Subdivide the interval [t/a, (t+s)/a) into intervals with length  $n_a$ . The term on the right side of the last equality sign of Eq. (13) then can be rewritten as:

$$\begin{split} &\lim_{a} Eh(x^{a,N}(t_{j}), \ j \leq \nu) \\ & \times \left( \sum_{k=t/a}^{(t+s)/a} aH(f_{x}(x_{k}^{N})) \sum_{i=1}^{\lambda} \tilde{z}_{n}(i) I_{\{f(x^{N}+z_{n}(i))=\min_{\mu \in \Lambda_{k}} f(x_{k}^{N}+\mu)\}} q_{N}(x_{k}^{N}) \right) \\ &= \lim_{a} Eh(x^{a,N}(t_{j}), \ j \leq \nu) \\ & \times \left( \sum_{l\delta_{a}=t}^{(t+s)} \delta_{a} \frac{1}{n_{a}} \sum_{k \in L_{a}} E_{ln_{a}} H(f_{x}(x_{k}^{N})) \sum_{i=1}^{\lambda} \tilde{z}_{k}(i) I_{\{f(x^{N}+z_{n}(i))=\min_{\mu \in \Lambda_{k}} f(x_{k}^{N}+\mu)\}} q_{N}(x_{k}^{N}) \right), \end{split}$$

where  $L_a$  denotes the interval  $ln_a \leq k \leq ln_a + n_a - 1$ , and  $E_{ln_a}$  denotes the conditioning on  $\mathcal{F}_{ln_a}$ , the  $\sigma$ -algebra generated by  $\{x_0, z_k(i), k < ln_a, i = 1, 2, ..., \lambda\}$ . Since  $x^N(t_j)$  for  $j \leq \nu$  are  $\mathcal{F}_{ln_a}$ -measurable, this conditioning can be inserted.

Define the piecewise constant interpolation of

$$\frac{1}{n_a} \sum_{k \in L_a} E_{ln_a} H(f_x(x_k^N)) \sum_{i=1}^{\lambda} \tilde{z}_k(i) I_{\{f(x_k^N + z_k(i)) = \min_{\mu \in \Lambda_k} f(x_k^N + \mu)\}} q_N(x_k^N)$$

on  $[l\delta_a, l\delta_a + \delta_a)$  as  $\tilde{H}^a(\tau)$ . Then

$$\lim_{n} Eh(x^{a,N}(t_{j}), \ j \leq \nu) \quad \left( \int_{t}^{t+s} \tilde{H}^{a}(\tau) du - \sum_{l\delta_{1}=t}^{t+s} \delta_{a} \frac{1}{n_{a}} \sum_{k \in L_{a}} E_{ln_{a}} H(f_{x}(x_{k}^{N})) \right) \\ \times \sum_{i=1}^{\lambda} \tilde{z}_{k}(i) I_{\{f(x_{k}^{N}+z_{k}(i))=\min_{\mu \in \Lambda_{k}} f(x_{k}^{N}+\mu)\}} q_{N}(x_{k}^{N}) = 0.$$

In view of the above equation, we need only consider the limit of the function  $\tilde{H}^a(\cdot)$  as  $a \to 0$ .

By using the nested conditional expectation, for  $k \ge ln_a$ ,

$$E_{ln_{a}}H(f_{x}(x_{k}^{N}))\sum_{i=1}^{\lambda}\tilde{z}_{k}(i)I_{\{f(x^{N}+z_{k}(i))=\min_{\mu\in\Lambda_{k}}f(x_{k}^{N}+\mu)\}}q_{N}(x_{k}^{N})$$

$$= E_{ln_{a}}H(f_{x}(x_{k}^{N}))\sum_{i=1}^{\lambda}E_{k}\tilde{z}_{k}(i)I_{\{f(x_{k}^{N}+z_{k}(i))=\min_{\mu\in\Lambda_{k}}f(x_{k}^{N}+\mu)\}}q_{N}(x_{k}^{N})$$

$$= E_{ln_{a}}H(f_{x}(x_{k}^{N}))\zeta(x_{k}^{N})q_{N}(x_{k}^{N}) \qquad \text{by Theorem 3.1.}$$

In view of the continuity of the functions  $H(\cdot)$ ,  $f(\cdot)$ ,  $\zeta(\cdot)$  and  $q_N(\cdot)$ , we have

$$\frac{1}{n_a} \sum_{k \in L_a} H(f_x(x_k^N))\zeta(x_k^N)q_N(x_k^N) = \frac{1}{n_a} \sum_{k \in L_a} H(f_x(x_{ln_a}^N)\zeta(x_{ln_a}^N)q_N(x_{ln_a}^N) + o(1) \\
= H(f_x(x^{a,N}(l\delta_a)))\zeta(x^{a,N}(l\delta_a))q_N(x_{ln_a}^N) + o(1),$$

where  $o(1) \xrightarrow{a} 0$  in probability. Letting  $l\delta_a \to \tau$ , we can further replace

$$H(f_x(x^{a,N}(l\delta_a)))\zeta(x^{a,N}(l\delta_a))q_N(x^{a,N}(l\delta_a))$$
by  
$$H(f_x(x^N(\tau)))\zeta(x^N(\tau))q_N(x^N(\tau)).$$

The limit for the truncated process is thus proved.

Step 4: (The result for the un-truncated process). The proof is similar to that of Theorem 2 of Chapter 3 in Kushner (1984). Let  $P_{x(0)}(\cdot)$  and  $P^N(\cdot)$  be the measures induced by  $x(\cdot)$  and  $x^N(\cdot)$ , respectively. Due to the uniqueness of (10),  $P_{x(0)}(\cdot)$  is unique. For each  $T < \infty$ ,  $P_{x(0)}(B) = P^N(B)$  for each Borel subsets  $B \subset \mathcal{B}$ , where

$$\mathcal{B} = \{x(\cdot) \in D^d[0,\infty); \ x(t) \in S_N \text{ for each } t \leq T\}.$$

Observe that

$$P_{x(0)}\{\sup_{t\leq T}|x(t)|\leq N\}\xrightarrow{N} 1 \text{ as } N\to\infty.$$

The desired result thus follows. The proof of the theorem is completed.

# 4 Further asymptotic results

We derive further asymptotic properties in this section. In the first subsection, we derive an upper bound of the estimation errors and in the second subsection, we study the case a is small and n is large and obtain limit results.

#### 4.1 An upper bound on the estimation error

The result is recorded in Theorem 4.1 below. It indicates how the estimation errors depend on the step size, and gives us a way to assess the rate of convergence.

**Theorem 4.1.** Let the conditions of Theorem 3.3 be satisfied. Suppose that  $\theta$  is an asymptotically stable point of (10), and suppose that there is a twice continuously differentiable Liapunov function  $V(\cdot) : \mathbb{R}^d \to \mathbb{R}$ , such that  $V(x) \ge 0$  for all  $x, V(x) \xrightarrow{|x| \to \infty} \infty, V_{xx}(\cdot)$  is bounded, and  $V'_x(x)H(f_x(x))\zeta(x) < -\eta V(x)$  for all  $x \neq \theta$  and for some  $\eta > 0$ . Furthermore, let  $|H(x)| \le K(1+|x|)$  and  $|f_x(x)|^2 \le K(1+V(x))$ . Then the following statements hold:

- 1.  $EV(x_n) < \infty$  and hence  $\{x_n\}$  is tight in  $\mathbb{R}^d$ .
- 2. There is an  $N_a > 0$  and an  $a_0 > 0$  such that for all  $n \ge N_a$ , and all  $a \le a_0$ ,  $EV(x_n) = O(a).$

Remark: It has been shown in our Theorem 3.1 that the vector field  $\zeta(\cdot)$  is always in the downhill direction. Our analysis is based on the Liapunov stability theory. Although we assumed the existence of a Liapunov function, its actual form need not be known. As far as the theoretical development is concerned, there is no loss of generality to assume that  $\theta = 0$  since we can always translate the coordinate axes by subtracting  $\theta$ . Henceforth, we work with the case  $\theta = 0$ . This is rather convenient for notational concern.

Proof of Theorem 4.1. We shall only prove the second part of the theorem. The proof of the first part is easier.

Using the recursive formula (4), and owing to the fact that  $x_n$  is  $\mathcal{F}_n$ -measurable, together with the continuity of  $H(\cdot)$  and  $f(\cdot)$ ,

$$E_{n}V(x_{n+1}) - V(x_{n}) = aV'_{x}(x_{n})H(f_{x}(x_{n}))\sum_{\substack{i=1\\\lambda}}^{\lambda} \tilde{z}_{n}(i)I_{\{f(x_{n}+z_{n}(i))=\min_{\mu\in\Lambda_{n}}f(x_{n}+\mu)\}} + a^{2}H^{2}(f_{x}(x_{n}))\sum_{\substack{i=1\\\lambda}}^{\lambda} \tilde{z}'_{n}(i)I_{\{f(x_{n}+z_{n}(i))=\min_{\mu\in\Lambda_{n}}f(x_{n}+\mu)\}} \times V_{xx}(x_{n}^{+})\sum_{\substack{i=1\\\lambda}}^{\lambda} \tilde{z}_{n}(i)I_{\{f(x_{n}+z_{n}(i))=\min_{\mu\in\Lambda_{n}}f(x_{n}+\mu)\}} \leq aV'_{x}(x_{n})H(f_{x}(x_{n}))\zeta(x_{n}) + Ka^{2}H^{2}(f_{x}(x_{n}))E_{n}|\tilde{z}_{n}|^{2} \leq -\eta aV(x_{n}) + Ka^{2}(1+V(x_{n})),$$

$$(14)$$

where  $x_n^+$  is a point on the line segment joining  $x_n$  and  $x_{n+1}$ .

For sufficiently small a > 0, i.e., there exists an  $a_0 > 0$  such that for all  $0 < a \le a_0$ ,  $Ka^2 - \eta a \le -\eta/2a$ , and as a result

$$E_n V(x_{n+1}) \le (1 - \eta/2a) V(x_n) + Ka^2.$$

Taking expectation and iterating on the above inequality leads to

$$EV(x_{n+1}) \leq (1 - \eta a/2)^n EV(x_0) + Ka^2 \sum_{i=0}^n (1 - \eta a/2)^i$$
  
$$\leq (1 - \eta a/2)^n EV(x_0) + Ka.$$
(15)

Choose  $N_a$  such that for all  $n \ge N_a$ ,  $(1 - \eta a/2)^n \le Ka$ . Then the desired estimate follows from (15). This concludes the proof of the theorem.

To proceed, we note that Theorem 3.3 gives us a convergence result that is on arbitrarily large but still bounded time interval. One of our interests is to figure out what happens when a is small and n is large. This problem is treated in Section 4.2. As it was, we assume that  $\theta = 0$  throughout.

### **4.2** Convergence for $a \downarrow 0$ and $n \uparrow \infty$

For all problems, it is important to have a convergence result which is uniform in t. The way to study the problem is similar to what has been done in Theorem 3.3. However, we wish to have the time variable tends to infinity. Introduce another sequence  $\{t_a\}$  such that  $t_a \to \infty$  as  $a \to 0$ . The significance of the differential equation (10) is that its stationary points correspond to the points  $f_x(x) = 0$  we are searching for. In lieu of  $x^a(\cdot)$  as defined in Section 3, consider  $x^a(\cdot + t_a)$ . Note that the weak convergence result alone does not imply that  $x^a(\cdot + t_a)$  converges weakly to a stationary solution to (10). In what follows, we establish the convergence of this sequence.

**Theorem 4.2.** Under the conditions of Theorem 4.1,  $x^a(\cdot + t_a)$  converges to  $\theta = 0$  weakly.

Proof. The proof is quite similar to that of Theorem 3.3 and to a corresponding result in Kushner and Yin (1987), so we will be very brief. For each  $T < \infty$ , consider the pair  $(x^a(\cdot + t_a), x^a(\cdot - T + t_a))$ . The tightness can be obtained as in the previous case. Therefore, we can extract a convergent subsequence, still use a as the index and denote the limit by  $(x(\cdot), x_T(\cdot))$ . It is clear that  $x(0) = x_T(T)$  due to the construction. The value of  $x_T(0)$  may be not known. However, in accordance with Theorem 4.1, the values of it belongs to a set that is tight. As a result, by the stability argument, for any  $\hat{\varepsilon} > 0$ , there is a  $T_{\hat{\varepsilon}}$  such that for all  $T \ge T_{\hat{\varepsilon}}$ ,  $P(|x_T(T)| > \hat{\varepsilon}) \le \hat{\varepsilon}$ . This completes the proof of the theorem.

### 5 Algorithms with decreasing step size

There are times that we may wish to use decreasing step size algorithms. This section is concentrated on the study of such algorithms related to the  $(1, \lambda)$  strategy. Consider

$$x_{n+1} = x_n + a_n H(f_x(x_n)) \sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x_n + z_n(i)) = \min_{\mu \in \Lambda_n} f(x_n + \mu)\}},$$
(16)

where  $\{a_n\}$  is a sequence of nonnegative real numbers such that

$$a_n \xrightarrow{n} 0, \quad \sum_{n=1}^{\infty} a_n = \infty, \quad \sum_{n=1}^{\infty} a_n^2 < \infty.$$

A typical choice of  $\{a_n\}$  is  $a_n = O(1/n^{\gamma})$ , with  $1/2 < \gamma \leq 1$ . In what follows, we derive the w.p.1 convergence result and then establish an upper bound on the estimation errors.

#### 5.1 W.p.1 convergence

In this section, we obtain a w.p.1 convergence result by using the ordinary differential equation method (see Benveniste, Métivier, and Priouret 1990, Kushner and Clark 1978, Ljung 1977, and the references therein). For future use define

$$\hat{S} = \{x; \ H(f_x(x))\zeta(x) = 0\},\$$
  
$$\psi_n(x) = \sum_{i=1}^{\lambda} \tilde{z}_n(i) I_{\{f(x+z_n(i))=\min_{\mu \in \Lambda_n} f(x+\mu)\}}$$

**Theorem 5.1.** Let (A) hold. Suppose that there is a twice continuously differentiable Liapunov function  $V(\cdot) : \mathbb{R}^d \to \mathbb{R}$ , such that  $V(x) \ge 0$  for all  $x, V(x) \xrightarrow{|x| \to \infty} \infty, V_{xx}(\cdot)$  is bounded, and  $V'_x(x)H(f_x(x))\zeta(x) < -\eta_0$  for all x and for some  $\eta_0 > 0$ , and  $|V'_x(x)H(f_x(x))\zeta(x)| \le K(1+V(x))$ . Furthermore,  $|H(x)| \le K(1+|x|)$  and  $|f_x(x)|^2 \le K(1+|V'_x(x)H(f_x(x))\zeta(x)|)$ . Then  $\{x_n\}$  is bounded w.p.1. In addition,  $x_n$  converges to  $\hat{S}$  w.p.1, i.e.,

$$P\left(\lim_{n\to\infty}\rho(x_n,\hat{S})=0\right)=1,$$

where  $\rho(\cdot)$  denotes the usual distance function such that  $\rho(x, \hat{S}) = \inf_{y \in \hat{S}} \rho(x, y)$ . In particular, if  $\hat{S} = \{\theta\}$ , a singleton set, then  $x_n \xrightarrow{n} \theta$  w.p.1.

Proof. We first note owing to the assumptions of the theorem

$$H^{2}(f_{x}(x)) \leq K(1 + |V'_{x}(x)H(f_{x}(x))\zeta(x)|).$$

Direct computation yields that

$$E_{n}V(x_{n+1}) - V(x_{n}) = a_{n}V'_{x}(x_{n})H(f_{x}(x_{n}))E_{n}\psi_{n}(x_{n}) +a_{n}^{2}H^{2}(f_{x}(x_{n}))E_{n}\psi'_{n}(x_{n})V_{xx}(x_{n}^{+})\psi_{n}(x_{n}) = a_{n}(1 + a_{n}\tilde{\varepsilon}_{n})V'_{x}(x_{n})H(f_{x}(x_{n}))\zeta(x_{n}) + a_{n}^{2}\varepsilon_{n} \leq -Ka_{n}\eta_{0} + a_{n}^{2}\varepsilon_{n} < 0,$$
(17)

for some  $N_0$  and all  $n \ge N_0$ , where  $x_n^+$  is point with all of its components sitting between  $x_n$ and  $x_{n+1}$ ;  $\{\tilde{\varepsilon}_n\}$  and  $\{\varepsilon_n\}$  are sequences of uniformly bounded random variables.

In view of (17), for  $n \ge N_0$ ,  $V(x_n)$  is a supermartingale. Since  $V(x_n) \ge 0$ , the limit of  $V(x_n)$  exists w.p.1. Since  $V(x) \xrightarrow{|x| \to \infty} \infty$ , the boundedness of  $\{x_n\}$  (in the sense of w.p.1) follows.

To complete the proof of the theorem, we apply a result from Kushner and Clark (1978). First rewrite the recursion as

$$x_{n+1} = x_n + a_n H(f_x(x_n))\zeta(x_n) + a_n H(f_x(x_n))[\psi_n(x_n) - \zeta(x_n)].$$
(18)

In accordance with Theorem 2.4.2 of Kushner and Clark (1978), we need only verify that

$$m_n = \sum_{i=1}^n a_i H(f_x(x_i))[\psi_i(x_i) - \zeta(x_i)] \text{ converges w.p.1.}$$
(19)

It is readily seen that  $m_n$  is a martingale. Since  $\{x_n\}$  is bounded w.p.1,  $H(f_x(x_n))$  is also bounded w.p.1 by the continuity of  $H(\cdot)$  and  $f_x(\cdot)$ . Consequently, since  $x_i$  is  $\mathcal{F}_i$ -measurable,

$$\sum_{i=1}^{\infty} a_i^2 H^2(f_x(x_i)) E_i |\psi_i(x_i) - \zeta(x_i)|^2$$
  

$$\leq K \sum_{i=1}^{\infty} a_i^2 [E|\tilde{z}_n(i)|^2 + E|\zeta(x_i)|^2] < \infty,$$

since  $\{x_i\}$  is bounded w.p.1 and  $\zeta(\cdot)$  is continuous. Owing to the local martingale convergence theorem in Chow (1965), (19) holds. Now apply Theorem 2.4.2 of Kushner and Clark, the desired results follow.

## **5.2** $EV(x_n) = O(1/n^{\gamma})$ for $a_n = 1/n^{\gamma}$

Next we obtain an upper bound for the estimation errors. Since the techniques and details are similar to that of Theorem 4.1, we shall omit the proof.

**Theorem 5.2.** Let the conditions of Theorem 5.1 be satisfied, and  $a_n = 1/n^{\gamma}$  for some  $1/2 < \gamma \leq 1$ . Suppose that  $\theta$  is an asymptotically stable point of (10), and suppose that in addition to the conditions of Theorem 5.1, the Liapunov function  $V(\cdot)$  satisfies  $V'_x(x)H(f_x(x))\zeta(x) < -\eta V(x)$  for all  $x \neq \theta$  and for some  $\eta > 0$ . Then there is an N > 0 such that for all  $n \geq N$ ,  $EV(x_n) = O(n^{-\gamma})$ .

### 6 Concluding remarks

In this work, asymptotic properties of the  $(1, \lambda)$  evolution strategy was developed by use of stochastic approximation methods. The evolutionary algorithm was rewritten in a recursive form and then the analytic tools in stochastic approximation were employed to carry out the investigation. We considered both constant step size and decreasing step size algorithms. Under suitable conditions, we have obtained the convergence and the error bounds of the underlying algorithm. Our current effort lies in studying more complex situations, and extend the results to evolution strategies with noisy evolution.

This paper is our first effort in using stochastic approximation method to analyze the evolution strategies. One of the immediate questions is can we extend the results to non-smooth functions  $f(\cdot)$ ? We believe the answer is affirmative. It seems that we can deal with non-smooth functions by use of the non-smooth analysis techniques in conjunction with stochastic approximation methods. The analysis for the corresponding recursive procedures then becomes that of set-valued, and the differential equations become differential inclusions. Much more details need to be given serious thoughts, and deserve further study and in depth investigation.

Recently, several modifications on the standard stochastic approximation procedures were proposed by Polyak, Ruppert and Bather (see Yin and Yin (1994) or Yin (to appear) for more details and references), which result in asymptotical optimality. Such attempts have soon attracted much attention. One of the ingredients of their approach is the use of arithmetic averaging. It is conceivable such an idea may well be suited for studying the evolutionary algorithms.

A further promising observation is that stochastic approximation methods provide a route to convert discrete time EAs to *equivalent* continuous time evolutionary processes so that we can apply the well developed mathematical apparatus to analyze continuous time stochastic processes to the investigation of evolutionary algorithms that are more complex than the  $(1, \lambda)$  EA considered here.

### Appendix: Distribution of $\cos \omega$

Proof of Lemma 3.2. Let  $X \sim N(0, I_d)$ . Then U = X/|X| is uniformly distributed on a hypersphere surface of dimension d and X and |X| are stochastically independent (see Fang, Kotz, and Ng 1990). We wish to answer the question: what is the distribution of  $\cos \omega$ , where  $\omega$  is an angle between a vector y and u? Without loss of generality, choose  $y = e_1 = (1, 0, \dots, 0)'$ . Then

$$\cos(\omega) = e'_1 u = \frac{e_1 X}{|X|} = \frac{X_1}{|X|}$$

and

$$Z = \cos^2 \omega = \frac{X_1^2}{\sum_{i=1}^d X_i^2} = \frac{X_1^2}{X_1^2 + \sum_{i=2}^d X_i^2}$$

Thus,  $X_1^2 \sim \chi_1^2$  and  $\sum_{i=2}^d X_i^2 \sim \chi_{d-1}^2$ , so that  $Z \sim \text{Beta}\left(\frac{1}{2}, \frac{d-1}{2}\right)$  with probability density function (p.d.f.)

$$f(z) = \frac{z^{-1/2} (1-z)^{(d-3)/2}}{B(1/2, (d-1)/2)} \cdot I_{[0,1]}(z) .$$

Transformation of the above density with  $C = Z^{1/2}$  leads to

$$f(c) = \frac{2}{B(1/2, (d-1)/2)} (1-c)^{(d-3)/2} (1+c)^{(d-3)/2} \cdot I_{[0,1]}(c) ,$$

which is the p.d.f. of  $|\cos \omega|$ . Due to symmetry of the cosine we may modify the above p.d.f. to obtain

$$f(c) = \frac{1}{B(1/2, (d-1)/2)} (1-c)^{(d-3)/2} (1+c)^{(d-3)/2} \cdot I_{[-1,1]}(c)$$

so that the distribution function of  $\cos \omega$  is

$$F_c(x) = \frac{1}{B(1/2, (d-1)/2)} \int_{-1}^x (1-c)^{(d-3)/2} (1+c)^{(d-3)/2} dc .$$

As a result, direct substitutions yield

$$F_{c}(x) = \frac{2^{d-2}}{B(1/2, (d-1)/2)} \int_{0}^{(x+1)/2} y^{(d-3)/2} (1-y)^{(d-3)/2} dy$$
  
=  $\frac{1}{B((d-1)/2, (d-1)/2)} \int_{0}^{(x+1)/2} y^{(d-3)/2} (1-y)^{(d-3)/2} dy$  (20)

where we used the relation  $B(1/2, (d-1)/2) = 2^{d-2} B((d-1)/2, (d-1)/2)$ . Thus,  $\cos \omega$  possesses a Beta distribution on [-1, 1].

### References

Bäck, T., G. Rudolph, and H.-P. Schwefel (1993). Evolutionary programming and evolution strategies: Similarities and differences. In D. Fogel and W. Atmar (Eds.), *Proceedings of the 2nd Annual Conference on Evolutionary Programming*, pp. 11–22. La Jolla (CA): Evolutionary Programming Society.

Bäck, T. and H.-P. Schwefel (1993). An overview of evolutionary algorithms for parameter optimization. Evolutionary Computation 1(1), 1–23.

Benveniste, A., M. Métivier, and P. Priouret (1990). Adaptive Algorithms and Stochastic Approximation. Berlin: Springer.

Chow, Y. (1965). Local convergence of martingales and the law of large numbers. Ann. Math. Statist. 36, 552–558.

David, H. (1970). Order Statistics. New York: Wiley.

Ethier, S. and T. Kurtz (1986). *Markov Processes, Characterization and Convergence*. New York: Wiley.

Fang, K.-T., S. Kotz, and K.-W. Ng (1990). Symmetric Multivariate and Related Distributions. London and New York: Chapman and Hall.

Fogel, D. B. (1992). An analysis of evolutionary programming. In D. B. Fogel and W. Atmar (Eds.), *Proceedings of the 1st Annual Conference on Evolutionary Programming*, San Diego (CA), pp. 43–51. Evolutionary Programming Society.

Fogel, L. J., A. J. Owens, and M. J. Walsh (1966). Artificial Intelligence through Simulated Evolution. New York: Wiley.

Holland, J. (1962). Outline for a logical theory of adaptive systems. J. Assoc. Comp. Machinery 3, 297-314.

Kushner, H. (1984). Approximation and Weak Convergence Methods for Random Processes, with applications to Stochastic Systems Theory. Cambridge (MA): MIT Press.

Kushner, H. and D. Clark (1978). Stochastic Approximation Methods for Constrained and Unconstrained Systems. Springer.

Kushner, H. and G. Yin (1987). Asymptotic properties of distributed and communicating stochastic approximation algorithms. SIAM J. Control Optim. 25, 1266–1290.

Ljung, L. (1977). Analysis of recursive stochastic algorithms. *IEEE Trans. Automatic Control 22*, 551–575.

De Jong, K. (1975). An analysis of the behavior of a class of genetic adaptive systems. Ph. D. thesis, University of Michigan.

Rechenberg, I. (1965, August). Cybernetic solution path of an experimental problem. Royal Aircraft Establishment, Library translation No. 1122, Farnborough, Hants., UK.

Rechenberg, I. (1973). Evolutionsstrategie: Optimierung technischer Systeme nach Prinzipien der biologischen Evolution. Stuttgart: Frommann-Holzboog Verlag.

Rudolph, G. (1994). Convergence of non-elitist strategies. In Proceedings of the First IEEE Conference on Computational Intelligence, Vol. 1, pp. 63-66. IEEE Press.

Schwefel, H.-P. (1965). Kybernetische Evolution als Strategie der experimentellen Forschung in der Strömungstechnik. Diploma thesis, Technische Universität Berlin.

Schwefel, H.-P. (1977). Numerische Optimierung von Computer-Modellen mittels der Evolutionsstrategie. Interdisciplinary systems research; 26. Basel: Birkhäuser.

Schwefel, H.-P. (1981). Numerical Optimization of Computer Models. Chichester: Wiley.

Schwefel, H.-P. (1993, August). Evolving Evolutionary Algorithms, Invited Opening Lecture, SIAM Conference on Simulation and Monte Carlo Methods.

Yin, G. (to appear). Adaptive filtering with averaging. IMA Volume in Applied Math.

Yin, G., G. Rudolph, and H.-P. Schwefel (1995). Establishing connections between evolutionary algorithms and stochastic approximation. *Informatica* 6(1), 93–116.

Yin, G. and K. Yin (1994). Asymptotically optimal rate of convergence of smoothed stochastic recursive algorithms. *Stochastics Stochastic Rep.* 47, 21–46.